# ELLIPTIC COMPLEXES AND INDEX THEORY 

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## 1. Sobolev Theory on $\mathbb{R}^{n}$

As general references for this section, see the books [Nar], Ch.3, [Gil], Ch.1, [Hor], Ch I, II and [Rud], Ch. 6,7.
1.1. Test functions and distributions. We introduce some standard notation. For a multi-index $\alpha=$ $\left(\alpha_{1}, . ., \alpha_{n}\right)$ of length $n$, the symbol $|\alpha|:=\sum_{i} \alpha_{i}$, and $\alpha!:=\alpha_{1}!, . ., \alpha_{n}!$.

For derivatives, we denote:

$$
d_{x}^{\alpha}=\left(\frac{\partial}{\partial x_{1}}\right)^{\alpha_{1}} \ldots \ldots\left(\frac{\partial}{\partial x_{n}}\right)^{\alpha_{n}}, \quad \partial_{j}=\left(\frac{\partial}{\partial x_{j}}\right), \quad D_{j}=\frac{1}{\sqrt{-1}} \partial_{j}
$$

Finally $x_{1}^{\alpha_{1}} x_{2}^{\alpha_{2}} \ldots x_{n}^{\alpha_{n}}$ will be denoted simply by $x^{\alpha}$.
We can define some standard function spaces on $\mathbb{R}^{n}$. For us, functions will always be complex valued.
Definition 1.1.1 (Standard function spaces on $\mathbb{R}^{n}$ ). . The function spaces defined below are all complex vector spaces, and to define a topology on them, it is enough to define convergence to zero.
(i):

$$
C^{\infty}\left(\mathbb{R}^{n}\right)=\left\{\text { smooth functions on } \mathbb{R}^{n}\right\}
$$

Define $f_{n} \rightarrow 0$ if $d_{x}^{\alpha} f_{n} \rightarrow 0$ uniformly on compact sets for all $|\alpha| \geq 0$. This space is also sometimes denoted $\mathcal{E}$ by analysts.
(ii):

$$
C_{0}^{\infty}\left(\mathbb{R}^{n}\right)=\left\{f \in C^{\infty}\left(\mathbb{R}^{n}\right): \lim _{x \rightarrow \infty} d_{x}^{\alpha} f=0 \text { for all }|\alpha| \geq 0\right\}
$$

This is the space of all smooth functions whose derivatives of all orders vanish at infinity. The topology in this space is the subspace topology from $C^{\infty}\left(\mathbb{R}^{n}\right)$. It is often denoted $\mathcal{E}_{0}$.
(iii):

$$
C_{c}^{\infty}\left(\mathbb{R}^{n}\right)=\left\{f \in C^{\infty}\left(\mathbb{R}^{n}\right): \text { support } f \text { is compact }\right\}
$$

Its topology is defined by $f_{n} \rightarrow 0$ if there exists a compact set $K$ such that $\operatorname{supp} f_{n} \subset K$ for all $n$ and $d_{x}^{\alpha} f_{n} \rightarrow 0$ uniformly on $K$ for all $|\alpha| \geq 0$. Note that this is not the subspace topology from $C^{\infty}\left(\mathbb{R}^{n}\right)$, for if we define a take a non-zero function $\psi$ on $\mathbb{R}$ with compact support $[-1,1]$ say, and let $f_{n}(x)=\psi(x-n)$, (which has support $[n-1, n+1]$ ), then $f_{n} \rightarrow 0$ in $C^{\infty}(\mathbb{R})$, but not in $C_{c}^{\infty}(\mathbb{R})$. It is, in fact easily seen to be strictly finer than the subspace topology. This space is denoted $\mathcal{D}$ by analysts, and also called the space of test functions.
(iv):

$$
\mathcal{S}\left(\mathbb{R}^{n}\right)=\left\{f \in C^{\infty}\left(\mathbb{R}^{n}\right): \sup _{x \in \mathbb{R}^{n}}\left|x^{\alpha} d_{x}^{\beta} f(x)\right| \leq C_{\alpha \beta} \text { for all }|\alpha|,|\beta| \geq 0\right\}
$$

This is called the Schwartz space of rapidly decreasing functions. Define the topology by declaring $f_{n} \rightarrow 0$ if

$$
\sup _{x \in \mathbb{R}^{n}}\left|x^{\alpha} d_{x}^{\beta} f_{n}(x)\right| \rightarrow 0
$$

for each $|\alpha|,|\beta| \geq 0$.

It is an easy exercise to see that there are natural inclusions:

$$
C_{c}^{\infty}\left(\mathbb{R}^{n}\right) \subset \mathcal{S}\left(\mathbb{R}^{n}\right) \subset C_{0}^{\infty}\left(\mathbb{R}^{n}\right) \subset C^{\infty}\left(\mathbb{R}^{n}\right)
$$

all of which are continuous, and all of which are strict. The reader may also check that the inclusion $C_{c}^{\infty} \subset C^{\infty}$ is dense, (by using multiplication with cutoff functions $\phi_{n}$ which are identically 1 on a ball of radius $n$ and identically zero outside a ball of radius say $2 n$ ), and hence all the inclusions above are dense.

On $\mathcal{S}$ or $C_{c}^{\infty}$, we may introduce the $L_{p}$-norm defined by:

$$
\|f\|_{L_{p}}:=\left(\int_{\mathbb{R}^{n}}|f|^{p} d x\right)^{\frac{1}{p}}
$$

for $1 \leq p<\infty$. Upon completing either of these two spaces with respect to this norm, one gets the Banach space $L_{p}\left(\mathbb{R}^{n}\right)$. For $p=\infty$, this is false, as can be seen by looking at the non-zero constant functions. $L_{\infty}\left(\mathbb{R}^{n}\right)$ is got by taking all measurable functions on $\mathbb{R}^{n}$ which are essentially bounded.
1.2. The Fourier Transform and Plancherel Theorem. In the sequel we will simply write $C^{\infty}$ for $C^{\infty}\left(\mathbb{R}^{n}\right)$, and so on, if no confusion is likely. Also, to eliminate annoying powers of $2 \pi$, we introduce the measure (volume element) $d x$ on $\mathbb{R}^{n}$ by the formula:

$$
d x:=(2 \pi)^{-n / 2} d x_{1} \ldots d x_{n}
$$

Definition 1.2.1. For $f \in \mathcal{S}$, define the Fourier Transform of $f$ by the formula

$$
\widehat{f}(\xi):=\int_{\mathbb{R}^{n}} f(x) e^{-i \xi \cdot x} d x
$$

which makes sense for any $f \in L_{1}$, and in particular for $f \in \mathcal{S}$. Here $\xi . x=\sum_{i} \xi_{i} x_{i}$ is the usual Euclidean inner product of vectors in $\mathbb{R}^{n}$. Similarly, for $f \in \mathcal{S}$, define the Inverse Fourier Transform of $f$ by the formula:

$$
f^{\vee}(\xi):=\widehat{f}(-\xi)=\int_{\mathbb{R}^{n}} f(x) e^{i \xi \cdot x} d x
$$

Finally, for $f, g \in \mathcal{S}$, define the convolution product

$$
f * g(x):=\int_{\mathbb{R}^{n}} f(x-y) g(y) d y=\int_{\mathbb{R}^{n}} f(-z) g(z+x) d z
$$

It is easy to verify (taking limits inside the integral sign after appealing to Lebesgue's Dominated Convergence Theorem) that $f * g$ is also in $\mathcal{S}$, and that $f * g=g * f$.

Before proving the main proposition of this section, we need a couple of useful lemmas. Note that the Gaussian function

$$
\psi(x)=e^{-\frac{|x|^{2}}{2}}
$$

is in $\mathcal{S}$. Also its integral $\int_{\mathbb{R}^{n}} \psi(x) d x=1$.

Lemma 1.2.2. For the Gaussian $\psi$ above, we have $\widehat{\psi}=\psi$.
Proof: We have:

$$
\begin{aligned}
\widehat{\psi}(\xi) & =\int e^{-i \xi \cdot x} e^{-\frac{|x|^{2}}{2}} d x \\
& =e^{-\frac{|\xi|^{2}}{2}} \int e^{\frac{(x+i \xi) \cdot(x+i \xi)}{2}} d x
\end{aligned}
$$

Let $x=\left(x_{1}, . ., x_{n}\right)$ and $\xi=\left(\xi_{1}, . ., \xi_{n}\right)$. By choosing a rectangular contour in $\mathbb{C}$ with vertices $-a, a,-a+i \xi_{1}, a+$ $i \xi_{1}$, noting that the integral of the holomorphic function $e^{-z_{1}^{2} / 2}$ around this contour is zero, and also that the contributions along the vertical edges $(-a+i t)$ and $(a+i t)$ for $0 \leq t \leq \xi_{1}\left(\right.$ if $\left.\xi_{1} \geq 0\right)\left(\right.$ resp. $\xi_{1} \leq t \leq 0$ if $\left.\xi \leq 0\right)$ converge to zero as $a_{1} \rightarrow \infty$, we see that:

$$
\int_{\mathbb{R}} e^{-\left(x_{1}+i \xi_{1}\right)^{2} / 2} d x_{1}=\int_{\mathbb{R}} e^{-x_{1}^{2} / 2} d x_{1}=\sqrt{2 \pi}
$$

apply the argument variable by variable to conclude that:

$$
\int_{\mathbb{R}^{n}} e^{-(x+i \xi) \cdot(x+i \xi) / 2} d x=(2 \pi)^{-\frac{n}{2}} \int_{\mathbb{R}} \int_{\mathbb{R}} \ldots \int_{\mathbb{R}} e^{-\sum_{j}\left(x_{j}+i \xi_{j}\right)^{2} / 2} d x_{1} \ldots d x_{n}=1
$$

which proves our assertion.
Lemma 1.2.3 (Approximate identities). Let $\phi \in \mathcal{S}$, such that $\phi(x) \geq 0$ for all $x$ and $\int \phi(x) d x=1$. For $\epsilon>0$, define the approximate identity or mollifier:

$$
\phi_{\epsilon}(x)=\epsilon^{-n} \phi(x / \epsilon)
$$

Then for any $f \in \mathcal{S}$, we have $f * \phi_{\epsilon}$ converges uniformly to $f$ as $\epsilon \rightarrow 0$.
Proof: Since $\int_{\mathbb{R}^{n}} \phi_{\epsilon} d x=1$ for each $\epsilon>0$, we have:

$$
\begin{aligned}
\left|\left(f * \phi_{\epsilon}\right)(x)-f(x)\right| & =\left|\int \phi_{\epsilon}(y)(f(x-y)-f(x)) d y\right| \\
& \leq \int\left|\phi_{\epsilon}(y)(f(x-y)-f(x))\right| d y
\end{aligned}
$$

Let $C>0$ be such that $\int|f(x)| d x \leq C$. Now let $\eta>0$ be any positive number. Choose a $\delta>0$ (by uniform continuity of $f$ ) such that $|f(x-y)-f(x)| \leq \eta$ for all $|y| \leq \delta$, and all $x$. Then

$$
\int_{|y| \leq \delta}\left|\phi_{\epsilon}(y)(f(x-y)-f(x))\right| d y \leq \eta \int_{|y| \leq \delta} \phi_{\epsilon}(y) d y \leq \eta \int_{\mathbb{R}^{n}} \phi_{\epsilon}(y) d y=\eta
$$

for all $\epsilon>0$.
Now choose an $\epsilon_{0}>0$ small enough so that $\int_{|y|>\delta} \phi_{\epsilon}(y) d y<\eta / 2 C$ for $\epsilon \leq \epsilon_{0}$. Then, we have:

$$
\int_{|y|>\delta}\left|\phi_{\epsilon}(y)(f(x-y)-f(x))\right| d y \leq \int_{|y|>\delta} \phi_{\epsilon}(y)(2 C) d y \leq \eta
$$

for $\epsilon<\epsilon_{0}$. Combining the integrals for $|y| \leq \delta$ and $|y|>\delta$, we get:

$$
\left|f * \phi_{\epsilon}(x)-f(x)\right| \leq 2 \eta \text { for all } \epsilon<\epsilon_{0}
$$

independent of $x$. That is, $f * \phi_{\epsilon} \rightarrow f$ uniformly as $\epsilon \rightarrow 0$.
Remark 1.2.4. It follows from the above that one can take any non-negative compactly supported function $\phi$ and define the approximate identities $\phi_{\epsilon}$. Similarly, by starting with the Gaussian $\psi$ defined above, we get that the functions:

$$
\psi_{\epsilon}(x)=\epsilon^{-n} e^{-|x|^{2} / 2 \epsilon^{2}}
$$

are approximate identities.

Proposition 1.2.5. We have the following facts about the Fourier transform on the Schwartz class $\mathcal{S}$.
(i): The map $f \mapsto \widehat{f}$ is an isomorphism of $\mathcal{S}$ with itself, of order 4. In fact,

$$
(\widehat{f})^{\wedge}(x)=f(-x),(\widehat{f})^{\vee}(x)=f(x) \text { for all } f \in \mathcal{S}
$$

(The second formula is called the Fourier Inversion Formula.)
(ii): For all multi-indices $\alpha$,

$$
\left(D_{x}^{\alpha} f\right)^{\wedge}(\xi)=\xi^{\alpha} \widehat{f} ; \quad D_{\xi}^{\alpha} \widehat{f}(\xi)=(-1)^{|\alpha|}\left(x^{\alpha} f\right)^{\wedge}
$$

In particular, by the first formula, if $P$ is an $n$-variable polynomial with complex coefficients, then for the constant coefficient differential operator $P(D)$ we have:

$$
(P(D) f)^{\wedge}=P(\xi) \widehat{f}
$$

(iii):

$$
\widehat{f} \widehat{g}=(f * g)^{\wedge}, \quad \widehat{f} * \widehat{g}=(f g)^{\wedge}
$$

(iv): (Plancherel Theorem) The map $f \mapsto \widehat{f}$ is a unitary isomorphism of $\mathcal{S}$ to itself with respect to $L_{2}$-norm. Thus (in view of (i) above), it extends to a unitary isomorphism of $L_{2}\left(\mathbb{R}^{n}\right)$ to itself.
(v): (Riemann-Lebesgue Lemma) There is an inclusion:

$$
\left(L_{1}\left(\mathbb{R}^{n}\right)\right)^{\wedge} \subset C_{0}\left(\mathbb{R}^{n}\right)
$$

where the space on the right is the space of all continuous functions vanishing at $\infty$, with the topology of uniform convergence on compact sets.

Proof: We first prove (ii). Since all derivatives of $f \in \mathcal{S}$ are also in $\mathcal{S}$, and hence Lebesgue integrable, one can use the Lebesgue Dominated Convergence Theorem, and differentiate under the integral sign to get:

$$
\begin{aligned}
D_{\xi}^{\alpha} \widehat{f}(\xi) & =\int_{\mathbb{R}^{n}} D_{\xi}^{\alpha}\left(e^{-i \xi \cdot x} f(x)\right) d x \\
& =\int_{\mathbb{R}^{n}}(-1)^{|\alpha|} x^{\alpha} e^{-i \xi \cdot x} f(x) d x=(-1)^{|\alpha|}\left(x^{\alpha} f\right)^{\wedge}
\end{aligned}
$$

In particular, we have that $\widehat{f}$ is a smooth function. This proves the second part of (ii). To prove the first part, one uses repeated integration by parts and the fact that all derivatives of $f$ vanish at $\infty$ to conclude that:

$$
\begin{aligned}
\left(D_{x}^{\alpha} f\right)^{\wedge}=\int_{\mathbb{R}^{n}} e^{-i \xi \cdot x} D_{x}^{\alpha} f(x) d x & =(-1)^{|\alpha|} \int_{\mathbb{R}^{n}}\left(D_{x}^{\alpha} e^{-i \xi \cdot x}\right) f(x) d x \\
& =\xi^{\alpha} \widehat{f}(\xi)
\end{aligned}
$$

This proves (ii). Thus we also have $(P(D) f)^{\wedge}=P(\xi) \widehat{f}$. From (ii) it also follows that

$$
\xi^{\alpha} D_{\xi}^{\beta} \widehat{f}= \pm \xi^{\alpha}\left(x^{\beta} f\right)^{\wedge}= \pm\left(D_{x}^{\alpha}\left(x^{\beta} f\right)\right)^{\wedge}
$$

and since the function on the right is bounded by the $L_{1}$ norm of $D_{x}^{\alpha}\left(x^{\beta} f\right)$ (a Schwartz class function), it follows that $\widehat{f} \in \mathcal{S}$ as well.
(v) is easy by noting that for an $L^{1}$ function $f$, we have: $\|\widehat{f}\|_{\infty}=\sup _{\xi \in \mathbb{R}^{n}}|\widehat{f}(\xi)| \leq\|f\|_{1}$, and that for any $L_{1}$ function $f$, there is a sequence of $f_{n} \in \mathcal{S}$ with $\left\|f-f_{n}\right\|_{1} \rightarrow 0$. Which implies that $\widehat{f}_{n} \rightarrow \widehat{f}$ uniformly, so that since $\widehat{f}_{n} \in \mathcal{S}$, we have $\lim _{\xi \rightarrow \infty} \widehat{f}(\xi)=0$. This proves (v).

To prove (i), define the operator $T: \mathcal{S} \rightarrow \mathcal{S}$ by $T f(x)=(\widehat{f})^{\wedge}(-x)$. We need to show that $T f(x)=f(x)$ for all $x$. First suppose $f \in \mathcal{S}$ with $f(0)=0$. Then, by the first order Taylor formula we may write:

$$
f(x)=\sum_{j=1}^{n} x_{j} g_{j}(x)
$$

where $g_{j}$ are some smooth functions. Let $\phi$ be a non-negative compactly supported function which is identically $=1$ in a neighbourhood of the origin. Then

$$
f(x)=\phi(x) f(x)+(1-\phi(x)) f(x)=\sum_{j} x_{j} \phi g_{j}+\sum_{j} x_{j}\left(\frac{x_{j}(1-\phi) f}{|x|^{2}}\right)
$$

Since $\phi$ has compact support, the functions $\phi g_{j} \in \mathcal{S}$. On the other hand, since $\phi \equiv 1$ near the origin, the functions $\left(\frac{x_{j}(1-\phi) f}{|x|^{2}}\right) \in \mathcal{S}$ as well. Thus:

$$
f=\sum_{j=1}^{n} x_{j} h_{j}
$$

where $h_{j} \in \mathcal{S}$. However, by (ii) proved above, we have

$$
\widehat{\left(x_{j} h_{j}\right)}=i \frac{\partial \widehat{h}_{j}}{\partial \xi_{j}}
$$

so that $\widehat{f}=i \sum_{j} \frac{\partial \widehat{h}_{j}}{\partial \xi_{j}}$. Thus:

$$
T f(0)=(\widehat{f})^{\wedge}(0)=\int_{\mathbb{R}^{n}} \sum_{j} \frac{\partial \widehat{h}_{j}}{\partial \xi_{j}} d \xi
$$

But by the divergence theorem, the last integral is the limit:

$$
\lim _{R \rightarrow \infty} \int_{S(R)} \mathbf{h} . \nu d \mu
$$

where $S(R)$ is the sphere of radius $R$, and $\nu$ is the unit normal to $S(R)$, and $\mathbf{h}=\left(\hat{h}_{1}, . ., \hat{h}_{n}\right)$, and $d \mu$ is the suitably normalised measure on the sphere. Since $\|\mathbf{h}\|$ decreases faster than all powers of $R$, and the volume of $S(R)$ grows as $R^{n-1}$, the limit above is zero. This proves that $T f(0)=f(0)$ for $f(0)=0$.

Now for $f \in \mathcal{S}$ arbitrary, we write:

$$
f=f(0) \psi+(f-f(0) \psi)=f(0) \psi+g
$$

where $\psi$ is the Gaussian. Clearly, $g \in \mathcal{S}$, with $g(0)=0$. Thus $T f(0)=f(0)(T \psi)(0)+(T g)(0)$. But since $\psi$ is its own Fourier transform, have $T \psi(0)=\psi(0)=1$, and $T g(0)=0$ by the case done in the last para, we have $T f(0)=f(0)$ for all $f \in \mathcal{S}$. Finally, to deduce the result for all points, we just translate coordinates. That is, for $f \in \mathcal{S}$, and $a \in \mathbb{R}^{n}$, define $g(x)=f(x+a)$, so that $g(0)=f(a)$, and $g$ is also in $\mathcal{S}$. Then

$$
\begin{aligned}
\operatorname{cccf}(a)=g(0)=(T g)(0) & =\iint e^{-i \xi \cdot x} f(x+a) d x d \xi \\
& =\iint e^{-i \xi \cdot x} e^{i \xi \cdot a} f(x) d x d \xi \\
& =\int e^{i \xi \cdot a} \widehat{f}(\xi) d \xi \\
& =(\widehat{f})^{\wedge}(-a)=T f(a)
\end{aligned}
$$

This proves the first part of (i). The second part (about the inverse Fourier transform) follows immediately. Thus (i) is proved.

To see (iii), note that:

$$
\begin{aligned}
(\widehat{f} \widehat{g})(\xi) & =\iint e^{-i \xi \cdot x} f(x) e^{-i \xi \cdot y} g(y) d x d y \\
& =\iint e^{-i \xi \cdot(x-y)} f(x-y) e^{-i \xi \cdot y} g(y) d x d y \\
& =\iint e^{-i \xi \cdot x} f(x-y) g(y) d x d y=\widehat{f * g}(\xi)
\end{aligned}
$$

where we have used Fubini to get the last line, because the double integral is absolutely convergent (since $f, g \in \mathcal{S}$ ), and a change of variables in the second line. The second part of (iii) follows immediately from (i) by replacing $f$ and $g$ by $\widehat{f}$ and $\widehat{g}$ respectively.

It finally remains to prove (iv), the Plancherel Theorem. We denote the $L_{2}$ inner product of $f$ and $g$ by $(f, g)=\int f(x) \bar{g}(x) d x$, which is $\mathbb{C}$-linear in the first slot, and $\mathbb{C}$-antilinear in the second. We compute for $f, g \in \mathcal{S}:$

$$
\begin{aligned}
(f, \widehat{g})=\int f(x) \overline{\hat{g}}(x) d x & =\iint f(x) e^{i x \cdot y} \bar{g}(y) d y d x \\
& =\int\left(\int f(x) e^{i x \cdot y} d x\right) \bar{g}(y) d y \\
& =(\widehat{f}(-x), g)
\end{aligned}
$$

where Fubini is used to interchange the order integration in an absolutely convergent double integral. Replacing $\widehat{f}(-y)$ by $g(y)$, we have $\widehat{g}=f$ by (i) above, so that:

$$
(\widehat{g}, \widehat{g})=(g, g)
$$

which shows (using (i)) that the Fourier transform is a unitary map on $\mathcal{S}$ with respect to $L_{2}$-norm. Thus it extends to a unitary isomorphism of $L^{2}\left(\mathbb{R}^{n}\right)$, since $\mathcal{S}$ is dense in it. This proves the proposition.

### 1.3. Distributions.

Definition 1.3.1 (Distributions). We define a distribution $T$ on $\mathbb{R}^{n}$ to be an element of the topological vector space dual of $C_{c}^{\infty}=\mathcal{D}$. That is, $T$ is a linear functional on $C_{c}^{\infty}$ and continuous with respect to the topology on it. The space of all distributions on $\mathbb{R}^{n}$ is clearly a complex vector space, and denoted $\mathcal{D}^{\prime}$. Distributions $T \in \mathcal{D}^{\prime}$ which extend to a continuous linear functional on the larger space $\mathcal{S}$ are called tempered distributions, and the vector space of tempered distributions is denoted $\mathcal{S}^{\prime}$. Finally, distributions which extend as a continuous linear functional on all of $C^{\infty}=\mathcal{E}$ are called compactly supported distributions, and their vector space is denoted by $\mathcal{E}^{\prime}$. Clearly, we have the inclusions of vector spaces:

$$
\mathcal{E}^{\prime} \subset \mathcal{S}^{\prime} \subset \mathcal{D}^{\prime}
$$

Here are some basic examples:
Example 1.3.2. Let $f$ be a measurable and locally $L_{1}$ function on $\mathbb{R}^{n}$. Then $f$ defines the distribution $T_{f} \in \mathcal{D}^{\prime}$ by the formula:

$$
T_{f}(g)=\int_{\mathbb{R}^{n}} f(x) g(x) d x \quad \text { for } g \in C_{c}^{\infty}
$$

which makes sense since $g$ is compactly supported. By the way the topology is defined on $C_{c}^{\infty}$ and the dominated convergence theorem, it follows that $T_{f}\left(g_{n}\right) \rightarrow 0$ in $\mathbb{C}$ if $g_{n} \rightarrow 0$ in $C_{c}^{\infty}$.

Example 1.3.3. Let $f$ be a measurable function on $\mathbb{R}^{n}$ such that $(1+|x|)^{-N} f(x)$ is in $L_{1}\left(\mathbb{R}^{n}\right)$, for some $N \in \mathbb{N}$. Such functions are called tempered functions. Then defining $T_{f}$ by the same formula as in the example above, and letting $g \in \mathcal{S}$, we get a tempered distribution. The formula makes sense because, for the $N$ as above:

$$
\int_{\mathbb{R}^{n}} f(x) g(x) d x=\int_{\mathbb{R}^{n}}(1+|x|)^{-N} f(x)(1+|x|)^{N} g(x) d x
$$

and we have that the function $(1+x \mid)^{N} g(x)$ is bounded since $g \in \mathcal{S}$, and $(1+|x|)^{-N} f(x)$ is $L_{1}$. Again, the proof of its continuity is a consequence of the Dominated Convergence Theorem, and the topology defined earlier on $\mathcal{S}$. In particular, since $(1+|x|)^{-n-1}$ is integrable on $\mathbb{R}^{n}$, all polynomials, bounded continuous functions, or continuous functions with at most polynomial growth define tempered distributions.

Note that if we take a function like $f(x)=e^{x}$, it can be checked that this is a distribution which is not tempered, so the inclusion $\mathcal{S}^{\prime} \subset \mathcal{D}^{\prime}$ is strict.

Example 1.3.4. Let $f$ be a compactly supported function, and define an element $T_{f}$ of $\mathcal{E}^{\prime}$ via the same formula as in the above two examples, but $g \in \mathcal{E}$. It is checked easily that this is a compactly supported distribution.

If one wants to see a distribution which is not defined by a function, it is the very celebrated Dirac distribution of the next example.

Example 1.3.5 (Dirac distribution). Define the distribution $\delta_{a}$ by the formula:

$$
\delta_{a}(g)=g(a) \text { for } g \in C^{\infty}
$$

Again, it is trivial to check continuity, so that $\delta_{a} \in \mathcal{E}^{\prime}$.

Exercise 1.3.6. Show that the inclusion $\mathcal{E}^{\prime} \subset \mathcal{S}^{\prime}$ is also strict.

Remark 1.3.7. A locally $L^{1}$-function $f$ which defines a tempered distribution (via integration against $g \in \mathcal{S}$ ) need not be a tempered function in the sense of Example 1.3.3 above. For example, the locally $L_{1}$ function $e^{x} \sin e^{x}$ defines a tempered distribution on $\mathbb{R}$, (because it is the derivative of the bounded continuous function $-\cos e^{x}$, which is therefore a tempered distribution by 1.3 .3 above, and the fact proved in the next subsection that all derivatives of tempered distributions are tempered distributions). However, it is not a tempered function, as we check below. For each $N$, we have a $C_{N}>0$ such that:

$$
(1+|x|)^{-N}\left|e^{x}\right| \geq C_{N} e^{x / 2}
$$

for all $x \in[0, \infty)$, and thus we have an inequality of the integrals:

$$
\int_{\mathbb{R}}(1+|x|)^{-N}\left|e^{x} \sin e^{x}\right| d x \geq C_{N} \int_{1}^{\infty} e^{x / 2}\left|\sin e^{x}\right| d x \geq C_{N} \int_{0}^{\infty}\left|\frac{\sin y}{\sqrt{y}}\right| d y
$$

by a change of variables $y=e^{x}$. The right hand integral is infinite by comparing with the infinite series $\sum_{n} n^{-1 / 2}$. Some authors (e.g. Folland) define a tempered function to be a locally $L_{1}$ function which is a tempered distribution, to avoid this inconsistency.

We will see later after defining convolutions that if $f$ is a real-valued non-negative locally integrable function on $\mathbb{R}^{n}$, then it is a tempered function in our sense if it is a tempered distribution. The rapid oscillation of say $e^{x} \sin e^{x}$ which causes the problem above is thereby eliminated.

As a final example of a distribution which is not a function, we have:

Example 1.3.8. Fix a multi-index $\alpha$, and a point $a \in \mathbb{R}^{n}$. Then the higher derivative $D_{x \mid a}^{\alpha}$ at $a$ clearly maps $\mathcal{E} \rightarrow \mathbb{C}$ in a continuous fashion with respect to the given topology, and defines a compactly supported distribution. For $\alpha=(0,0, . .0)$, we recover the Dirac distribution. When we later define derivatives of distributions, we will see that this distribution is nothing but $\pm D_{x}^{\alpha} \delta_{a}$.

Definition 1.3.9 (Support of a distribution). For an open subset $U \subset \mathbb{R}^{n}$, we say that the distribution $T \in \mathcal{D}^{\prime}$ vanishes on $U$ if $T(f)=0$ for all $f$ with compact support in $U$. For example, the Dirac distribution $\delta_{a}$ vanishes on $\mathbb{R}^{n} \backslash\{a\}$. Similarly the distribution $D_{x \mid a}^{\alpha}$ vanishes on $\mathbb{R}^{n} \backslash\{a\}$. By using a partition of unity subordinate to an open covering $\left\{U_{i}\right\}_{i \in \Lambda}$, one easily sees that if a distribution $T$ vanishes on $U_{i}$ for each $i \in \Lambda$, then it vanishes on the union $U=\cup_{i \in \Lambda} U_{i}$. Hence there is a largest open set $U$ (possibly empty) on which a distribution $T$ vanishes. The complement of this open set is called support of $T$, and denoted $\operatorname{supp} T$.

Lemma 1.3.10 ( $\mathcal{E}^{\prime}$ and distributions of compact support). A distribution $T \in \mathcal{D}^{\prime}$ is in $\mathcal{E}^{\prime}$ iff $\operatorname{supp} T$ is compact. (Hence the terminology "compactly supported distribution" for elements of $\mathcal{E}^{\prime}$.)

Proof: Suppose $\operatorname{supp} T=K$ a compact set. Let $\psi \in \mathcal{D}$ be a compactly supported smooth real-valued function with $0 \leq \psi \leq 1$ and $\psi \equiv 1$ on $K$. For a function $\phi \in \mathcal{E}$, define:

$$
T(\phi)=T(\psi \phi)
$$

Note that this definition is independent of the cut-off function $\psi$ chosen above, for if $\psi_{1}$ is another cut-off function satisfying the same properties as $\psi$ above, then $\psi \phi-\psi_{1} \phi$ will be a smooth compactly supported function whose support lies in $K^{c}$, so that $\operatorname{supp} T=K$ will imply that $T(\psi \phi)=T\left(\psi_{1} \phi\right)$. Now if $\phi_{n} \rightarrow 0$ in $\mathcal{E}$, we have (on applying Leibniz formula for derivatives of products) that $\psi \phi_{n}$ are compactly supported with support contained in the fixed compact set $L:=\operatorname{supp} \psi$ for all $n$, and that $D^{\alpha}\left(\psi \phi_{n}\right) \rightarrow 0$ uniformly on $L$. Thus $\psi \phi_{n} \rightarrow 0$ in $\mathcal{D}$, and hence $T\left(\phi_{n}\right)=T\left(\psi \phi_{n}\right) \rightarrow 0$ since $T \in \mathcal{D}^{\prime}$. Thus $T \in \mathcal{E}^{\prime}$.

Conversely, suppose supp $T$ is not compact, so $T$ does not vanish on $\mathbb{R}^{n} \backslash \overline{B(0, n)}$ for each ball $B(0, n)$ of radius $n=1,2,,,,$. Thus there exists a function $\phi_{n}$ with compact support $K_{n} \subset \mathbb{R}^{n} \backslash \overline{B(0, n)}$ with

$$
T\left(\phi_{n}\right)=\lambda_{n} \neq 0 \quad \text { for } \quad n=1,2, \ldots
$$

Then it is trivial to verify that the functions $f_{n}:=\lambda_{n}^{-1} \phi_{n}$ converge to 0 in $\mathcal{E}$, since on each compact set $L \subset \mathbb{R}^{n}$, we have $f_{n} \equiv 0$ on $L$ for $n$ large enough. On the other hand $T\left(f_{n}\right)=\lambda_{n}^{-1} T\left(\phi_{n}\right)=1$ for all $n$, so that $T$ is not continuous on $\mathcal{E}$, and hence $T \notin \mathcal{E}^{\prime}$. The lemma follows.

More examples of distributions will emerge as soon as we define some basic operations on distributions. Since tempered distributions are the ones of interest to us, we will concentrate mainly on them.
1.4. New distributions out of old. The most important operation on distributions is that of differentiation. Historically, distributions were invented by Dirac, to differentiate functions which had singularities, i.e. points of non-differentiability. Dirac realised that these are not going to be functions, but it was possible to do some self-consistent manipulations with them, so he called them "generalised functions". It took another thirty years for Laurent Schwartz to rigorise these ideas mathematically, and thanks to him, every distribution can be differentiated to get another distribution

The starting point is to note that if $f \in \mathcal{E}$ and $g \in \mathcal{D}$, then we have

$$
\int_{\mathbb{R}^{n}} D_{x}^{\alpha} f(x) g(x) d x=(-1)^{|\alpha|} \int_{\mathbb{R}^{n}} f(x) D_{x}^{\alpha} g(x) d x
$$

by using integration by parts, and noting that $\lim _{|x| \rightarrow \infty} f g=0$ because of compact support of $g$. The same identity holds if $f \in \mathcal{E}$ and of slow (at most polynomial, for all derivatives) growth in $x$, and $g \in \mathcal{S}$. Thus it makes sense to make the:

Definition 1.4.1 (Derivative of a distribution). For $T \in \mathcal{D}^{\prime}$, define the distribution $D_{x}^{\alpha} T$ by:

$$
D_{x}^{\alpha} T(g)=(-1)^{|\alpha|} T\left(D_{x}^{\alpha} g\right) \quad g \in \mathcal{D}
$$

If $g_{n} \rightarrow 0$ in $\mathcal{D}$, then by definition, $D_{x}^{\alpha} g_{n} \rightarrow 0$ in $\mathcal{D}$ as well, and hence $D_{x}^{\alpha} T$ defined as above is a continuous linear functional on $\mathcal{D}$. Hence it is also in $\mathcal{D}^{\prime}$. The factor $(-1)^{|\alpha|}$ has been chosen for consistency with derivatives of smooth functions, i.e. if $f \in \mathcal{E}=C^{\infty}$, the distribution $T_{f}$ defined by $f$ will satisfy $D_{x}^{\alpha} T_{f}=T_{D_{x}^{\alpha} f}$, viz. it is the distribution defined by $D_{x}^{\alpha} f$ in view of the last paragraph. The derivative of a distribution is often called a distributional derivative.

Exercise 1.4.2. For a fixed $a \in \mathbb{R}$, consider the distribution defined by the locally $L_{1}$ Heaviside function:

$$
\begin{aligned}
f: \mathbb{R} & \rightarrow \mathbb{R} \\
x & \mapsto \\
& \text { for } x<a \\
x & \mapsto
\end{aligned} \text { for } x \geq a
$$

(This is just the indicator (or characteristic) function $\chi_{[a, \infty)}$ of $[a, \infty)$.) Show that the distributional derivative $\frac{d f}{d x}$ is the Dirac distribution $\delta_{a}$.

Exercise 1.4.3. For $T \in \mathcal{S}^{\prime}$ a tempered distribution, $D_{x}^{\alpha} T$ is also tempered. If $T \in \mathcal{E}^{\prime}$ is any compactly supported distribution, then so is $D_{x}^{\alpha} T$. For any distribution $T \in \mathcal{D}^{\prime}$, the support of the derivative obeys

$$
\operatorname{supp} D_{x}^{\alpha} T \subset \operatorname{supp} T
$$

Definition 1.4.4 (Multiplication by a smooth function). If $f \in \mathcal{E}$, then the linear multiplication mapping $\mathcal{D} \rightarrow \mathcal{D}$ defined by $g \mapsto f g$ is clearly continuous. Thus we may define for a distribution $T \in \mathcal{D}^{\prime}$ the product $f T$ by the formula:

$$
f T(g)=T(f g) \quad \text { for } g \in \mathcal{D}
$$

By the remark above, $f T$ is also a distribution. Likewise for $\mathcal{E}$, the mapping $\mathcal{E} \rightarrow \mathcal{E}$ defined by $g \mapsto f g$ is continuous, and we can again define $f T$ as a compactly supported distribution for $T \in \mathcal{E}^{\prime}$ a compactly supported distribution by the same procedure as above.

The story for tempered distributions is different. Multiplication by an arbitrary smooth function $f$ does not send the Schwartz space $\mathcal{S}$ to itself. The best we can do is to observe that if $f$ is a smooth function of slow growth (i.e. $\left|D_{x}^{\beta} f\right|<C_{\beta}(1+|x|)^{N_{\beta}}$ for each $\beta$ ), then $g \mapsto f g$ is a continuous linear operator $\mathcal{S} \rightarrow \mathcal{S}$. Hence, by the procedure above, we can define $f T$ for $T \in \mathcal{S}^{\prime}$ and $f$ a smooth function of slow growth.

Finally, we come to convolution of functions and distributions. For a function $g$, we define:

$$
g^{x}(y):=g(x-y)
$$

so that for smooth functions $f, g$, their convolution (whenever it is defined) maybe expressed as

$$
(f * g)(x)=\int_{\mathbb{R}^{n}} f(y) g^{x}(y) d y
$$

By a change of variables, $f * g=g * f$. It is clear that the linear mapping $g \mapsto g^{x}$ is a continuous map which takes $\mathcal{E}$ to $\mathcal{E}, \mathcal{D}$ to $\mathcal{D}$ and $\mathcal{S}$ to $\mathcal{S}$. Taking our cue from this, it is natural to make the following definition:

Definition 1.4.5 (Convolution of a distribution with a function). Let $T \in \mathcal{D}$ be a distribution, and $f$ a smooth function of compact support. Then define the function $f * T$ by the formula:

$$
(f * T)(x)=T\left(f^{x}\right) \quad \text { for } f \in \mathcal{E}
$$

Similarly, if $T \in \mathcal{S}^{\prime}$ is a tempered distribution and $f \in \mathcal{S}$, or if $T \in \mathcal{E}^{\prime}$ is a compactly supported distribution and $f \in \mathcal{E}$ is any smooth function. These restrictions are natural, in view of the fact that even functions $f, g$ need to obey some decay conditions in order to be convolved.

Example 1.4.6 (Convolution with the Dirac distribution). Let $g \in \mathcal{E}$ be any smooth function, and $\delta_{0}$ be the Dirac distribution. Then the convolution $g * \delta_{0}$ is the function $g$. (This shows that the identity element for the convolution product is a distribution). For, by definition,

$$
\left(g * \delta_{0}\right)(x)=\delta_{0}\left(g^{x}\right)=g^{x}(0)=g(x)
$$

Lemma 1.4.7. Whenever it makes sense by the definition above, the convolution $f * T$ is a smooth function. Furthermore, we have the identities:

$$
D^{\alpha}(f * T)=f * D^{\alpha} T=D^{\alpha} f * T
$$

Proof: We just prove it for the first partial derivative with respect to $x_{1}$, viz. $\partial_{1}=i D_{1}$. Let $e_{1}$ denote the unit vector $(1,0, . ., 0)$, and the case of $f \in \mathcal{D}$ and $T \in \mathcal{D}^{\prime}$. For a smooth function $f \in \mathcal{E}$, we have the Taylor formula:

$$
f^{x+h e_{1}}(y)-f^{x}(y)=f\left(x+h e_{1}-y\right)-f(x-y)=h g(x, h, y)+h^{2} r(x, h, y)
$$

where $g$ and $r$ are smooth in all the variables, and $g(x, 0, y)=\left(\partial_{1} f\right)(x-y)=\left(\partial_{1} f\right)^{x}(y)$. Because the supremum norm

$$
\sup _{y \in K,|h| \leq \epsilon}|r(x, h, y)| \leq C(K)
$$

for any compact set $K \subset \mathbb{R}^{n}$, it follows that, for a fixed $x$, and as a function of $y$ :

$$
\lim _{h \rightarrow 0}\left(\frac{f^{x+h e_{1}}-f^{x}}{h}\right) \rightarrow g(x, 0,-)=\left(\partial_{1} f\right)^{x}
$$

uniformly on compact sets. Similarly all $y$-derivatives of the functions $g_{h}:=\frac{f^{x+h e}-f^{x}}{h}$ converge uniformly to the corresponding derivatives of $\left(\partial_{1} f\right)^{x}$ on all compact sets as $h \rightarrow 0$. If the function $f$ is in $\mathcal{D}$, and compactly supported in $K$ say, then it is easy to check that for all $|h|<1$, the functions $g_{h}$ are all supported in the fixed compact set $K^{\prime}=x-K+\overline{B(0,1)}$, and $g_{h} \rightarrow\left(\partial_{1} f\right)^{x}$ as $h \rightarrow 0$ in $\mathcal{D}$ as well. Thus by the continuity and linearity of $T \in \mathcal{D}^{\prime}$ we have:

$$
\partial_{1}(f * T)=\lim _{h \rightarrow 0}\left(\frac{T\left(f^{x+h e_{1}}\right)-T\left(f^{x}\right)}{h}\right)=T\left[\lim _{h \rightarrow 0}\left(\frac{f^{x+h e_{1}}-f^{x}}{h}\right)\right]=T\left(\left(\partial_{1} f\right)^{x}\right)=\partial_{1} f * T
$$

Also note that we have:

$$
\left(\partial_{1} f\right)^{x}(y)=\left(\partial_{y_{1}} f\right)^{x}(y)=\left(\partial_{y_{1}} f\right)(x-y)=-\partial_{y_{1}}\left(f^{x}\right)(y)
$$

so that:

$$
\partial_{1} f * T=T\left(\left(\partial_{1} f\right)^{x}\right)=-T\left(\partial_{1} f^{x}\right)=\left(\partial_{1} T\right)\left(f^{x}\right)=f * \partial_{1} T
$$

by the definition of derivative of a distribution. This proves the lemma.

## Exercise 1.4.8.

(i): Let $T \in \mathcal{D}^{\prime}$ be a distribution, and let

$$
\rho(x, y): \mathbb{R}^{n} \times \mathbb{R}^{n} \rightarrow \mathbb{C}
$$

be a smooth function of compact support in $\mathbb{R}^{2 n}$, say $\operatorname{supp} \rho \subset K \times K$ for some compact $K \subset \mathbb{R}^{n}$. Thus, the function $\int_{x \in \mathbb{R}^{n}} \rho(x, y) d x$ is a smooth function (of $\left.y\right)$ in $C_{c}^{\infty}\left(\mathbb{R}^{n}\right)$, which is supported in $K$. Then show that

$$
T\left(\int_{x \in \mathbb{R}^{n}} \rho(x, y) d x\right)=\int_{\mathbb{R}^{n}} T(\rho(x, y)) d x
$$

where on the right hand side, $T$ is operating on the function $\rho(x, y)$ considered as a function of $y$, and thus $T(\rho(x, y))$ is a function of $x$. (Hint: Find a sequence of Riemann sums, which are functions of $y$, say $S_{n}(y):=\sum_{j} \rho\left(x_{j}, y\right) \Delta_{j}$ where the $\Delta_{j}$ 's are the volumes of cubes of side $\frac{1}{n}$ covering the compact set $K$, $x_{j}$ the centre of $\Delta_{j}$, and show that $S_{n} \rightarrow \int \rho(x,-) d x$ inside $\mathcal{D}\left(\mathbb{R}^{n}\right)$, and use the continuity and linearity of $T$.)
(ii): If $f$ and $g$ are compactly supported functions in $\mathcal{D}$, show that $\rho(x, y):=f(x) g(y-x)$ is in $C_{c}^{\infty}\left(\mathbb{R}^{n} \times \mathbb{R}^{n}\right)$, and conclude that $f * g$ is a smooth compactly supported function, and by (i) above, we further have:

$$
T(f * g)=\int_{x} f(x) T\left(\widetilde{g}^{x}\right) d x=(\widetilde{g} * T)(f)
$$

where $\widetilde{g}(z):=g(-z)$ and $\widetilde{g} * T$ is a function being regarded as a distribution.

Here is an important application of convolutions of functions with distributions. We can use compactly supported approximate identities (see the Lemma 1.2.3) to approximate any distribution $T \in \mathcal{D}^{\prime}$ by smooth functions of compact support. First we need a topology on $\mathcal{D}^{\prime}$ to make sense of the notion of approximation.

Definition 1.4.9 (Weak-star topology on $\mathcal{D}^{\prime}$ ). Say that a sequence of distributions $T_{n} \rightarrow 0$ in $\mathcal{D}^{\prime}$ if for each $\phi \in \mathcal{D}$, the sequence $T_{n}(\phi) \rightarrow 0$. This is the topology of pointwise convergence in any dual vector space, and is usually called the weak-star topology. On the subspaces $\mathcal{S}^{\prime}$ (resp. $\mathcal{E}^{\prime}$ ) of tempered (resp. compactly supported) distributions, we induce the subspace topology from this weak star topology on $\mathcal{D}^{\prime}$.

Proposition 1.4.10 (Approximation of distributions by compactly supported functions). The space of smooth compactly supported functions $\mathcal{D}=C_{c}^{\infty}\left(\mathbb{R}^{n}\right)$ is dense in the topological vector space $\mathcal{D}^{\prime}$ of distributions on $\mathbb{R}^{n}$.

Proof: We first make a remark about convolution with approximate identities (see the Lemma 1.2.3). Take a real-valued $\phi \in C_{c}^{\infty}$, with $\phi(x) \geq 0$ all $x \in \mathbb{R}^{n}$ and $\int \phi(x) d x=1, \operatorname{supp} \phi \subset \overline{B(0,1)}$ and $\phi(x)=\phi(-x)$ (even function). Then $\operatorname{supp} \phi_{\epsilon} \subset \overline{B(0, \epsilon)}$, and for a compactly supported function $g \in C_{c}^{\infty}$ with $\operatorname{supp} g=K$ a compact set, we have:

$$
\operatorname{supp}\left(g * \phi_{\epsilon}\right) \subset K^{\epsilon}:=\{x: d(x, K) \leq \epsilon\}
$$

This is because if $x$ is outside the set $K^{\epsilon}$ on the right, $(x-y)$ will lie outside $K$ for all $y \in \overline{B(0, \epsilon)}$, and hence $g(x-y)$ will be zero. For $y \notin \overline{B(0, \epsilon)}, \phi_{\epsilon}(y)$ will be zero. Thus the product $g(x-y) \phi_{\epsilon}(y)$ will be identically zero, and hence

$$
g * \phi_{\epsilon}(x)=\int_{\mathbb{R}^{n}} g(x-y) \phi_{\epsilon}(y)=0
$$

for $x \notin K^{\epsilon}$.
Now let $T \in \mathcal{D}^{\prime}$ be a distribution. Let $\psi_{j} \geq 0$ be compactly supported cutoff functions which are identically 1 on $\overline{B(0, j)}$ and identically zero outside $V_{j}:=\overline{B(0, j+1)}$, say. We claim that the distributions $\psi_{j} T$ converge to $T$ in $\mathcal{D}^{\prime}$. Indeed, for a fixed $g \in \mathcal{D}$

$$
\psi_{j} T(g)=T\left(\psi_{j} g\right) \rightarrow T(g)
$$

because $\psi_{j} g \equiv g$ for $j$ large enough, $g$ being compactly supported. So by the definition of the weak star topology, we have $\psi_{j} T \rightarrow T$.

Now we claim that the function $\phi_{\epsilon} *\left(\psi_{j} T\right)$ is a smooth function compactly supported in $V_{j}^{\epsilon}$. We already know by the Lemma 1.4.7 above that the convolution $\phi_{\epsilon} *\left(\psi_{j} T\right)$ is a smooth function. Clearly, it is a compactly supported function iff it is a compactly supported distribution. To show that $\phi_{\epsilon} * \psi_{j} T$ vanishes on the complement of $V_{j}^{\epsilon}$, let $g \in \mathcal{D}$ be a smooth function with $\operatorname{supp} g=L$ a compact set, and satisfying $L \cap V_{j}^{\epsilon}=\phi$. Then, by (ii) of Exercise 1.4.8 above:

$$
\begin{equation*}
\left(\phi_{\epsilon} * \psi_{j} T\right)(g)=\left(\widetilde{\phi}_{\epsilon} * \psi_{j} T\right)(g)=\psi_{j} T\left(\phi_{\epsilon} * g\right)=T\left(\psi_{j}\left(\phi_{\epsilon} * g\right)\right) \tag{1}
\end{equation*}
$$

By the first para above, $\operatorname{supp}\left(\phi_{\epsilon} * g\right)=L^{\epsilon}$. The support of $\psi_{j}$ is contained in $V_{j}$. Since $L \cap V_{j}^{\epsilon}=\phi$, we have $L^{\epsilon} \cap V_{j}=\phi$, so the function of $y$ given by $\psi_{j}\left(\phi_{\epsilon} * g\right)$ above is the identically zero function, and so $T$ applied to it is therefore zero. This shows that $\phi_{\epsilon} * \psi_{j} T$ is a distribution compactly supported in $V_{j}^{\epsilon}$, and hence a smooth function of compact support.

We now claim that for a fixed $j$, the family of distributions (compactly supported smooth functions by the above) $\phi_{\epsilon} * \psi_{j} T$ converge to $\psi_{j} T$ in $\mathcal{D}^{\prime}$ as $\epsilon \rightarrow 0$. By the Lemma 1.2 .3 and the fact that $\operatorname{supp} \psi_{j}=V_{j}$ a compact set, we have $\psi_{j}\left(\phi_{\epsilon} * g\right) \rightarrow \psi_{j} g$ uniformly on $V_{j}$ as $\epsilon \rightarrow 0$. Hence the right hand side of the equation (1) above converges to $T\left(\psi_{j} g\right)$ as $\epsilon \rightarrow 0$ by the continuity of $T$, and the claim follows.

Since $\psi_{j} T \rightarrow T$ in $\mathcal{D}^{\prime}$ as $j \rightarrow \infty$, and the compactly supported smooth functions $\phi_{\epsilon} * \psi_{j} T \rightarrow \psi_{j} T$ as $\epsilon \rightarrow 0$, it follows that $C_{c}^{\infty}\left(\mathbb{R}^{n}\right)$ is dense in $\mathcal{D}^{\prime}$.

Now we come to one of the chief reasons why the Schwartz space $\mathcal{S}$ and tempered distributions were introduced. We have already observed in (ii) of the Proposition 1.2.5 that

$$
\xi^{\alpha} D_{\xi}^{\beta} \widehat{f}= \pm\left(D_{x}^{\alpha}\left(x^{\beta} f\right)\right)^{\wedge} \quad \text { for } \quad f \in \mathcal{S}
$$

Hence if $f_{n} \rightarrow 0$ as $n \rightarrow \infty$ in $\mathcal{S}, x^{\beta} D_{x}^{\alpha} f_{n} \rightarrow 0$ uniformly on $\mathbb{R}^{n}$, for all $\alpha, \beta$. By Leibnitz's rule for the derivatives of products, it follows that $D_{x}^{\alpha}\left(x^{\beta} f_{n}\right) \rightarrow 0$ uniformly on $\mathbb{R}^{n}$. Thus $\left\|D_{x}^{\alpha}\left(x^{\beta} f_{n}\right)\right\|_{1} \rightarrow 0$. By the fact that $\|\widehat{g}\|_{\infty} \leq\|g\|_{1}$ for $g \in \mathcal{S}$ and the equation above it follows that

$$
\left.\left\|\xi^{\alpha} D_{\xi}^{\beta} \widehat{f}_{n}\right\|_{\infty}=\| D_{x}^{\alpha}\left(x^{\beta} f\right)\right)^{\wedge}\left\|_{\infty}=\right\| D_{x}^{\alpha}\left(x^{\beta} f_{n}\right) \|_{1} \rightarrow 0 \quad \text { as } n \rightarrow \infty
$$

That is, $\widehat{f}_{n} \rightarrow 0$ in $\mathcal{S}$ as $n \rightarrow \infty$. Thus, with the topology introduced earlier on $\mathcal{S}$, we have:

$$
\uparrow \mathcal{S} \rightarrow \mathcal{S}
$$

is a continuous linear map of topological vector spaces. Hence it makes sense to make the following:

Definition 1.4.11 (Fourier transform of a tempered distribution). Let $T \in \mathcal{S}^{\prime}$ be a tempered distribution. Define the Fourier transform $\widehat{T}$ by the formula:

$$
\widehat{T}(g)=T(\widehat{g}) \quad \text { for } \quad g \in \mathcal{S}
$$

By the remarks above, $\widehat{T}$ is also a tempered distribution. We leave it as an easy exercise to check that this definition is consistent with the definition for functions, i.e. for an $L^{1}$-function (which defines the tempered distribution $T_{f}$ via integration as indicated in the Example 1.3.3), then we have $T_{\widehat{f}}=\widehat{T_{f}}$. (Just mimic the proof of (iv) of Proposition 1.2.5 without the complex conjugation).

Analogously, since the inverse Fourier transform and transform differ by reflection of the function, we define the inverse Fourier transform $T^{\vee}$ of a tempered distribution $T$ by the formula:

$$
T^{\vee}(g)=T\left(g^{\vee}\right) \quad \text { for } \quad g \in \mathcal{S}
$$

We have the following proposition about the distributional Fourier transform. .

Proposition 1.4.12. Let $T \in \mathcal{S}^{\prime}$ be a tempered distribution. Then the Fourier transform ${ }^{\wedge} \mathcal{S}^{\prime} \rightarrow \mathcal{S}^{\prime}$ satisfies:
(i): The map ${ }^{\wedge}$ is a continuous linear isomorphism of $\mathcal{S}^{\prime}$ of period 4, and we have

$$
(\widehat{T})^{\vee}=T \quad \text { for all } T \in \mathcal{S}^{\prime}
$$

(ii): For a polynomial $P=P\left(X_{1}, . ., X_{n}\right)$, we have:

$$
(P(D) T)^{\wedge}=P(\xi) \widehat{T}, \quad(P(x) T)^{\wedge}=P(-D) \widehat{T}
$$

(iii): If $f \in L_{1}\left(\mathbb{R}^{n}\right)$, then $T_{\widehat{f}}=\widehat{T_{f}}$.
(iv): $(\phi * T)^{\wedge}=\widehat{\phi} \widehat{T}$ for $\phi \in \mathcal{S}$.

Proof: If $T_{n} \rightarrow T$ in $\mathcal{S}^{\prime}$, we have by definition that $T_{n}(\widehat{g}) \rightarrow T(\widehat{g})$ for each $g \in \mathcal{S}$. That is, $\widehat{T}_{n}(g) \rightarrow \widehat{T}(g)$, for each $g \in \mathcal{S}$, which again, by definition, implies $\widehat{T}_{n} \rightarrow \widehat{T}$ in $\mathcal{S}^{\prime}$. This shows that the Fourier transform is a continuous map. The rest of (i),(ii) and (iii) follow immediately by applying the relevant parts of the Proposition 1.2.5.

To see (iv), note that:

$$
(\phi * T)^{\wedge}(g)=(\phi * T)(\widehat{g})=T(\widetilde{\phi} * \widehat{g})
$$

by (ii) of Exercise 1.4.8. But by (iii) of Proposition 1.2 .5 , we have $\widetilde{\phi} * \widehat{g}=\widehat{\widehat{\phi}} * \widehat{g}=(\widehat{\phi} g)^{\wedge}$, so the last expression above is precisely $(\widehat{T})(\widehat{\phi} \widehat{g})=(\widehat{\phi} \widehat{T})(g)$. The proposition follows.

To deduce some more crucial facts about $\widehat{T}$, we need an elementary but very useful lemma about "locally convex topological vector spaces".

Lemma 1.4.13. Let $V$ be a topological vector spaces whose topology is defined by a "family of seminorms" $\left\{p_{\alpha}\right\}_{\alpha \in \Lambda}$, viz., a sequence $x_{n} \in V$ converges to zero iff $p_{\alpha}\left(x_{n}\right) \rightarrow 0$ for all $\alpha \in \Lambda$. Then a linear map $T: V \rightarrow \mathbb{C}$ is continuous iff there exists a constant $C>0$ and a finite subfamily $p_{\alpha_{1}}, . ., p_{\alpha_{k}}$ of seminorms such that:

$$
|T x| \leq C \sum_{i=1}^{k} p_{\alpha_{j}}(x) \text { for all } x \in V
$$

Proof: We define the "semiball" (?) in $V$ around 0 with respect to the seminorm $p_{\alpha}$ in the obvious manner:

$$
B_{\alpha}(0, \epsilon)=\left\{x \in V: p_{\alpha}(x)<\epsilon\right\}
$$

and note that by the definition of a seminorm all of these are convex sets, and since each $p_{\alpha}$ is continuous, they are open. Hence their finite intersections are also convex, open, and contain $x$. Define a new topology $\mathcal{T}$ on $V$ by declaring a neighbourhood base of 0 to be the family of finite intersections:

$$
\mathcal{N}(0):=\left\{\cap_{i=1}^{k} B_{\alpha_{i}}\left(0, \epsilon_{i}\right): \epsilon_{i}>0, \quad\left\{\alpha_{1}, . ., \alpha_{k}\right\} \subset \Lambda, k=1,2, \ldots,\right\}
$$

and the neighbourhood base around $x$ by $\mathcal{N}(x):=x+\mathcal{N}(0)$. It is clear that if $p_{\alpha}\left(x_{n}\right) \rightarrow 0$ for all $\alpha \in \Lambda$, then $x_{n} \rightarrow 0$ in the topology $\mathcal{T}$, because $x_{n}$ will eventually lie in every basic neighbourhood. On the other hand, if there exists an $\alpha \in \Lambda$ such that $p_{\alpha}\left(x_{n}\right)$ does not converge to zero, then there exists an $\epsilon>0$ and some subsequence $x_{n_{k}}$ such that $p_{\alpha}\left(x_{n_{k}}\right)>\epsilon$ for all $k$. That is, $x_{n_{k}} \notin B_{\alpha}(x, \epsilon)$ for all $k$, so the sequence $\left\{x_{n}\right\}$ will fail to eventually belong to this open neighbourhood $B_{\alpha}(0, \epsilon)$, and hence does not converge to 0 in the topology $\mathcal{T}$. Thus $\mathcal{T}$ is exactly the topology defined by the family of seminorms $\left\{p_{\alpha}\right\}_{\alpha \in \Lambda}$.

Since $T$ is continuous, there is an open neighbourhood $U$ of 0 such that $|T x|<1$ for all $x \in U$. Since $\mathcal{N}(0)$ is a neighbourhood basis of 0 , we may assume without loss of generality that $U=\cap_{i=1}^{k} B_{\alpha_{i}}\left(0, \epsilon_{i}\right)$. Set $\epsilon=\min _{1 \leq i \leq k}\left\{\epsilon_{i}\right\}$ and $C=\epsilon^{-1}$.

Let $x \in V$. If $p_{\alpha_{j}}(x)=0$ for all $j=1, . ., k$, then by the definition of $U$, it follows that $\lambda x \in U$ for all $\lambda>0$, and by the choice of $U$ it follows that $|T(\lambda x)|=\lambda|T x|<1$ for all $\lambda>0$, so that $T x=0$, and certainly

$$
|T x|=0 \leq C \sum_{i=1}^{k} p_{\alpha_{i}}(x)
$$

for $C$ as above. On the other hand, if $p_{\alpha_{j}}(x)>0$ for some $j$, observe that $|y|=\epsilon x / \sum_{i=1}^{k} p_{\alpha_{i}}(x)$ satisfies $p_{\alpha_{j}}(y)<\epsilon$ for all $j=1, . ., k$, so that $y \in U$ and $|T y|<1$. Which is the same as saying that:

$$
|T x|<(\epsilon)^{-1} \sum_{i=1}^{k} p_{\alpha_{i}}(x)
$$

So, since $C=\epsilon^{-1}$, we have the desired inequality in both cases, i.e. for all $x \in V$, and the lemma follows.

Remark 1.4.14 (: Caution!). The topology of the topological vector space $\mathcal{E}$ is determined by the family of seminorms

$$
\left\{p_{\alpha, K}: p_{\alpha, K}(f):=\sup _{x \in K}\left|D_{x}^{\alpha} f\right|, \alpha \text { a multi-index, } K \text { a compact subset of } \mathbb{R}^{n}\right\}
$$

Similarly, the Schwartz space $\mathcal{S}$ is defined by the family of seminorms:

$$
\left\{p_{\alpha, \beta}: \quad p_{\alpha, \beta}(f):=\sup _{x \in \mathbb{R}^{n}}\left|x^{\alpha} D_{x}^{\beta} f\right|, \quad \alpha, \beta \text { multi-indices }\right\}
$$

However, the topology of $\mathcal{D}$ is not given by the family of seminorms which define $\mathcal{E}$. As we noted in (iii) of the Definition 1.1.1, if we take a fixed compactly supported function $\psi \neq 0$ supported on $[-1,1] \subset \mathbb{R}$, then define $f_{n}(x)=\psi(x-n)$ by translating, we have $p_{\alpha, K} f_{n} \rightarrow 0$ for each $\alpha$ and $K$, but $f_{n} \nrightarrow 0$ in $\mathcal{D}$.

Proposition 1.4.15. Let $T \in \mathcal{E}^{\prime}$ be a compactly supported distribution. For $\xi \in \mathbb{R}^{n}$, denote $e_{\xi}(x):=e^{i \xi \cdot x}=$ $e_{x}(\xi)$. Then the tempered distribution $\widehat{T}$ is the function:

$$
\widehat{T}(\xi)=T\left(e_{-\xi}\right)
$$

It is a smooth function of slow growth (see the Definition 1.4.4).

Proof: Let us first check the identity above, whose right side, viz the function $\nu(\xi):=T\left(e_{-\xi}\right)$ makes sense because $T$ is in $\mathcal{E}^{\prime}$. Then let $K:=\operatorname{supp} T$ be the compact support of $T$ (in view of the Lemma 1.3.10), and let $\psi(\xi)$ be a compactly supported function which is identically 1 on $K$. It is trivial to check that $\psi T=T$. Now let $g \in \mathcal{D}$ be a compactly supported function. Then

$$
\begin{aligned}
\widehat{T}(g) & =\psi T(\widehat{g})=T\left[\psi(\xi) \int g(x) e_{-\xi}(x) d x\right]=T\left[\int \psi(\xi) g(x) e_{-x}(\xi) d x\right] \\
& =\int g(x) T\left(\psi e_{-x}\right) d x=\int g(x)(\psi T)\left(e_{-x}\right) d x \\
& =\int g(x) T\left(e_{-x}\right) d x=\int g(x) \nu(x) d x=T_{\nu}(g)
\end{aligned}
$$

by using (i) of the Exercise 1.4.8 applied to the compactly supported function $\rho(x, \xi)=\psi(\xi) g(x) e_{-x}(\xi)$.
Now the smoothness of the function $\nu(\xi)$ easily follows by applying continuity of $T$, and that $T$ is compactly supported so acts on all smooth functions. To check slow growth, we first note that for the family of seminorms $\left\{p_{\alpha, L}: \alpha\right.$ a multi-index and $L$ a compact subset of $\left.\mathbb{R}^{n}\right\}$
which define the topology of $\mathcal{E}$, we have:

$$
p_{\alpha, L}\left(x^{\beta} e_{-\xi}\right)=\sup _{x \in L}\left|D_{x}^{\alpha}\left(x^{\beta} e^{-i \xi \cdot x}\right)\right| \leq C(L)(1+|\xi|)^{N(\alpha, \beta)}
$$

By the Lemma 1.4.13, since $T$ is continuous, there exists a finite subfamily $p_{\alpha_{j}, L_{j}}$ such that

$$
\left|D_{\xi}^{\beta} \nu(\xi)\right|=\left| \pm T\left(D_{\xi}^{\beta} e_{-\xi}\right)\right|=\left|T\left(x^{\beta} e_{-\xi}\right)\right| \leq C \sum_{j=1}^{k} p_{\alpha_{j}, L_{j}}\left(x^{\beta} e_{-\xi}\right) \leq C \sum_{j=1}^{k} C\left(L_{j}\right)(1+|\xi|)^{N\left(\alpha_{j}, \beta\right)}
$$

which is clearly bounded by $C(1+|\xi|)^{N}$ for $N=\max _{j} N\left(\alpha_{j}, \beta\right)$. The proposition follows.
Since the constant function 1 is a locally $L_{1}$ function satisfying $(1+|x|)^{-N} .1 \in L_{1}\left(\mathbb{R}^{n}\right)$ for all $N>n$, by the Example 1.3.3 it is a tempered distribution. The Dirac distribution is in fact a compactly supported distribution, and hence a tempered distribution. Thus it makes sense to take the Fourier transforms of these distributions. Indeed we have the:

Corollary 1.4.16. The Dirac distribution $\delta_{0}$ and the constant function 1 are Fourier transforms of each other.

Proof: By the above Proposition 1.4.15, we have

$$
\widehat{\delta}_{0}(\xi)=\delta_{0}\left(e_{-\xi}\right)=1
$$

for all $\xi$. The fact that $\widehat{1}=\delta_{0}$ then follows from the Fourier inversion formula in (i) of the Proposition 1.4.12, since $\delta$ and 1 are invariant under the reflection $x \mapsto-x$. One can check it directly as well, for if $g \in \mathcal{S}$, we have:

$$
\widehat{1}(g)=1(\widehat{g})=\int \widehat{g} d x=\widehat{\widehat{g}}(0)=g(-0)=g(0)=\delta_{0}(g)
$$

Exercise 1.4.17. Using (ii) of the Proposition 1.4.12, show that polynomials in $\xi=\left(\xi_{1}, . ., \xi_{n}\right)$ are exactly the Fourier transforms of tempered distribution defined as finite linear combinations of derivatives of the Dirac distribution $\delta_{0}$, namely distributions $T$ of the form:

$$
T=\sum_{i=1}^{k} c_{k} D_{x}^{\alpha_{i}} \delta_{0}
$$

where $\alpha_{i}$ are some multi-indices, and $c_{i} \in \mathbb{C}$.

Indeed, we have the following interesting characterisation of distributions whose support is a point.

Proposition 1.4.18 (Distributions with point support). Let $a \in \mathbb{R}^{n}$, and let $T \in \mathcal{D}^{\prime}$ with $\operatorname{supp} T=\{a\}$. Then

$$
T=\sum_{i=1}^{k} c_{k} D_{x}^{\alpha_{i}} \delta_{a}
$$

where $\delta_{a}$ is the Dirac distribution at $a, \alpha_{i}$ are some multi-indices, and $c_{i} \in \mathbb{C}$.

Proof: By translation, we may assume that $a=0$. Let $\psi \in \mathcal{D}$ be a cutoff function such that $\psi \geq 0, \psi \equiv 1$ on $B\left(0, \frac{1}{2}\right)$ and $\psi \equiv 0$ outside $B(0,1)$. Since $T$ is supported in the point $\{0\}$, it follows that $T((1-\psi) \phi)=0$ for all $\phi \in \mathcal{E}$, and hence $T(\phi)=T(\psi \phi)$ for all $\phi \in \mathcal{E}$. Note that by Leibnitz's formula for the derivatives of a product, we have for a compact set $K$ and $\alpha$ a multi-index, the inequality:

$$
p_{\alpha, K}(\psi \phi)=\sup _{x \in K}\left|D^{\alpha}(\psi \phi)\right| \leq C \sum_{|\beta| \leq|\alpha|} \sup _{\|x\| \leq 1}\left|D^{\beta} \phi\right|
$$

where $C$ depends on $\sup _{\|x\| \leq 1}\left|D^{\gamma} \psi\right|$ for various $|\gamma| \leq|\alpha|$.
Combining the above fact with the Lemma 1.4.13, we have an inequality:

$$
\begin{equation*}
|T(\phi)|=|T(\psi \phi)| \leq C \sum_{i=1}^{k} p_{\alpha_{i}, K_{i}}(\psi \phi) \leq C \sum_{|\alpha| \leq N} \sup _{\|x\| \leq 1}\left|D^{\alpha} \phi(x)\right| \quad \text { for all } \phi \in \mathcal{E} \tag{2}
\end{equation*}
$$

where $N=\max _{1 \leq i \leq k}\left|\alpha_{i}\right|$. Now we make the following:
Claim: Let $\phi \in \mathcal{E}$ such that $D^{\alpha} \phi(0)=0$ for all $|\alpha| \leq N$. Then $T(\phi)=0$.

Consider the sequence of functions $\phi_{k} \in \mathcal{E}$ defined by

$$
\phi_{k}(x)=\phi(x)(1-\psi(k x)), \quad x \in \mathbb{R}^{n}
$$

where $\psi$ is the cut-off function defined in the first paragraph. The fact that $D^{\alpha} \phi(0)=0$ for $|\alpha| \leq N$ will imply by Taylor's formula for $D^{\beta} \phi$ around the origin that there exists a $\delta>0$ such that

$$
\left|D^{\beta} \phi(x)\right| \leq C\|x\|^{N+1-|\beta|} \quad \text { for all } \quad\|x\|<\delta, \quad|\beta| \leq N
$$

Note that the function $\psi(k x)$ and all its derivatives are supported in the ball $B(0,1 / k)$, and for $k$ large enough, this ball is contained in $B(0, \delta)$.

Now let $\alpha$ be a multi-index such that $|\alpha| \leq N$. Then we have

$$
\left|D^{\alpha}(\phi(x) \psi(k x))\right|=0 \text { for }\|x\|>\frac{1}{k}, \text { and all } k
$$

On the other hand, for $\|x\|<\frac{1}{k}$, and $k$ large enough so that $\frac{1}{k}<\delta$, we have:

$$
\begin{aligned}
\left|D^{\alpha}(\phi(x) \psi(k x))\right| & \leq C \sum_{|\beta| \leq|\alpha|}\left|D^{\beta} \phi(x) \| D^{\alpha-\beta} \psi(k x)\right| \\
& \leq C \sum_{|\beta| \leq|\alpha|}\|x\|^{N+1-|\beta|} k^{|\alpha|-|\beta|}\left\|D^{\alpha-\beta} \psi\right\|_{\infty} \\
& \leq C \sum_{|\beta| \leq|\alpha|}\left(k^{-1}\right)^{N+1-|\beta|} k^{|\alpha|-|\beta|} \\
& \leq C k^{|\alpha|-N-1} \leq C k^{-1}
\end{aligned}
$$

Summing up, we have:

$$
\sup _{x \in \mathbb{R}^{n}}\left|D^{\alpha}\left(\phi_{k}(x)-\phi(x)\right)\right|=\sup _{x \in \mathbb{R}^{n}}\left|D^{\alpha}(\phi(x) \psi(k x))\right| \leq C k^{-1} \quad \text { for } \quad|\alpha| \leq N \quad \text { and } k \gg 0
$$

Plugging this fact into the inequality (2) above, we find that: $\lim _{k \rightarrow \infty}\left|T \phi_{k}-T \phi\right|=0$, i.e.

$$
\lim _{k \rightarrow \infty} T\left(\phi_{k}\right)=T(\phi)
$$

Now note that $\phi_{k}$ are compactly supported in the region $\{1 / 2 k \leq\|x\|<\infty\}$, and hence compactly supported in $\mathbb{R}^{n} \backslash\{0\}$. Since $\operatorname{supp} T=\{0\}, T\left(\phi_{k}\right)=0$ for all $k$. Thus $T(\phi)=\lim _{k \rightarrow \infty} T\left(\phi_{k}\right)=0$ and our claim follows.

Now, to show that $T=\sum_{|\alpha| \leq N} c_{\alpha} D^{\alpha} \delta_{0}$, it is enough to show that the Fourier transform $\widehat{T}$ is a polynomial (see the preceding Exercise 1.4.17). By the Proposition 1.4.15, we have for the $N$ chosen above that

$$
\begin{aligned}
\widehat{T}(\xi)=T\left(e_{-\xi}\right) & =T\left(\sum_{0 \leq k \leq N} \frac{[-i(\xi \cdot x)]^{k}}{k!}+\phi\right) \\
& =\sum_{0 \leq|\alpha| \leq N}(-i)^{|\alpha|} \frac{\xi^{\alpha} T\left(x^{\alpha}\right)}{\alpha!}=\sum_{0 \leq|\alpha| \leq N} c_{\alpha} \xi^{\alpha}
\end{aligned}
$$

since $T(\phi)=0$ by the Claim above (all its derivatives of order $\leq N$ vanish at 0 ). This proves the proposition.

Finally, we have the following description of compactly supported distributions.

Proposition 1.4.19. Let $T \in \mathcal{E}^{\prime}$ be a compactly supported distribution. Then there exists a continuous function $g \in C_{0}\left(\mathbb{R}^{n}\right)$ such that:

$$
T=\sum_{i=1}^{k} c_{i} D_{x}^{\alpha_{i}} g
$$

Proof: By the Proposition 1.4.15, the Fourier transform $\widehat{T}$ is a smooth function, say $F$, of slow growth. That is, there exist $C>0$ and $N$ such that:

$$
|F(\xi)| \leq C(1+|\xi|)^{N}
$$

Since $\left(1+|\xi|^{2}\right)^{-s}$ is in $L^{1}\left(\mathbb{R}^{n}\right)$ for any $s>n / 2$, it follows that the function:

$$
G(\xi)=\left(1+|\xi|^{2}\right)^{-M} F(\xi)
$$

is in $L^{1}\left(\mathbb{R}^{n}\right)$ for $M=N+n$, say. So, by the Riemann-Lebesgue lemma (v) of Proposition 1.2.5 above, the function $g=G^{\vee}=\widetilde{G}^{\wedge}$ is in $C_{0}\left(\mathbb{R}^{n}\right)$, and $G=\widehat{g}$. Also

$$
F(\xi)=\left(1+\sum_{i=1}^{n} \xi_{i}^{2}\right)^{M} G(\xi)=P(\xi) \widehat{g}(\xi)=(P(D) g)^{\wedge}(\xi)
$$

where $P(\xi)=\left(1+|\xi|^{2}\right)^{M}$ is a polynomial, by (ii) of Proposition 1.2.5. But then:

$$
T=(\widehat{T})^{\vee}=F^{\vee}=\left((P(D) g)^{\wedge}\right)^{\vee}=P(D) g
$$

which proves the proposition.

Remark 1.4.20 (Tempered distributions given by non-negative $L_{1, l o c}$ functions). We saw with the example of $e^{x} \cos e^{x}$ in the Remark 1.3.7 that an $L_{1, l o c}$ function which is a tempered distribution is not necessarily a tempered function in the sense of Example 1.3.3. However, if $f \in L_{1, l o c}\left(\mathbb{R}^{n}\right)$, and $f$ is non-negative, then the distribution $T_{f}$ defined by $f$ is a tempered distribution implies that the function $f$ is a tempered function in the sense of Example 1.3.3. For it is enough to prove, for example, that for some $N,(1+|x|)^{-N} f$ is integrable on say $\{|x| \geq 2\}$, because every locally integrable function is integrable on $\overline{B(0,2)}$. Let $\psi \in C_{c}^{\infty}(\mathbb{R})$ be a real valued non-negative function with $\psi \equiv 1$ on the interval $[-1 / 2,1 / 2]$, and $\equiv 0$ outside the interval $[-1,1]$. For $a \geq 2$, define on $\mathbb{R}^{n}$ the radially symmetric non-negative function:

$$
\psi_{a}(x)=\psi(|x|-a)
$$

which is compactly supported in the annulus $\{a-1 \leq|x| \leq a+1\}$. Since $\psi_{a}$ are radially symmetric, it is easy to check that the Schwartz seminorms of these functions are majorised as:

$$
p_{\alpha \beta}\left(\psi_{a}\right)=\sup _{x}\left|x^{\alpha} D_{x}^{\beta} \psi_{a}\right| \leq(a+1)^{|\alpha|}\left|\partial_{r}^{\beta} \psi\right| \leq C_{\beta}(a+1)^{|\alpha|}
$$

where $C_{\beta}$ is independent of $a$. Now, since $f$ is a tempered distribution, we apply the Lemma 1.4.13 above to conclude that for $a \geq 2$ we have:

$$
\int_{a-\frac{1}{2} \leq|x| \leq a+\frac{1}{2}} f(x) d x \leq \int_{\mathbb{R}^{n}} f(x) \psi_{a}(x) d x \leq \sum_{i=1}^{k} p_{\alpha_{i}, \beta_{i}}\left(\psi_{a}\right) \leq C(a+1)^{N}
$$

for some $N$, and $C$ independent of $a$. That is, the integral of $f$ on the annulus $\left\{a-\frac{1}{2} \leq|x| \leq a+\frac{1}{2}\right\}$ is of polynomial growth in $a$. From this it is easy to check that $f$ is a tempered function.

As a consequence of the above discussion, a function $f \in L_{1, l o c}$ is a tempered function iff $|f|$ is a tempered distribution.

## 2. Distributions and Partial Differential Equations

2.1. Motivation from Electrostatics. We recall that in electromagnetism, the Maxwell equations imply that for a smooth charge distribution $g \in C^{\infty}$, the scalar electrostatic potential is given by a function $\phi$, where $\phi$ is a solution to the inhomogeneous Laplace equation

$$
\Delta \phi=-\sum_{i=1}^{3} \partial_{i}^{2} \phi=g
$$

Classically, it was known that the potential due to a unit point charge at the origin was given by $\phi(x)=C|x|^{-1}$ by the inverse square law, so the potential at $x$ due to the "infinitesimal" charge element $g(y) d y$ situated at $y$
would be $C|x-y|^{-1} g(y) d y$. Since the scalar potential is additive, the total potential at $x$ due to the entire charge distribution would be the integral:

$$
\begin{equation*}
\phi(x)=C \int|x-y|^{-1} g(y) d y \tag{3}
\end{equation*}
$$

This looks like the convolution of the "function" $C|x|^{-1}$ and $g$. Only $C|x|^{-1}$ is not a function. It is however, a tempered distribution, indeed it is a tempered function as is easily checked by using polar coordinates. So although the expression in (3) above doesn't quite make sense unless we justify the convergence of the integral above, we can try to see if it can be recast as a convolution of a the tempered distribution $C|x|^{-1}$ which will (a) rigorise and (b) generalise the above heuristic argument.

### 2.2. Fundamental solutions.

Definition 2.2.1. Let $L$ be a linear differential operator with constant coefficients on $\mathbb{R}^{n}$. That is, $L=P(D)$ where $P$ is an $n$-variable polynomial. We say that the distribution $T$ is a fundamental solution of $L$ if

$$
L T=\delta_{0}
$$

as an identity of distributions. We are not suggesting that they exist in general. If, however, $T$ is a tempered distribution, then its Fourier transform $\widehat{T}$ must also be a tempered distribution, so also $L T$. By taking the Fourier transform of $L T=\delta_{0}$ and applying (ii) of Proposition 1.4.12 and Corollary 1.4.16 we see that $\widehat{T}$ must satisfy the identity:

$$
P(\xi) \widehat{T}=1
$$

of tempered distributions. More on this later.

The reason to look for fundamental solutions is the following proposition.
Proposition 2.2.2. Let $g \in \mathcal{S}$ be a rapidly decreasing function, and assume that $T$ is a tempered distribution which is a fundamental solution of $L=P(D)$. Then the smooth function $\phi:=g * T$ is a smooth solution of $L \phi=g$. Similarly, if $g \in \mathcal{D}$ is a compactly supported function and $T$ is any distributional fundamental solution to $L$.

Proof: By the Lemma 1.4.7, we have in all the cases cited above that $\phi$ is a smooth function. Furthermore, by the same lemma, and the Example 1.4.6, we have:

$$
L(\phi)=P(D)(g * T)=g * P(D) T=g * \delta_{0}=g
$$

This proves our proposition.
This is a "soft analysis" method of solving the inhomogeneous equation $L \phi=g$, given a fundamental solution. Finding a fundamental solution, however, is not a "soft" activity. We illustrate with a few examples below.

Let us define the following linear first order differential operators on $\mathbb{R}^{2}$ :

$$
\bar{\partial}:=\frac{1}{2}\left(\partial_{x}+i \partial_{y}\right), \quad \partial:=\frac{1}{2}\left(\partial_{x}-i \partial_{y}\right)
$$

(The operator $\bar{\partial}$ is called the Cauchy-Riemann operator). Note that $4 \partial \bar{\partial}=4 \bar{\partial} \partial=-\Delta$ where $\Delta=-\partial_{x}^{2}-\partial_{y}^{2}$ is the Laplace operator on the plane.

Proposition 2.2.3 (Cauchy Problem). On $\mathbb{R}^{2}$, the tempered distribution $2(x+i y)^{-1}=2 / z$ is a fundamental solution to $\bar{\partial}$. The tempered distribution $-\log |z|$ is a fundamental solution to $\Delta$.

Proof: We recall that our volume element on $\mathbb{R}^{2}$ is $d V=(2 \pi)^{-1} d x d y$. By using polar coordinates, for example, $d V=(2 \pi)^{-1} r d r d \theta$, and it is readily verified that $2 / z$ and $\log |z|$ are tempered functions, and hence tempered distributions by Example 1.3.3.

For $f=P+i Q$ a complex valued function, the 1-form $f d z$ on $\mathbb{R}^{2}$ denotes $(P+i Q)(d x+i d y)=(P d x-$ $Q d y)+i(Q d x+P d y)$. If $W \subset \mathbb{R}^{2}$ is any open set, and $\Omega \subset W$ is a compact domain with smooth boundary $\partial \Omega$, then we have the Green Formulas:

$$
\int_{\partial \Omega} Q d x+P d y=2 \pi \int_{\Omega}\left(\partial_{x} P-\partial_{y} Q\right) d V, \quad \int_{\partial \Omega} P d x-Q d y=-2 \pi \int_{\Omega}\left(\partial_{x} Q+\partial_{y} P\right) d Q
$$

Since $\bar{\partial} f=\frac{1}{2}\left(\partial_{x}+i \partial_{y}\right)(P+i Q)=\frac{1}{2}\left(\partial_{x} P-\partial_{y} Q\right)+i\left(\partial_{y} P+\partial_{x} Q\right)$, we can write the two Green formulas above for the 1 -form $f d z$ as the single formula:

$$
\begin{equation*}
\int_{\partial \Omega} f d z=4 \pi i \int_{\Omega}(\bar{\partial} f) d V \tag{4}
\end{equation*}
$$

Now we claim that the tempered distribution $\frac{2}{z}$ is a fundamental solution of $\bar{\partial}$ on $\mathbb{R}^{2}$.
For, let $\phi \in \mathcal{S}$ be a smooth function. Then note that on $\mathbb{R}^{2} \backslash\{0\}$, the smooth function $\frac{2}{z}$ is holomorphic, so that on $\mathbb{R}^{2} \backslash\{0\}$ we have by the Leibnitz formula that $\bar{\partial}(2 \phi / z)=(2 / z) \bar{\partial} \phi$. For $\epsilon>0, R>0$, let $\Omega_{\epsilon, R} \subset \mathbb{R}^{2} \backslash\{0\}$ denote the annulus $\epsilon \leq|z| \leq R$. Choose $R \gg 0$ so that the support of $\phi$ is contained in $B(0, R)$. We apply Green's formula (4) to the function $f(z)=2 \phi / z, W=\mathbb{R}^{2}-\{0\}$ and $\Omega=\Omega_{\epsilon, R}$, to obtain:

$$
\begin{aligned}
\int_{\Omega_{\epsilon, R}}\left(\frac{2}{z}\right) \bar{\partial} \phi d V & =\int_{\Omega_{\epsilon, R}} \bar{\partial}(2 \phi / z) d V=\int_{\Omega_{\epsilon, R}} \bar{\partial} f d V \\
& =\frac{1}{4 \pi i}\left[\int_{S(R)} f(z) d z-\int_{S(\epsilon)} f(z) d z\right] \\
& =\frac{i}{2 \pi} \int_{S(\epsilon)} \frac{\phi(z)}{z} d z
\end{aligned}
$$

where $S(r)$ denotes the circle of radius $r$ centred at the origin and oriented counterclockwise, and the integral over $S(R)$ vanishes because $\phi \equiv 0$ on $S(R)$ by the choice of $R$. From the fact that $\int_{S(r)} d z / z=2 \pi i$, it follows that:

$$
\int_{\mathbb{R}^{2}} \frac{2}{z} \bar{\partial} \phi d V=\lim _{\epsilon \rightarrow 0} \int_{\Omega_{\epsilon, R}} \frac{2}{z} \bar{\partial} \phi d V=\lim _{\epsilon \rightarrow 0} \frac{i}{2 \pi} \int_{S(\epsilon)} \frac{\phi(z)}{z} d z=-\phi(0)
$$

From this it follows that:

$$
\bar{\partial}(2 / z)(\phi)=-(2 / z)(\bar{\partial} \phi)=-\int_{\mathbb{R}^{2}}\left(\frac{2}{z}\right) \bar{\partial} \phi d V=\phi(0)=\delta_{0}(\phi)
$$

for all $\phi \in \mathcal{S}$. Thus $\bar{\partial}(2 / z)=\delta_{0}$, and the assertion for the Cauchy Riemann operator follows.
The statement for the Laplacian follows by first checking that $\log |z|$ is a tempered distribution, and obeys the distributional identity:

$$
4 \partial \log |z|=\left(\partial_{x}-i \partial_{y}\right)\left(\log \left(x^{2}+y^{2}\right)\right)=\frac{2}{z}
$$

as distributions on $\mathbb{R}^{2}$. This is clear enough as an identity of functions on $\mathbb{R}^{2} \backslash\{0\}$, but has to be verified as an identity of distributions on $\mathbb{R}^{2}$, which involves using the annuli $\Omega_{\epsilon, R}$ etc., and writing down a $\partial$ analogue of the $\bar{\partial}$ Green's formula that we had in (4) above. We leave these details to the reader.

Then it follows that:

$$
\Delta(-\log |z|)=4 \bar{\partial} \partial(-\log |z|)=\bar{\partial}(4 \partial \log |z|)=\bar{\partial}(2 / z)=\delta_{0}
$$

by the fact that $(2 / z)$ is a fundamental solution to $\bar{\partial}$ proved above. The proposition follows.

Proposition 2.2.4 (Fundamental solutions to $\Delta$ on $\mathbb{R}^{n}, n \neq 2$ ). Let $n \neq 2$. A fundamental solution to $\Delta$ on $\mathbb{R}^{n}$ is given by the tempered distribution:

$$
T=\frac{(2 \pi)^{n / 2} r^{-n+2}}{(2-n) \omega_{n-1}}
$$

where $\omega_{n-1}:=\operatorname{Vol} S^{n-1}$.

Proof: There is the following special case of Stokes's Theorem (=Gauss's divergence theorem) for a domain $\Omega$ with smooth boundary $\partial \Omega$ contained in an open subset $U \subset \mathbb{R}^{n}$, and a smooth vector field $\mathbf{v}(x)=$ $\left(v_{1}(x), . ., v_{n}(x)\right)$ on $U$.

$$
\int_{\Omega}\left(\sum_{i} \partial_{i} v_{i}\right) d V=(2 \pi)^{-n / 2} \int_{\partial \Omega}(\mathbf{v} . \nu) d S
$$

where $d S$ is the induced surface measure on $\partial \Omega$ from the Euclidean measure $d x_{1} \ldots d x_{n}$ on $\mathbb{R}^{n}$, and $\nu$ denotes the outward normal vector field on $\partial \Omega$. (The factor of $(2 \pi)^{-n / 2}$ comes because for us $d V=(2 \pi)^{-n / 2} d x_{1} \ldots d x_{n}$.)

If we substitute for $\mathbf{v}$ the particular vector field $\mathbf{v}=f \nabla g$, where $f, g$ are smooth functions on $U$, we have the formula:

$$
\int_{\Omega}(\nabla f . \nabla g) d V-\int_{\Omega}(f \Delta g) d V=(2 \pi)^{-n / 2} \int_{\partial \Omega}\left(f \partial_{\nu} g\right) d S
$$

where $\partial_{\nu} g:=\nabla g . \nu$ is the normal derivative vector field of $g$ on $\partial \Omega$. (Remember that $\Delta=-\sum_{i} \partial_{i}^{2}$ ).
Interchanging the roles of $f$ and $g$, and subtracting, we have the Green Formula:

$$
\begin{equation*}
\int_{\Omega}(f \Delta g-g \Delta f) d V=(2 \pi)^{-n / 2} \int_{\partial \Omega}\left(g \partial_{\nu} f-f \partial_{\nu} g\right) d S \tag{5}
\end{equation*}
$$

Now note that the function $f(x):=\|x\|^{-n+2}=r^{-n+2}$ is tempered, and on a radially symmetric function it is easily checked by using polar coordinates that:

$$
\Delta=-\partial_{r}^{2}-(n-1) r^{-1} \partial_{r}
$$

so that $\Delta f=\Delta\left(r^{-n+2}\right) \equiv 0$ on $\mathbb{R}^{n} \backslash\{0\}$. Now let $\Omega_{\epsilon, R}=\{x: \epsilon \leq\|x\| \leq R\}$. If $g$ is a smooth function of compact support, and supp $g \subset B(0, R)$, we will have $g=\partial_{\nu} g \equiv 0$ on the sphere $S(R)$ of radius $R$. Thus from the Green formula (5) it follows that:

$$
\begin{aligned}
\int_{\Omega_{\epsilon, R}} f \Delta g & =-(2 \pi)^{-n / 2} \int_{S(\epsilon)} g \partial_{\nu} f d S=(2 \pi)^{-n / 2} \int_{S(\epsilon)} g \partial_{r} f d S=(2 \pi)^{-n / 2}(2-n) \int_{S(\epsilon)} g \cdot r^{-n+1} d S \\
& =(2 \pi)^{-n / 2}(2-n) \int_{S(\epsilon)} g \cdot \epsilon^{1-n} d S=(2 \pi)^{-n / 2}(2-n) \int_{S(1)} g(\epsilon x) d S
\end{aligned}
$$

It is clear that as $\epsilon \rightarrow 0$, the expression above converges to

$$
(2 \pi)^{-n / 2}(2-n) \omega_{n-1} g(0)
$$

where $\omega_{n-1}:=\operatorname{Vol} S(1)$ is the volume of the unit sphere in $\mathbb{R}^{n}$. Thus we have:

$$
(\Delta f)(g)=\int_{\mathbb{R}^{n}} f \Delta g d V=(2 \pi)^{-n / 2}(2-n) \omega_{n-1} \delta_{0}(g)
$$

which shows that a fundamental solution is as asserted.

Remark 2.2.5. In general there is nothing unique about a fundamental solution. For example, since the Cauchy-Riemann operator $\bar{\partial}$ annihilates every holomorphic function $f$, the distribution $2 / z+f(z)$ is also a fundamental solution for $\bar{\partial}$. Likewise for the Laplacian $\Delta$, adding on a harmonic function (i.e. a function annihilated by $\Delta$ ) will also provide a fundamental solution. For a general constant coefficient linear partial differential operator $L$, all the fundamental solutions constitute the affine space:

$$
\phi+\operatorname{ker}\left\{L: \mathcal{D}^{\prime} \rightarrow \mathcal{D}^{\prime}\right\}
$$

where $\phi$ is one fundamental solution. It is a consequence of the elliptic regularity theorem to be proved later that any distribution in the kernel of $\bar{\partial}$ or $\Delta$ is actually a function, a smooth function in fact.

## 3. Sobolev Theory

We will define certain Hilbert spaces which provide the ideal ones for studying differential operators, and more generally the "pseudodifferential operators" to be introduced later.

### 3.1. Sobolev Spaces.

Definition 3.1.1. Let $s \in \mathbb{R}$. The Sobolev space $H_{s}\left(\mathbb{R}^{n}\right)$ is defined as:

$$
H_{s}\left(\mathbb{R}^{n}\right)=\left\{f \in \mathcal{S}^{\prime}: \widehat{f} \text { is a measurable function and } \int_{\mathbb{R}^{n}}\left(1+|\xi|^{2}\right)^{s}|\widehat{f}(\xi)|^{2} d \xi<\infty\right\}
$$

For $f, g \in H_{s}\left(\mathbb{R}^{n}\right)$, their Sobolev inner-product is defined by

$$
(f, g)_{s}:=\int_{\mathbb{R}^{n}}\left(1+|\xi|^{2}\right)^{s} \widehat{f}(\xi) \overline{\widehat{g}}(\xi) d \xi
$$

which is finite by applying the Cauchy-Schwartz inequality.

## Remark 3.1.2.

(i): By the Plancherel Theorem in (iv) of 1.2 .5 , we have $H_{0}\left(\mathbb{R}^{n}\right)=L_{2}\left(\mathbb{R}^{n}\right)$.
(ii): Note that for any $s \in \mathbb{R}$, we have that the function $\rho_{s}(\xi):=\left(1+|\xi|^{2}\right)^{s / 2}$ is a slowly increasing function. Thus mlutiplication by this function is an isomorphism $\mathcal{S} \rightarrow \mathcal{S}$. Then if we define the linear operator:

$$
\begin{aligned}
\Lambda_{s}: \mathcal{S}^{\prime} & \rightarrow \mathcal{S}^{\prime} \\
f & \mapsto\left(\rho_{s} \widehat{f}\right)^{\vee}
\end{aligned}
$$

it follows that this operator is a continuous isomorphism (with inverse $\Lambda_{-s}$ ). Hence, in view of the Plancherel Theorem (iv) of 1.2 .5 , we have the description:

$$
H_{s}=\left\{f \in \mathcal{S}^{\prime}: \Lambda_{s} f \in L_{2}\left(\mathbb{R}^{n}\right)\right\}
$$

and since $\Lambda_{s}$ is an isomorphism, it follows that $H_{s}$ is isomorphic to $L_{2}$ as a Hilbert space. In particular it is a separable Hilbert space.
(iii): For each $t \leq s$, we have $\left(1+|\xi|^{2}\right)^{t} \leq\left(1+|\xi|^{2}\right)^{s}$, so $H_{s} \subset H_{t}$ for all $s \geq t$.
(iv): If $T \in \mathcal{E}^{\prime}$ is a compactly supported distribution, then by the Proposition 1.4.15, it follows that $\widehat{T}$ is a function (namely $T\left(e_{-\xi}\right)$ ) which is of slow growth. That means its modulus square is also a function of slow growth, and it will be integrable against $\left(1+|\xi|^{2}\right)^{s}$ for some $s$. Hence $T$ will be in the corresponding $H_{s}\left(\mathbb{R}^{n}\right)$. On the other hand every non-compactly supported Schwartz class function in in each $H_{s}$, but not in $\mathcal{E}^{\prime}$.
(v): Not every tempered distribution is in some $H_{s}$. For, the constant function 1 is a tempered distribution, but since its Fourier transform is $\delta_{0}$, which is not a function, 1 does not belong to any $H_{s}$. Thus in view of (iv) above, we have strict containments:

$$
\mathcal{E}^{\prime} \subset H_{-\infty}:=\cup_{s \in \mathbb{R}} H_{s} \subset \mathcal{S}^{\prime}
$$

(vi): Since $\Lambda_{s}$ is an isomorphism on $\mathcal{S}$, and so $\Lambda_{s} f \in L_{2}$ for every $f \in \mathcal{S}$ and every $s$, it follows that:

$$
\mathcal{S} \subset H_{\infty}:=\cap_{s \in \mathbb{R}} H_{s}
$$

Also since $\Lambda_{s}$ is a Hilbert space isometry of $H_{s}$ to $H_{0}=L_{2}, \Lambda_{s}(\mathcal{S})=\mathcal{S}$, and $\mathcal{S}$ is dense in $L_{2}$, it follows that $\mathcal{S}$ is dense in each $H_{s}$.
(vii): However, the containment:

$$
\mathcal{S} \subset H_{\infty}:=\cap_{s} H_{s}
$$

is also strict. For example, the function $f(x)=\left(1+x^{2}\right)^{-1}$ on $\mathbb{R}$ has the Fourier transform $e^{-|\xi|}$, which is integrable against all powers of $\left(1+|\xi|^{2}\right)^{s}$, so $f \in H_{\infty}$. However, $f \notin \mathcal{S}$, because its Fourier transform $e^{-|\xi|}$ is not smooth, so not in $\mathcal{S}$.

Exercise 3.1.3. Prove the slightly stronger statement than (vi) above, viz. that $\mathcal{D}=C_{c}^{\infty}$ is dense in each $H_{s}$. Thus in view of (v) and this statement, one could also define $H_{s}$ as the completion of the inner product spaces $\mathcal{D}$ or $\mathcal{S}$ with respect to the Sobolev inner product $(-,-)_{s}$.

We now have a very elementary proposition about these Sobolev Spaces.

Proposition 3.1.4 (Some Facts on Sobolev Spaces).
(i): The inclusion $H_{s} \hookrightarrow H_{t}$ for $s \geq t$, defined in (iii) of 3.1.2 is a continuous (=bounded) operator. If $f \in \mathcal{S}$, the multiplication operator $u \mapsto f u$ is a continuous (=bounded) operator.
(ii): If $m \geq 0$ is a non-negative integer, then on the vector subspace $\mathcal{E} \cap H_{m}\left(\mathbb{R}^{n}\right)$ of $H_{m}$, the Sobolev $m$-norm is equivalent to the norm defined by:

$$
\|f\|^{2}=\sum_{|\alpha| \leq m} \int_{\mathbb{R}^{n}}\left|D_{x}^{\alpha} f\right|^{2} d x
$$

Thus for such an $m, H_{m}$ can be described as the completion of $\mathcal{S}$ or $\mathcal{D}$ with respect to this norm.
(iii): If $P$ is a polynomial of degree $k$, then for the linear constant coefficient differential operator $P(D)$, we have:

$$
P(D): H_{s} \rightarrow H_{s-k}
$$

is a continuous(=bounded) operator of Hilbert spaces.
(iv): The sesquilinear pairing:

$$
\begin{aligned}
\mathcal{S} \times \mathcal{S} & \rightarrow \mathbb{C} \\
f, g & \mapsto \int_{\mathbb{R}^{n}} f(x) \bar{g}(x) d x=\langle f, g\rangle
\end{aligned}
$$

extends to a sesquilinear pairing of $H_{s} \times H_{-s}$, also denoted $\langle-,-\rangle$ which satisfies:

$$
|\langle f, g\rangle| \leq\|f\|_{s}\|g\|_{-s} ; \quad\|f\|_{s}=\sup _{0 \neq g \in H_{s}} \frac{|\langle f, g\rangle|}{\|g\|_{-s}}
$$

$\langle-,-\rangle$ is therefore a perfect pairing and identifies $H_{-s}$ with the Hilbert space dual $\left(H_{s}\right)^{*}$ of $H_{s}$.

## Proof:

(i) is trivial from the fact for $t \leq s$ we have $\left(1+|\xi|^{2}\right)^{t} \leq\left(1+|\xi|^{2}\right)^{s}$ and all $\xi$. The second statement is also straightforward, and left as an exercise.

For (ii), we note that there exists a constant $C$ such that:

$$
\frac{1}{C}\left(1+|\xi|^{2}\right)^{m} \leq \sum_{\alpha \leq m}\left|\xi^{2 \alpha}\right| \leq C\left(1+|\xi|^{2}\right)^{m}, \quad \xi \in \mathbb{R}^{n}
$$

from which it follows by (ii) of the Proposition 1.2.5 that:

$$
\frac{1}{C}\left(1+|\xi|^{2}\right)^{m}|\widehat{f}(\xi)|^{2} \leq \sum_{\alpha \leq m}\left|\left(D^{\alpha} f\right)^{\wedge}(\xi)\right|^{2} \leq C\left(1+|\xi|^{2}\right)^{m}|\widehat{f}(\xi)|^{2}
$$

and all $\xi \in \mathbb{R}^{m}$. The result follows by integrating the above two inequalities over $\mathbb{R}^{n}$, and the Plancherel Theorem (iv), 1.2.5.
(iii) is also clear from the fact that $(P(D) f)^{\wedge}(\xi)=P(\xi) \widehat{f}(\xi)$, and that $|P(\xi)|^{2} \leq C\left(1+|\xi|^{2}\right)^{k}$ for some $C>0$ and all $\xi \in \mathbb{R}^{n}$, if $k=\operatorname{deg} P$.

To see (iv), note that for $f, g \in \mathcal{S}$, we have by Plancherel:

$$
\langle f, g\rangle=\langle\widehat{f}, \widehat{g}\rangle=\int \widehat{f}(\xi)\left(1+|\xi|^{2}\right)^{s / 2} \widehat{\widetilde{g}}(\xi)\left(1+|\xi|^{2}\right)^{-s / 2} d \xi \leq\|f\|_{s}\|\widetilde{g}\|_{-s}=\|f\|_{s}\|g\|_{-s}
$$

by using the Cauchy-Schwartz inequality. To see that equality is achieved in the inequality, choose $g$ such that $\widehat{\widetilde{g}}=\widehat{f}\left(1+|\xi|^{2}\right)^{s}$. This yields the rest of (iv), and the proposition follows.

Remark 3.1.5. (iii) of the Proposition 3.1.4 above is the reason for introducing Sobolev spaces, i.e. in order to view differential operators as being bounded operators between Hilbert spaces.
3.2. Sobolev Embedding Theorem. There is a criterion for a function to be a $k$ times continuously differentiable function which can be stated in terms of Sobolev spaces.

Proposition 3.2.1 (Sobolev Embedding Theorem or Sobolev Lemma). Let $k \geq 0$ be an integer. If $s>k+\frac{n}{2}$, then:
(i): $H_{s}\left(\mathbb{R}^{n}\right) \subset C_{0}^{k}\left(\mathbb{R}^{n}\right)$, where the right hand space denotes the space of $k$ times continuously differentiable functions $f$ with $D_{x}^{\alpha} f$ vanishing at $\infty$ for all $|\alpha| \leq k$.
(ii): $\left\|D_{x}^{\alpha} f\right\|_{\infty} \leq C_{\alpha}\|f\|_{s}$. Indeed if we define the norm

$$
\|f\|_{\infty, k}=\sup _{|\alpha| \leq k}\left\|D_{x}^{\alpha} f\right\|_{\infty}
$$

on $C_{k}^{\infty}\left(\mathbb{R}^{n}\right)$, then the inclusion $H_{s} \subset C_{0}^{k}$ of (i) above is continuous.

Proof: We first make the following:
Claim: If $f \in \mathcal{S}^{\prime}$ is a tempered distribution such that $\left(D_{x}^{\alpha} f\right)^{\wedge}$ is a function in $L_{1}\left(\mathbb{R}^{n}\right)$, then $f \in C_{0}^{k}\left(\mathbb{R}^{n}\right)$. Also, $\left\|D_{x}^{\alpha} f\right\|_{\infty}=\sup _{x \in \mathbb{R}^{n}}\left|D_{x}^{\alpha} f\right| \leq\left\|\left(D_{x}^{\alpha} f\right)^{\wedge}\right\|_{1}$.

If $g:=\left(D_{x}^{\alpha} f\right)^{\wedge} \in L^{1}$, then by the Riemann-Lebesgue Lemma (v) of Proposition 1.2.5, we have $D_{x}^{\alpha} f=g^{\vee}$ is in $C_{0}\left(\mathbb{R}^{n}\right)$. The last statement is clear from the fact that $\left\|g^{\vee}\right\|_{\infty} \leq\|g\|_{1}$.

In view of the above claim, all we have to do is show that if $f \in H_{s}\left(\mathbb{R}^{n}\right)$ for $s \geq k+\frac{n}{2}$, then $\left(D_{x}^{\alpha} f\right)^{\wedge}$ is an $L_{1}$ function. But then:

$$
\begin{aligned}
\int\left|\left(D_{x}^{\alpha} f\right)^{\wedge}(\xi)\right| d \xi & =\int|\xi|^{\alpha}|\widehat{f}(\xi)| \xi=\int|\xi|^{\alpha}\left(1+|\xi|^{2}\right)^{-s / 2}\left(1+|\xi|^{2}\right)^{s / 2}|\widehat{f}(\xi)| d \xi \\
& \leq\left(\int|\xi|^{2 \alpha}\left(1+|\xi|^{2}\right)^{-s} d \xi\right)^{1 / 2}\|f\|_{s}
\end{aligned}
$$

by the Cauchy-Schwartz inequality. Since:

$$
|\xi|^{2|\alpha|}\left(1+|\xi|^{2}\right)^{-s} \leq\left(1+|\xi|^{2}\right)^{k-s}
$$

and $s-k>n / 2$, the integral $\int|\xi|^{2 \alpha}\left(1+|\xi|^{2}\right)^{-s} d \xi$ is finite, and we therefore have:

$$
\left\|\left(D_{x}^{\alpha} f\right)^{\wedge}\right\|_{1} \leq C\|f\|_{s}
$$

which implies by the Claim above that:

$$
\left\|D_{x}^{\alpha} f\right\|_{\infty} \leq C\|f\|_{s} \quad \text { for all } \quad|\alpha| \leq k
$$

This proves both (i) and (ii) and the proposition follows.

Corollary 3.2.2. As a consequence of the entire subsection, we have the following chain of inclusions:

$$
\mathcal{S} \subset H_{\infty} \subset H_{-\infty} \subset \mathcal{S}^{\prime}
$$

each of which is strict. Note also that by the Sobolev Lemma above,

$$
H_{\infty} \subset C_{0}^{\infty}
$$

Note also that the Dirac distribution $\delta_{0}$ belongs to $H_{-s}$ for all $s>\frac{n}{2}$ since $\widehat{\delta_{0}}=1$, and

$$
\left\|\delta_{0}\right\|_{-s}=\int\left(1+|\xi|^{2}\right)^{-s} d \xi<\infty
$$

for all $s>\frac{n}{2}$. In general, the more negative the $s$, the more singular the tempered distributions that will be included in $H_{s}$.

Remark 3.2.3. The Sobolev Lemma above is crucial for proving regularity (smoothness) of distributional solutions to elliptic differential operators.
Exercise 3.2.4. By the Sobolev Lemma, $H_{\infty} \subset C_{0}^{\infty}$. Is $C_{0}^{\infty}$ a subset of $H_{-\infty}$ ?
3.3. Rellich's Lemma. The other crucial lemma about the Sobolev spaces is a statement about the inclusion $H_{s} \subset H_{t}$ for $s>t$. Before we prove it we state the following lemma about locally compact metric spaces.
Proposition 3.3.1 (Arzela-Ascoli Theorem). Let $X$ be a locally compact $\sigma$-compact metric space ( $\sigma$-compactness means $X$ is a countable union of compact subsets). Let $\left\{f_{k}\right\}$ be a sequence of complex valued functions satisfying:
(i): $\left\{f_{k}\right\}$ is equicontinuous. That is for each $x \in X$ and each $\epsilon>0$, there is a neighbourhood $U_{x}$ of $x$ such that $\left|f_{k}(x)-f_{k}(y)\right|<\epsilon$ for all $y \in U_{x}$ and all $k$.
(ii): $\left\{f_{k}\right\}$ is pointwise bounded, i.e. the set $\left\{f_{k}(x): k \in \mathbb{N}\right\}$ is a bounded set for each $x \in X$.

Then there exists a function $f \in C(X)$ and a subsequence $\left\{f_{k_{m}}\right\}$ of $\left\{f_{k}\right\}$ such that $f_{k_{m}} \rightarrow f$ uniformly on compact sets.

Proof: See Rudin's Real and Complex Analysis, or Folland's Real Analysis: Modern Techniques and their Applications.

Proposition 3.3.2 (Rellich's Lemma). Let $s>t$, so that $H_{s} \subset H_{t}$. Let $\left\{f_{k}\right\}$ be a sequence in $H_{s}$ such that:
(i): There exists a compact set $K$ such that for all $k$, the support of (the tempered distribution) $f_{k}$ is contained in $K$.
(ii): $\left\{f_{k}\right\}_{k \in \mathbb{N}}$ is a bounded set in $H_{s}$.

Then there is a subsequence of $\left\{f_{k}\right\}$ which converges in $H_{t}$.

Proof: First note that for $\xi, \eta \in \mathbb{R}^{n}$, we have by the triangle inequality that:

$$
|\xi|^{2} \leq 2\left(|\xi-\eta|^{2}+|\eta|^{2}\right)
$$

which implies that:

$$
\left(1+|\xi|^{2}\right) \leq 2\left(1+|\xi-\eta|^{2}\right)\left(1+|\eta|^{2}\right)
$$

Thus if $s \geq 0$, we have:

$$
\left(1+|\xi|^{2}\right)^{s / 2} \leq C\left(1+|\xi-\eta|^{2}\right)^{s / 2}\left(1+|\eta|^{2}\right)^{s / 2}
$$

where $C$ is a constant depending on $s$. Similarly, if $s<0$, we can apply the above inequality to $|s|=-s$ and interchange the roles of $\xi$ and $\eta$ to obtain the so called Peetre inequality for all $s$

$$
\left(1+|\xi|^{2}\right)^{s / 2} \leq C\left(1+|\xi-\eta|^{2}\right)^{|s| / 2}\left(1+|\eta|^{2}\right)^{s / 2}
$$

Since $f_{k} \in H_{s}$, by definition $\widehat{f}_{k}$ is a function for each $k$. Let $\phi \in \mathcal{D}$ be a smooth compactly supported function which is $\equiv 1$ on $K$. Then $f_{k}=\phi f_{k}$ as distributions, and by (iv) of the Proposition 1.4.12 we have $\widehat{f_{k}}=\widehat{\phi} * \widehat{f_{k}}$. Thus:

$$
\left|\widehat{f}_{k}(\xi)\right|=\left|\left(\widehat{\phi} * \widehat{f}_{k}\right)(\xi)\right|=\left|\int \widehat{\phi}(\xi-\eta) \widehat{f}_{k}(\eta) d \eta\right| \leq \int\left|\widehat{\phi}(\xi-\eta) \widehat{f}_{k}(\eta)\right| d \eta
$$

which together with Peetre's inequality above implies that:

$$
\left(1+|\xi|^{2}\right)^{s / 2}\left|\widehat{f}_{k}(\xi)\right| \leq C \int\left|\widehat{\phi}(\xi-\eta)\left(1+|\xi-\eta|^{2}\right)\right|^{|s| / 2}\left|\left(1+|\eta|^{2}\right)^{s / 2} \widehat{f}_{k}(\eta)\right| d \eta
$$

for all $s$. Applying the Cauchy-Schwartz inequality to the integral on the right, we have:

$$
\begin{equation*}
\left(1+|\xi|^{2}\right)^{s / 2}\left|\widehat{f}_{k}(\xi)\right| \leq C\|\phi\|_{|s|}\left\|f_{k}\right\|_{s} \leq C^{\prime} \quad \text { for all } k \tag{6}
\end{equation*}
$$

since $\left\{f_{k}\right\}$ is a bounded sequence in $H_{s}$.
We note that since $\widehat{f}_{k}$ are compactly supported distributions, by the Proposition 1.4.15 they are smooth functions. So, similarly, we have $d_{j} \widehat{f_{k}}=d_{j}\left(\widehat{\phi} * \widehat{f_{k}}\right)=d_{j} \widehat{\phi} * \widehat{f}_{k}$, and again a corresponding argument shows that:

$$
\left(1+|\xi|^{2}\right)^{s / 2}\left|d_{j} \widehat{f}_{k}(\xi)\right| \leq C^{\prime \prime} \text { for all } k
$$

This shows that $\widehat{f_{k}}$ and $d_{j} \widehat{f}_{k}$ are both uniformly bounded sequences of functions on each compact $L \subset \mathbb{R}^{n}$. In particular, the sequence $\widehat{f_{k}}$ is pointwise bounded, and the condition (ii) of the Arzela-Ascoli Theorem 3.3.1 is satisfied.

The uniform boundedness of all $d_{j} \widehat{f_{k}}$ implies by the Mean Value Theorem that on each compact $L \subset \mathbb{R}^{n}$, we have a uniform Lipschitz constant $C$ satisfying:

$$
\left|\widehat{f_{k}}(x)-\widehat{f_{k}}(y)\right| \leq C\|x-y\|
$$

for all $x, y \in L$ and all $k$. This shows that the sequence of functions $\left\{\widehat{f}_{k}\right\}$ is equicontinuous, and condition (i) of Arzela-Ascoli is satisfied. Thus there is a subsequence of $\left\{\widehat{f}_{k}\right\}$ which converges uniformly on compact subsets of $\mathbb{R}^{n}$. For notational convenience, denote this subsequence by $\left\{\widehat{f}_{k}\right\}$ as well.

Thus, for $t<s$, we have:

$$
\begin{align*}
\left\|f_{j}-f_{k}\right\|_{t}^{2} & =\int\left|\widehat{f}_{j}(\xi)-\widehat{f}_{k}(\xi)\right|^{2}\left(1+|\xi|^{2}\right)^{t} d \xi \\
& =\int_{|\xi| \geq r}\left|\widehat{f}_{j}(\xi)-\widehat{f}_{k}(\xi)\right|^{2}\left(1+|\xi|^{2}\right)^{t} d \xi+\int_{|\xi| \leq r}\left|\widehat{f}_{j}(\xi)-\widehat{f}_{k}(\xi)\right|^{2}\left(1+|\xi|^{2}\right)^{t} d \xi \tag{7}
\end{align*}
$$

for all $r>0$.
Since $t-s<0$, we have:

$$
\left(1+|\xi|^{2}\right)^{t}=\left(1+|\xi|^{2}\right)^{t-s}\left(1+|\xi|^{2}\right)^{s} \leq\left(1+r^{2}\right)^{t-s}\left(1+|\xi|^{2}\right)^{s} \text { for } \xi \geq r
$$

Thus, by the equation (6) (applied to $\widehat{f}_{j}-\widehat{f}_{k}$ replacing $\widehat{f}_{k}$ ), we get that the first integral in (7) is majorised by $C\left(1+r^{2}\right)^{t-s}$ for some $C>0$.

Given $\epsilon>0$, choose $r$ large enough that $C\left(1+r^{2}\right)^{t-s}<\epsilon$ then the first integral is $<\epsilon$. The second integral is $<\epsilon$ by choosing $k$ and $j$ large enough, since $\widehat{f}_{k}$ converges uniformly on the compact set $\{\xi \leq r\}$. This shows that $\left\{f_{k}\right\}$ is a Cauchy sequence in $H_{t}$, which is complete, so it converges. The proposition follows.

Exercise 3.3.3. Again show by considering a sequence of translates of a fixed function of compact support (whose supports thus march off to infinity) that the condition (i) of Rellich's Lemma cannot be dropped.

## 4. Globalisation to Compact Manifolds

In the sequel, $M$ will denote a paracompact, 2 nd countable, Hausdorff, oriented, $C^{\infty}$-manifold of dimension $n$. It is well known (by using partitions of unity) that such a manifold has a Riemannian metric on its tangent bundle $T M$, and by duality, on its cotangent bundle $T^{*} M$. We will very soon specialise to $M$ compact.

### 4.1. Smooth vector bundles and sections.

Definition 4.1.1 (Smooth vector bundles). A smooth manifold pair $\pi: E \rightarrow M$, with $\pi$ a smooth surjective submersion is called a smooth (or $C^{\infty}$ ) real (resp. complex) vector bundle of rank $k$ if:
(i): For each $x \in M$, the fibre $E_{x}:=\pi^{-1}(x)$ is a real (resp. complex) vector space of real (resp. complex) dimension $k$.
(ii): There exists an open covering $\left\{U_{i}\right\}_{i=1}^{\infty}$ of $M$ and smooth diffeomorphisms $\phi_{i}: \pi^{-1}\left(U_{i}\right) \rightarrow U_{i} \times \mathbb{F}^{k}$ (where $\mathbb{F}=\mathbb{R}($ resp. $\mathbb{C}))$ making the following diagram commute:

$$
\begin{array}{rll}
\pi^{-1}\left(U_{i}\right) & \xrightarrow{\phi_{i}} & U_{i} \times \mathbb{F}^{k} \\
\pi \searrow & & \swarrow \operatorname{pr}_{1} \\
& U_{i} &
\end{array}
$$

where $\mathrm{pr}_{1}$ denotes projection into the first factor.
(iii): For each $x \in U_{i}$, and all $i$, the composite map:

$$
E_{x}=\pi^{-1}(x) \xrightarrow{\phi_{i}}\{x\} \times \mathbb{F}^{k} \rightarrow \mathbb{F}^{k}
$$

is a linear isomorphism of real (esp. complex) vector spaces.
The smooth diffeos $\phi_{i}$ are called local charts or local trivialisations for the bundle, $E$ is called the total space and $M$ the base space of the bundle. The conditions (ii) and (iii) above simply say that the restricted bundles $E_{\mid U_{i}}: \pi^{-1}\left(U_{i}\right) \rightarrow U_{i}$ are trivial (i.e. product) bundles. When no confusion is likely, one simply writes $E$ to denote the bundle, instead of $\pi: E \rightarrow M$.

A smooth map $s: M \rightarrow E$ is called a smooth section of $E$ if $\pi \circ s=\operatorname{id}_{M}$. Using local trivialisations, it is easy to see that sections of the restricted bundles $E_{\mid U_{i}} \rightarrow U_{i}$ are in bijective correspondence with $\mathbb{F}^{k}$-valued smooth functions on $U_{i}$.

Example 4.1.2 (Some important bundles). Important examples of natural vector bundles on a smooth real (resp. complex) $n$-dimensional manifold $M$ are its real (resp. holomorphic) tangent bundle $T M$ (resp. $T_{\text {hol }} M$ ) and cotangent bundle $T^{*} M$ (resp. $T_{\text {hol }}^{*} M$. The local trivialsiations of these bundles arise naturally from a smooth atlas (resp. holomorphic atlas). We will usually be taking a real manifold of dimension $n$ and complexifying its real tangent and cotangent bundles, which will then become complex vector bundles of rank $n$ denoted respectively by $T_{\mathbb{C}} M$ and $T_{\mathbb{C}}^{*} M$. When $M$ happens to a complex manifold of complex dimension $n$, it can be viewed as a real manifold of dimension $2 n$, and $T_{\mathbb{C}} M=T_{h o l} M \oplus \overline{T_{h o l} M}$ and $T_{\mathbb{C}}^{*} M=T^{1,0} M \oplus T^{0,1} M$, where $T^{1,0}$ is the complex dual of $T_{h o l} M$ and $T^{0,1}$ the complex dual of $\overline{T_{h o l} M}$ (the conjugate bundle to $T_{\text {hol }} M$ ).

When one takes tensor or exterior powers of these bundles, one obtains other smooth bundles: $\otimes^{k} T_{\mathbb{C}} M$, the bundle of contravariant $k$-tensors, or $\otimes^{k} T_{\mathbb{C}}^{*} M$ the bundle of covariant $k$-tensors, or $\wedge^{p} T_{\mathbb{C}}^{*} M$, the bundle of complex valued differential $k$-forms. These associated bundles have natural trivialisations arising from the trivialisations of the tangent and cotangent bundles. For further details the reader may consult any standard differential topology or differential geometry text.

By proceeding componentwise, one easily defines the function spaces of $\mathbb{C}^{k}$-valued smooth functions:

$$
\mathcal{E}^{k}\left(\mathbb{R}^{n}\right)=\oplus_{i=1}^{k} \mathcal{E}\left(\mathbb{R}^{n}\right)
$$

and likewise $\mathcal{D}^{k}\left(\mathbb{R}^{n}\right)$, or $\mathcal{S}^{k}(\mathbb{R})^{n}$. So also the spaces of vector valued distributions $\mathcal{D}^{\prime k}\left(\mathbb{R}^{n}\right)$, tempered distributions $\mathcal{S}^{\prime k}\left(\mathbb{R}^{n}\right)$ and compactly supported distributions $\mathcal{E}^{\prime k}\left(\mathbb{R}^{n}\right)$.

Now let $E \rightarrow M$ be a smooth complex vector bundle on a paracompact real manifold $M$ of dimension $n$. We can choose, by refining if necessary, a covering $\mathcal{U}$ of $M$ by open sets $\left\{U_{i}\right\}_{i=1}^{\infty}$ such that:
(i): $U_{i}$ is diffeomorphic to $\mathbb{R}^{n}$, and $\bar{U}_{i}$ is compact, for each $i$.
(ii): $E_{\mid U_{i}}$ is a trivial bundle for each $i$.

Choose a partition of unity $\lambda_{i}$ subordinate to the open cover $\mathcal{U}$, so that $\operatorname{supp} \lambda_{i}$ is a compact subset of $U_{i}$ for each $i$, which is possible since $M$ is paracompact. Then, if we denote the space of smooth sections of $E$ by $C^{\infty}(M, E)$, in view of (i) and (ii) above we have a natural inclusion:

$$
\begin{aligned}
C^{\infty}(M, E) & \hookrightarrow \prod_{i=1}^{\infty} \mathcal{E}^{k}\left(\mathbb{R}^{n}\right) \\
s & \mapsto\left(\lambda_{i}\left(\phi_{i} \circ s\right)\right)_{i=1}^{\infty}
\end{aligned}
$$

Note that at each $x \in M$, only finitely many entries on the right have a non-zero value. Indeed, each compact subset $K \subset M$ meets at most finitely many $U_{i}^{\prime} s$, so that $K \cap\left(\operatorname{supp} \lambda_{i}\right)=\phi$ for all but finitely many $i$. If we denote $s_{i}:=\lambda_{i}\left(\phi_{i} \circ s\right)$, we may define seminorms:

$$
p_{\alpha, K}^{E}(s)=\sup _{i}\left(\sup _{K \cap\left(\operatorname{supp} \lambda_{i}\right)}\left|D_{x}^{\alpha}\left(s_{i}\right)\right|\right)
$$

where the quantiy in brackets on the right is the usual seminorm introduced earlier for $\mathcal{E}^{k}\left(\mathbb{R}^{n}\right)$. This defines a topology on $C^{\infty}(M, E)$.

Exercise 4.1.3. Verify that taking $M=\mathbb{R}^{n}, E$ a trivial real rank $k$ vector bundle (so that $\left.C^{\infty}(M, E)=\mathcal{E}^{k}\right)$ and a locally finite covering $\mathcal{U}$ by open balls $\left\{U_{i}\right\}_{i=1}^{\infty}$ (which are diffeomorphic to $\mathbb{R}^{n}$ ), and with $\lambda_{i}$ being a partition of unity subordinate to $\mathcal{U}$, the topology that is defined as above on $C^{\infty}(M, E)$ is the same as the topology introduced earlier on $\mathcal{E}^{k}$. (One needs to fix bounds on derivatives of $\lambda_{i}$ on their compact supports etc.)

Similarly for $C_{c}^{\infty}(M, E)$, the space of compactly supported smooth sections of $E$, we have the restriction of the above inclusion:

$$
\begin{aligned}
C_{c}^{\infty}(M, E) & \hookrightarrow \oplus_{i=1}^{\infty} \mathcal{E}^{k}\left(\mathbb{R}^{n}\right) \\
s & \mapsto \sum_{i} s_{i}
\end{aligned}
$$

where the $s_{i}$ are as above. We leave it as an exercise for the reader to define the topology on this space in a manner that is consistent (in the sense of the exercise above). We just remark that if $\left\{s_{n}\right\}$ is a sequence of smooth compactly supported sections all having support in some fixed compact set $K \subset M$, then $s_{n, i}$ above will be identically zero for all $i$ such that $i \notin F$, where $F=\left\{i: U_{i} \cap K \neq \phi\right\}$ is a finite set independent of $n$, and for each $i \in F$, all the $s_{n, i}$ will have support inside the compact set $\operatorname{supp} \lambda_{i} \cap K$.

Definition 4.1.4 (Distributions on manifolds). A continuous linear functional on $C_{c}^{\infty}(M, E)$ is called an $E$ valued distribution on $M$, and the space of these is denoted as $\mathcal{D}^{\prime}(M, E)$. Similarly, a continuous linear functional on $C^{\infty}(M, E)$ is called a compactly supported $E$-valued distribution on $M$, and their space denoted $\mathcal{E}^{\prime}(M, E)$. When $E$ is the trivial rank 1 (line) bundle on $M$, we just write $\mathcal{D}^{\prime}(M)$ (resp. $\mathcal{E}^{\prime}(M)$ ) for the respective spaces of distributions.

When $M$ is compact, $C_{c}^{\infty}(M, E)=C^{\infty}(M, E)$, and compactly supported $E$-valued distributions are exactly the same as $E$-valued distributions. One doesn't really need the space of tempered distributions on a manifold, their main use on $\mathbb{R}^{n}$ being the availability of Fourier transform, an operation that doesn't make global sense on a general manifold $M$.

Example 4.1.5 (Currents on a smooth manifold). In the particular case when $E=\wedge^{n-p} T_{\mathbb{C}}^{*} M$, the space of its smooth sections $C^{\infty}(M, E)$ is denoted $\bigwedge^{p}(M, \mathbb{C})$, and such a section is called a differential $(n-p)$-form. $E$-valued distributions on $M$ are known as p-currents on $M$. Likewise, compactly supported p-currents are elements of $\mathcal{E}^{\prime}\left(M, \wedge^{n-p} T_{\mathbb{C}}^{*} M\right)$. The reason for the indexing is that one may think of a differential $p$-form $\omega$ as a contnuous linear functional acting on the space $\bigwedge_{c}^{n-p}(M)$ via integration:

$$
T_{\omega}(\tau):=\int_{M} \omega \wedge \tau \quad \tau \in \wedge_{c}^{p}(M)
$$

where integration of an $n$-form on a singular $n$-cube is defined for the oriented manifold $M$ as usual, and where the support of $\tau$ can be covered by a finite union of $k$-cubes with the right orientations (i.e. a $k$-chain) etc. Clearly then, a differential $p$-form is a $p$-current by this indexing convention. Using the Stokes formula for a singular $k$-chain:

$$
\int_{\sigma} d \omega=\int_{\partial \sigma} \omega
$$

and the facts that (i) $d(\omega \wedge \tau)=d \omega \wedge \tau+(-1)^{\operatorname{deg}} \omega \omega \wedge d \tau$, and (ii) $\tau \in \bigwedge_{c}^{n-p}(M)$ implies that $\tau \equiv 0$ on the boundary of a sufficiently large $k$-chain covering the support of $\tau$, the reader can easily check that by defining the distributional exterior derivative of a $p$-current $T$ by $d T(\omega)=(-1)^{p+1} T(d \omega)$ for $\omega \in \wedge_{c}^{n-p}(M)$ leads to the consistency formula: $d T_{\omega}=T_{d \omega}$.

Indeed, if we denote the space of $p$-currents by $\mathcal{C}^{p}(M, \mathbb{C})$, there is the de-Rham complex of currents:

$$
\ldots \rightarrow \mathcal{C}^{p}(M, \mathbb{C}) \xrightarrow{d} \mathcal{C}^{p+1}(M, \mathbb{C}) \rightarrow \ldots
$$

with $d \circ d \equiv 0$, and the usual de-Rham complex is a subcomplex of this complex via the chain map $\omega \mapsto T_{\omega}$. It is a fact (using an approximation theorem analogous to the Proposition 1.4 .10 proved for $\mathbb{R}^{n}$ ) that this chain map is a chain homotopy equivalence.

Similarly, the singular $(n-p)$-chain $\sigma$ may be regarded as a compactly supported $p$-current via integration:

$$
T_{\sigma}(\tau):=\int_{\sigma} \tau \text { for } \tau \in \wedge^{n-p}(M, \mathbb{C})
$$

By Stokes's theorem, the distributional derivative $\partial T_{\sigma}$ defined by $\partial T_{\sigma}(\tau)=T_{\sigma}(d \tau)$ leads to the usual boundary operator on singular $(n-p)$-chains. In particular, an orientable $(n-p)$-dimensional submanifold $N$ of $M$ is an $(n-p)$ chain in $M$, and defines a $p$-current.

Analogously an infinite (Borel-Moore) locally finite ( $n-p$ )-chain maybe regarded as a $p$-current, acting on $\wedge_{c}^{n-p}(M, \mathbb{C})$ via the same integration formula as above. Again, the distributional derivative defined as above leads via Stokes to the usual geometric boundary. Thus $p$-currents (resp. compactly supported $p$-currents) are general enough to include both $(n-p)$-Borel-Moore chains (resp. singular $p$-chains) and differential $p$ forms (resp. compactly supported $p$-forms). One then shows that the cohomology of the complex of $p$ currents $\mathcal{C}^{*}(M, \mathbb{C})$ is the same as that of the Borel-Moore chain complex $\Delta_{n-*}^{B M}$, as well as the de Rham complex $\bigwedge^{*}(M, \mathbb{C})$. Similarly for compactly supported currents. Thus follow the standard Poincare duality isomorphisms of the Borel-Moore homology $H_{n-p}^{B M}(M, \mathbb{C})$ and the de-Rham cohomology $H_{d R}^{p}(M, \mathbb{C})$ ) (resp. singular homology $H_{n-p}(M, \mathbb{C})$ and compactly supported de Rham cohomology $\left.H_{d R, c}^{p}(M, \mathbb{C})\right)$

Remark 4.1.6. In all of the above, one has chosen a particular partition of unity, and a particular kind of open covering. One needs to check that everything defined above for $M$ is independent of these choices. One can actually define $\mathcal{E}(U)$ and $\mathcal{D}(U)$ for any open subset $U \subset \mathbb{R}^{n}$. Then one shows that if $U$ is a further locally finite union of $U_{i}$, an analogue of the exercise 4.1 .3 will imply that the "patching definition" of $\mathcal{D}^{\prime}(U)$ or $\mathcal{E}^{\prime}(U)$ is the same as the a priori definition. Then one uses common refinements, the partition of unity $\lambda_{i} \mu_{j}$ arising from different partitions of unity $\lambda_{i}$ and $\mu_{j}$, etc. to prove that these various choices are immaterial.
4.2. Sobolev spaces on a compact manifold. In this section $M$ is assumed to be compact throughout

Definition 4.2.1. Let $M$ be a compact manifold, and $E$ a smooth rank $k$ complex vector bundle on $M$. Again find a finite open covering $\left\{U_{i}\right\}_{i=1}^{N}$ satisfying:
(i): $U_{i}$ is diffeomorphic to $\mathbb{R}^{n}$ for each $i$ via a smooth diffeo $\psi_{i}$, and $\bar{U}_{i}$ is compact in $M$.
(ii): $E_{\mid U_{i}}$ is a trivial bundle for each $i$.
and let $\lambda_{i}$ be a partition of unity subordinate to this open covering. Via (i) and (ii) above, identify $U_{i}$ with $\mathbb{R}^{n}, E_{\mid U_{i}}$ with $U_{i} \times \mathbb{C}^{k}$, and using pushforward and pullback under these identifications, identify the Sobolev space $H_{s}\left(U_{i}, E\right)$ as $\left[H_{s}\left(\mathbb{R}^{n}\right)\right]^{k}:=\oplus_{i=1}^{k} H_{s}\left(\mathbb{R}^{n}\right)$. There is a natural Sobolev (direct sum) inner product on this last space, and the resulting Sobolev inner product on $H_{s}\left(U_{i}, E\right)$ is denoted $(-,-)_{i, s}$.

We now define:

$$
H_{s}(M, E):=\left\{f \in \mathcal{E}^{\prime}(M, E)=\mathcal{D}^{\prime}(M, E): \lambda_{i} f \in H_{s}\left(U_{i}, E\right) \quad \text { for each } i=1,2, . ., N\right\}
$$

In fact, we can define the Sobolev inner product on $H_{s}(M, E)$ by the formula:

$$
(f, g)_{s}:=\sum_{i}\left(\lambda_{i} f, \lambda_{i} g\right)_{i, s}
$$

Equip $M$ with a Riemannian metric $g$, which will be fixed once and for all. By the orientability of $M$ there results the global non-vanishing smooth section in $\bigwedge^{n}(M, \mathbb{C})$ called the Riemannian volume form, defined in a local coordinate system by:

$$
d V(x):=\sqrt{\operatorname{det} g_{i j}(x)} d x_{1} \wedge d x_{2} \wedge \ldots \wedge d x_{n}
$$

where $g_{i j}:=g\left(\partial_{i}, \partial_{j}\right)$ is the Gramm matrix of the metric. It is readily checked that the expression above for $d V$ is independent of the coordinate chart.

Similarly, one may equip the complex vector bundle $E$ with a Hermitian bundle metric denoted $\langle-,-\rangle$. If $f, g$ are sections in $C^{\infty}(M, E)$, the function $\langle f(x), g(x)\rangle$ is a smooth $\mathbb{C}$ valued function of $x \in M$, and we may define the global inner product:

$$
(f, g):=\int_{M}\langle f(x), g(x)\rangle d V(x)
$$

which is finite since $M$ is compact. This makes $C^{\infty}(M, \mathbb{C})$ a complex inner-product space, and we denote its completion by $L_{2}(M, E)$, the space of all measurable square integrable sections of $E$.

We can apply the results of the previous subsection and easily deduce the following:

Proposition 4.2.2 (Facts on Sobolev spaces on manifolds).
(i): $H_{0}(M, E) \equiv L_{2}(M, E)$ as Hilbert spaces.
(ii): $C^{\infty}(M, E)$ is dense in $H_{s}(M, E)$ for each $s \in \mathbb{R}$.
(iii): The sesquilinear pairing:

$$
\begin{aligned}
C^{\infty}(M, E) \times C^{\infty}(M, E) & \rightarrow \mathbb{C} \\
f, g & \mapsto(f, g)=\int_{M}\langle f(x), g(x)\rangle d V(x)
\end{aligned}
$$

extends to a sesquilinear pairing $H_{s}(M, E) \times H_{-s}(M, E) \rightarrow \mathbb{C}$ and identifies $H_{-s}(M, E)$ as the Hilbert space dual $\left[H_{s}(M, E)\right]^{*}$.
(iv): (Sobolev Embedding Theorem) There is a continuous inclusion $H_{s} \hookrightarrow C^{k}(M, \mathbb{C})$ for $s>k+n / 2$. This implies that $H_{\infty}(M, E):=\cap_{s \in \mathbb{R}} H_{s}(M, E) \subset C^{\infty}(M, E)$. Since $C^{\infty}(M, E) \subset H_{s}(M, E)$ for all $s$, we have the equality $H_{\infty}(M, E)=C^{\infty}(M, E)$.
(v): $H_{-\infty}(M, E):=\cup_{s \in \mathbb{R}} H_{s}(M, E)=\mathcal{D}^{\prime}(M, E)$
(vi): (Rellich's Lemma) For $s>t$, the inclusion:

$$
H_{s}(M, E) \rightarrow H_{t}(M, E)
$$

is a compact operator, viz. every bounded sequence in $H_{s}$ has a convergent subsequence in $H_{t}$.

Proof: Let $\left\{U_{i}\right\}_{i=1}^{N}$ and $\lambda_{i}$ be as in the beginning of this subsection. Since $K_{i}:=\operatorname{supp} \lambda_{i}$ are compact subsets of $U_{i}$, the measure $d V(x)$ and the Lebesgue measure on $U_{i} \simeq \mathbb{R}^{n}$ are equivalent on $K_{i}$. Similarly, the Hermitian bundle metric $\left\|\|\right.$ on $E$ and the Euclidean metric on $\mathbb{C}^{k}$ are equivalent on $K_{i}$. Hence, for a smooth section $f \in C^{\infty}(M, E)$, we see that the $L_{2}$-norm squared $\int_{M}\left\langle\lambda_{i} f, \lambda_{i} f\right\rangle d V(x)$ is equivalent to the Euclidean $L_{2}$-norm squared of $\lambda_{i} f$ regarded as an element of $\mathcal{E}^{k}$. Since $i=1, . ., N$, the first statement follows.

For (v), let $T=\sum_{i} \lambda_{i} T \in \mathcal{D}^{\prime}(M, E)$, and apply (iv) of Remark 3.1.2 to the compactly supported distributions $\lambda_{i} T$, for $i=1, . ., N$. The remaining statements are direct consequences of corresponding statements of the Propositions 3.1.4, 3.2.1 and 3.3.2 of the last subsection, combined with the remarks of the last paragraph. We leave them as an exercise.

## 5. Pseudodifferential Operators on $\mathbb{R}^{n}$

5.1. Motivation. When one wants to solve a differential equation on a manifold, one basically wants to "invert" a differential operator. This "inverse" is usually not a differential operator. For example, if one wants to solve the equation $\bar{\partial} f=g$ on the plane, for say $g \in \mathcal{S}$, one found in the Propositions 2.2.2 and 2.2.3 that the solution was $g *(2 / z)$, which is given by the integral:

$$
\int_{\mathbb{R}^{2}} \frac{g(w)}{w-z} d V(w)
$$

which is an integral operator acting on $g$. Thus, one needs to enlarge the class of differential operators to include more general operators. The key to this generalisation is the observation that if $P=\sum_{|\alpha| \leq d} a_{\alpha}(x) D_{x}^{\alpha}$ is a differential operator of degree $d, a_{\alpha}$ smooth functions, then for $f \in \mathcal{S}$ say, we have:

$$
P f(x)=\sum_{|\alpha| \leq d} a_{\alpha}(x) D_{x}^{\alpha} f=\sum_{|\alpha| \leq d} a_{\alpha}(x)\left(\widehat{D_{x}^{\alpha} f}\right)^{\vee}(x)=\sum_{|\alpha| \leq d} a_{\alpha}(x)\left(\xi^{\alpha} \widehat{f}\right)^{\vee}(x)=\int_{\mathbb{R}^{n}} e^{i x \cdot \xi} p(x, \xi) \widehat{f}(\xi) d \xi
$$

where $p(x, \xi)=\sum_{|\alpha| \leq d} a_{\alpha}(x) \xi^{\alpha}$ is called the symbol of the differential operator $P$. If the function $f$ was vector valued, taking values in $\mathbb{R}^{k}$, and $P f$ is $\mathbb{R}^{m}$-valued, then the $a_{\alpha}(x)$ would be $m \times k$ matrices, and the symbol $p(x, \xi)$ would be $m \times k$ matrix-valued.

### 5.2. Pseudodifferential operators.

Definition 5.2.1 (Pseudodifferential operators). Let $d \in \mathbb{Z}$. A matrix valued function:

$$
\begin{aligned}
p: \mathbb{R}^{n} \times \mathbb{R}^{n} & \rightarrow \operatorname{hom}_{\mathbb{C}}\left(\mathbb{C}^{k}, \mathbb{C}^{m}\right) \\
(x, \xi) & \mapsto p(x, \xi)
\end{aligned}
$$

is called a symbol of order $d$ if:
(i): $p$ is a smooth map.
(ii): For each pair of multi-indices $\alpha, \beta$, there exists a constant $C_{\alpha \beta}>0$ such that:

$$
\left|D_{x}^{\alpha} D_{\xi}^{\beta} p(x, \xi)\right| \leq C_{\alpha \beta}(1+|\xi|)^{d-|\beta|} \quad \text { for all } x, \xi \in \mathbb{R}^{n}
$$

(Note the norm on the left hand side of the inequality in (ii) is the Hilbert- Schmidt norm on hom $\mathbb{C}\left(\mathbb{C}^{k}, \mathbb{C}^{m}\right)$, defined by $|A|^{2}=\operatorname{tr} A A^{*}=\operatorname{tr} A^{*} A$.)

It is easily checked that the space of symbols of order $d$ form a $\mathbb{C}$ - vector space, which is denoted $S^{d}$. Clearly $S^{d} \subset S^{e}$ if $d \leq e$, and we denote $S^{\infty}=\cup_{d \in \mathbb{Z}} S^{d}$ and $S^{-\infty}=\cap_{d \in \mathbb{Z}} S^{d}$.

For a symbol $p(x, \xi)$ of order $d$, we define the corresponding pseudodifferential operator of order $d$, or $\psi D O$ for short, by the formula:

$$
P f=\int_{M} e^{i x \cdot \xi} p(x, \xi) f(\xi) d \xi
$$

which makes sense at least for $f \in \mathcal{D}^{k}$ of compact support. The space of $\psi D O$ 's of order $d$ is denoted $\Psi^{d}$. If $P$ is a pseudodifferential operator of order $d$, we denote its symbol $p(x, \xi)$ of order $d$ by $\sigma(P)$.
Example 5.2.2 (Linear Differential Operators). Clearly a linear differential operator $P=\sum_{|\alpha| \leq d} a_{\alpha}(x) D_{x}^{\alpha}$ of order $d$ is a $\Psi D O$ of order $d$.

Example 5.2.3 (Convolutions). Let $g \in \mathcal{S}$. Then by Proposition 1.4.15, its Fourier transform $\widehat{g}(\xi)$ is also in $\mathcal{S}$. We also have $D_{\xi}^{\beta} \widehat{g(\xi)} \in \mathcal{S}$ for each $\beta$, and by the rapid decay condition:

$$
\left\|\mid D_{\xi}^{\beta} \widehat{g(\xi)}\right\|_{\infty} \leq C_{\beta}\left(1+|\xi|^{2}\right)^{d}
$$

for each $d \geq 0$ and some $C_{\beta}>0$. Also $D_{x}^{\alpha} D_{\xi}^{\beta} \widehat{g(\xi)} \equiv 0$ for all $|\alpha|>0$, so that $\widehat{g}(\xi)$ is a symbol of every order $d$, and hence belongs to $S^{-\infty}$.

The corresponding $\psi D O$ is defined by:

$$
P f=\int e^{i x \cdot \xi} \widehat{g}(\xi) \widehat{f}(\xi) d \xi=(\widehat{g} \widehat{f})^{\vee}=g * f \quad \text { for } \quad f \in \mathcal{D}
$$

which is just convolution by $g$. It is a $\psi D O$ in $\Psi^{-\infty}$. Thus, in particular, convolution by a smooth compactly supported function is a pseudodifferential operator of infinite order. Convolution is not a differential operator. Hence $\psi D O$ 's are general enough to include both differential operators and integral operators like convolution.

## Remark 5.2.4.

(i): The foregoing example showed how the integral operator of convolution by a rapidly decreasing function defined a pseudodifferential operator. There is a converse to this, namely if $P$ is a $\psi D O$ in $\Psi^{-\infty}$, with symbol $\sigma(P)=p(x, \xi) \in S^{-\infty}$, (that is, the symbol is rapidly decreasing in the $\xi$ direction), then the $\psi D O P$ is an integral operator with smooth kernel. For, let $f \in \mathcal{D}$, then,

$$
\begin{aligned}
P f & =\int e^{i x \cdot \xi} p(x, \xi) \widehat{f}(\xi) d \xi=\int e^{i x . \xi} p(x, \xi) \int e^{-i y . \xi} f(y) d y d \xi \\
& =\int\left(\int e^{i .(x-y)} p(x, \xi) d \xi\right) f(y) d y=\int K(x, y) f(y) d y
\end{aligned}
$$

where the compact $y$-support of $f$ and the rapid decay of $p(x, \xi)$ in $\xi$ allows the interchange of the integrals, and where

$$
K(x, y):=\int e^{i(x-y)} p(x, \xi) d \xi=p^{\vee}(x, x-y)
$$

$p^{\vee}(x,-)$ being the partial inverse Fourier transform of $p$ in the $\xi$ variable. $p$ is rapidly decreasing in $\xi$, and smooth in $x$, so that $p^{\vee}$ is smooth in both variables, and $K(x, y)$ is smooth. Thus $P$ is an integral operator with smooth kernel $K$. Loosely speaking, a general $\psi D O$ is an integral operator with "distributional" kernel $K(x, y)=p^{\vee}(x, x-y)$, since $p^{\vee}$ is in general a distribution.
(ii): Not every integral operator $f \mapsto \int K(x, y) f(y) d y$ with $K(x, y)$ smooth leads to a smoothing operator. For example, taking the smooth kernel $K(\xi, y)=e^{-i \xi \cdot y}$ leads to the integral operator $f \mapsto \widehat{f}$, and say the $C^{\infty}$ function $f(x)=\left(1+x^{2}\right)^{-1} \in \mathcal{E}(\mathbb{R})$ has Fourier transform $\widehat{f}=e^{-|x|}$, which is not even $C^{1}$. However, the next proposition will show that pseudodifferential operators of order $d$ "reduce smoothness" by at most $d$, like constant coefficient differential operators of order $d$ (see (iii) of Proposition 3.1.4).

The way the definition of $\psi D O$ 's is set up, i.e. using the Fourier transform, it behaves well with respect to Schwartz spaces and tempered distributions. More precisely:

Proposition 5.2.5. For $P \in \Psi^{d}$ a $\psi D O$ of order $d$, we have that $P$ is a continuous linear operator of $\mathcal{S}^{k}$ to $\mathcal{S}^{m}$, and hence defines a continuous map of tempered $k$-vector valued distributions $\mathcal{S}^{\prime k}$ to $\mathcal{S}^{\prime m}$. If the $x$-support of $p$ is compact, (i.e. there exists a $K \subset \mathbb{R}^{n}$ such that $p(x,-) \equiv 0$ for all $x \notin K$ ), then $P$ is a bounded operator from $H_{s+d}\left(\mathbb{R}^{n}, \mathbb{C}^{k}\right)$ to $H_{s}\left(\mathbb{R}^{n}, \mathbb{C}^{k}\right)$.

Proof: For simplicity, we will take $k=m=1$, since it is the same argument, with moduli replaced by Hilbert Schmidt norms etc. Let $\Delta_{\xi}=-\sum_{i} \partial_{i}^{2}$ denote the Laplacian in the $\xi$-variable, whose symbol is $p(\xi, x)=|x|^{2}$. Then, for $f \in \mathcal{S}$, we have that $\widehat{f}$ is also in $\mathcal{S}$, and so by the definition of a symbol of order $d, p(x, \xi) \widehat{f}(\xi)$ is in $\mathcal{S}$ in the $\xi$ variable, by using Leibnitz formula. Thus the integral defining $\operatorname{Pf}(x)$ is finite for each $x$, and also we have the inequality:

$$
\left|\Delta_{\xi}^{N}[p(x, \xi) \widehat{f}(\xi)]\right| \leq C_{r, N}\left(1+|\xi|^{2}\right)^{-r}
$$

for any $r>0$. Hence:

$$
\begin{aligned}
\left|x^{2 N} P f(x)\right| & =\left|\int\left(\Delta_{\xi}^{N} e^{i x . \xi}\right) p(x, \xi) \widehat{f}(\xi) d \xi\right|=\left|\int e^{i x . \xi} \Delta_{\xi}^{N}(p(x, \xi) \widehat{f}(\xi)) d \xi\right| \\
& \leq C_{r, N} \int\left(1+|\xi|^{2}\right)^{-r} d \xi
\end{aligned}
$$

where we have used integration by parts for the last equality of the first line, since $p(x, \xi) \widehat{f}(\xi)$ is rapidly decreasing (Schwartz class) in $\xi$. Choosing $r>n / 2$ shows that $|x|^{2 N} P f$ is bounded for all $n$. For the higher derivatives $D_{x}^{\alpha}$ with respect to $x$, we differentiate under the integral sign with respect to $x$ and note
 $D_{x}^{\alpha-\gamma} p(x, \xi)$ by definition, and if $\widehat{f}(\xi) \in \mathcal{S}$, so is $|\xi|^{\gamma} \widehat{f}(\xi)$, so the same argument as above applies to each term in this sum, and we have $P f \in \mathcal{S}$.

To prove the second statement, let $K$ denote the $x$-support of $p(x, \xi)$. For $f \in \mathcal{S}$, we have:

$$
(P f)^{\wedge}(\eta)=\int e^{-i x \cdot \eta} e^{i x . \xi} p(x, \xi) \widehat{f}(\xi) d \xi d x=\int q(\eta-\xi, \xi) \widehat{f}(\xi) d \xi
$$

where the compact $x$-support and rapid decay in $\xi$ of $p(x, \xi) \widehat{f}(\xi)$ (since $f \in \mathcal{S}$ implies $\widehat{f} \in \mathcal{S}$ as well) justifies the change of integrals above. Here

$$
q(\eta, \xi):=\int e^{-i x \cdot \eta} p(x, \xi) d x
$$

is the partial Fourier transform of $p$ in the $x$-direction, which is a Schwartz class function in $\eta$ since $p$ has compact $x$-support. In the $\xi$ variable, $q(\eta, \xi)$ has the same growth properties as $p(x, \xi)$. Putting these two facts together, we have:

$$
|q(\eta, \xi)| \leq C_{k}\left(1+|\xi|^{2}\right)^{d / 2}\left(1+|\eta|^{2}\right)^{-k / 2}
$$

which implies that:

$$
\begin{equation*}
|q(\eta-\xi, \xi)| \leq C_{k}\left(1+|\xi|^{2}\right)^{d / 2}\left(1+|\eta-\xi|^{2}\right)^{-k / 2} \tag{8}
\end{equation*}
$$

where we will conveniently choose $k$ to be large enough later on.
Now, let $g \in \mathcal{S}$. Then by the Plancherel theorem (iv) of Proposition 1.2 .5 and the Cauchy-Schwartz inequality, we have:

$$
\begin{aligned}
\left|(P f, g)_{0}\right| & =\left|(\widehat{P f}, \widehat{g})_{0}\right|=\left|\int q(\eta-\xi, \xi) \widehat{f}(\xi) \overline{\widehat{g}}(\eta) d \xi d \eta\right| \\
& \leq \int|K(\eta, \xi)|^{1 / 2}\left|\left(1+|\xi|^{2}\right)^{s / 2} \widehat{f}(\xi)\right||K(\eta, \xi)|^{1 / 2}\left|\left(1+|\eta|^{2}\right)^{\frac{d-s}{2}} \bar{g}(\eta)\right| d \xi d \eta \\
& \leq\left(\int|K(\eta, \xi)| d \eta\left(1+|\xi|^{2}\right)^{s}|\widehat{f}(\xi)|^{2} d \xi\right)^{1 / 2}\left(\int|K(\eta, \xi)| d \xi\left(1+|\eta|^{2}\right)^{d-s}|\widehat{g}(\eta)|^{2} d \eta\right)^{1 / 2}(9)
\end{aligned}
$$

where

$$
K(\eta, \xi):=q(\eta-\xi, \xi)\left(1+|\xi|^{2}\right)^{-s / 2}\left(1+|\eta|^{2}\right)^{\frac{s-d}{2}}
$$

Because of the inequality (8) above, and Peetre's inequality, we have:

$$
\begin{aligned}
|K(\eta, \xi)| & =\left\lvert\, q(\eta-\xi, \xi)\left(1+|\xi|^{2}\right)^{-s / 2}\left(1+|\eta|^{2}\right)^{\frac{s-d}{2}}\right. \\
& \leq C_{k}\left(1+|\xi|^{2}\right)^{\frac{d-s}{2}}\left(1+|\eta|^{2}\right)^{\frac{s-d}{2}}\left(1+|\eta-\xi|^{2}\right)^{-k / 2} \\
& \leq C_{k}\left(1+|\eta-\xi|^{2}\right)^{\frac{|d-s|-k}{2}}
\end{aligned}
$$

This shows that by choosing $k$ so that $|d-s|-k<-n$, or $k>|d-s|+n$, the integrals:

$$
\int|K(\eta, \xi)| d \eta \leq A ; \quad \int|K(\eta, \xi)| d \xi \leq A
$$

where $A<\infty$ is independent of $\xi, \eta$, so that from the inequality (9) above, we have for $f, g \in \mathcal{S}$ :

$$
\left|(P f, g)_{0}\right| \leq A C_{k}\|f\|_{s}\|g\|_{d-s}
$$

By the density of $\mathcal{S}$ in $H_{s}$ and $H_{d-s}$, we have the same inequality for all $f \in H_{s}$ and all $g \in H_{d-s}$. Then, by (iv) of Proposition 3.1.4, we have for $f \in \mathcal{S}$ that:

$$
\|P f\|_{s-d}=\sup _{g \in H_{d-s} ; g \neq 0} \frac{\left|(P f, g)_{0}\right|}{\|g\|_{d-s}} \leq C\|f\|_{s}
$$

which proves that $P: H_{s} \rightarrow H_{s-d}$ is bounded, and the proposition follows.

Remark 5.2.6. Like the spaces $L_{p, l o c}$, one can define the localised Sobolev spaces:

$$
H_{s, l o c}\left(\mathbb{R}^{n}\right)=\left\{f \in \mathcal{S}^{\prime}: \psi f \in H_{s}\left(\mathbb{R}^{n}\right) \text { for all } \psi \in C_{c}^{\infty}\left(\mathbb{R}^{n}\right)\right\}
$$

Then if one drops the compact $x$-support condition on $\sigma(P)=p(x, \xi)$, one observes that the pseudodifferential operator $\psi P$ defined by $(\psi P) f(x):=\psi(x) P f(x)$ will have the symbol $\sigma(\psi P)=\psi(x) p(x, \xi)$, which will have compact $x$ - support, so that the previous proposition applied to $\psi P$ will yield the fact that $\|(\psi P) f\|_{s-d}<\infty$ for $f \in H_{s}\left(\mathbb{R}^{n}\right)$. That is, $P f \in H_{s-d, l o c}$ for $f \in \mathcal{H}_{s}$. In fact, if one defines a topology on $H_{s-d, l o c}$ by $f_{n} \rightarrow 0$ iff $\psi f_{n} \rightarrow 0$ for each $\psi \in C_{c}^{\infty}\left(\mathbb{R}^{n}\right)$, then the argument above shows that for a general $\psi D O \quad P$ we have $P: H_{s} \rightarrow H_{s-d, l o c}$ a continuous linear map.

## Exercise 5.2.7.

(i): Show that the obvious containment $H_{s} \subset H_{s, l o c}$ is strict for each $s$. In fact, find a function which is in $H_{s, l o c}$ for every $s$, but not in $H_{s}$ for any $s$.
(ii): Show that the localised analogue of the Sobolev lemma holds. That is, if a tempered distribution $f \in \mathcal{S}^{\prime}$ is in $H_{\infty, l o c}:=\cap_{s \in \mathbb{R}} H_{s, l o c}$ then $f \in C^{\infty}$. One can no longer conclude, of course, that $f$ or its derivatives vanish at $\infty$, i.e. in general $f$ won't be in $C_{0}^{\infty}$.

Corollary 5.2.8 (Infinitely smoothing operators). If $P$ is in $\Psi^{-\infty}=\cap_{d} \Psi^{d}$, then $P\left(H_{s}\right) \subset C^{\infty}$ for every $s$. In particular, $P\left(H_{-\infty}\right) \subset C^{\infty}$. (Such operators are called infinitely smoothing. Thus convolutions with $g \in \mathcal{S}$ are infinitely smoothing, by Example 5.2.3.)

Proof: Apply the Remark 5.2.6 and (ii) of the Exercise 5.2.7 above.
5.3. Some Technical Lemmas on $\psi D O^{\prime}$ s. We will need a few lemmas to perform operations with $\psi D O^{\prime} s$. We make a a couple of definitions first.

Definition 5.3.1. Let $p(x, \xi) \in S^{d}$ be a symbol of order $d$ with compact $x$-support $K$. For an open subset $U \subset \mathbb{R}^{n}$, we will say that $p \in S^{d}(U)$ if $K \subset U$. Clearly $S^{d}(U) \subset S^{d}(V)$ for $U \subset V$.

Definition 5.3.2. Let $p, q \in S^{d}(U)$. We will say $p \sim q$ if $p-q \in S^{-\infty}(U):=\cap_{d \in \mathbb{R}} S^{d}(U)$. If $d_{1}>d_{2}>\ldots>$ $d_{j}<\ldots$ is a sequence of real numbers with $d_{j} \rightarrow-\infty$, and $p_{j} \in S^{d_{j}}(U)$ for $j=1,2, \ldots$, we will say $p \sim \sum_{j} p_{j}$ if $p-\sum_{j=1}^{k-1} p_{j} \in S^{d_{k}}(U)$ for all $k$.

Lemma 5.3.3. Let $U$ be a relatively compact open set in $\mathbb{R}^{n}$, and let $d_{1}>d_{2}>\ldots>d_{j}>\ldots$ be a sequence of real numbers with $d_{j} \rightarrow-\infty$. Let $p_{j} \in S^{d_{j}}(U)$ for $j=1,2, \ldots$, . Then for any $V$ containing $\bar{U}$, there is a symbol $p \in S^{d_{1}}(V)$ and such that $p \sim \sum_{j} p_{j}$ in $S^{d_{1}}(V)$.

Proof: By definition, there are constants $C_{\alpha, \beta}^{j}$ satisfying:

$$
\left|D_{x}^{\alpha} D_{\xi}^{\beta} p_{j}(x, \xi)\right| \leq C_{\alpha, \beta}^{j}(1+|\xi|)^{d_{j}-|\beta|}
$$

for all $\alpha, \beta, j$.
Let $\psi \geq 0$ be a smooth function in $C_{c}^{\infty}\left(\mathbb{R}^{n}\right)$ with $\psi(x) \equiv 0$ for $|\xi| \leq 1$ and $\psi \equiv 1$ for $|\xi| \geq 2$. Let $1 \leq r_{1} \leq r_{2} \ldots<r_{j}<\ldots$ be a sequence of positive real numbers with $\lim _{j \rightarrow \infty} r_{j}=\infty$. We define the symbol:

$$
p(x, \xi)=\sum_{k=1}^{\infty} \psi\left(r_{k}^{-1} \xi\right) p_{k}(x, \xi)
$$

For a fixed $\xi,\left|r_{j}^{-1} \xi\right| \leq 1$ for $j$ large enough, so $\psi\left(r_{j}^{-1} \xi\right) \equiv 0$ for $j$ large enough, and the sum on the right is finite, and makes sense. Also, since the $x$-support of each $p_{j}$ is contained in $U$, the $x$-support of $p$ is contained in $\bar{U}$, which is compact. Thus the $x$-support of $p$ is contained in every open set $V \supset \bar{U}$.

To make $p$ a symbol in $S^{d_{1}}(V)$, we need to make a careful choice of $r_{j}$. For each multi-index $\gamma$, let $A_{\gamma}>0$ be a constant so that:

$$
\left|D_{\xi}^{\gamma} \psi(\xi)\right| \leq A_{\gamma} \text { for all } \xi
$$

Then, since $r_{j} \geq 1$ for all $j$, it follows that:

$$
\left|D_{\xi}^{\gamma} \psi\left(r_{j}^{-1} \xi\right)\right| \leq A_{\gamma} r_{j}^{-|\gamma|} \text { for each multi-index } \gamma \text { and all }|\xi| \leq 2 r_{j} ; \equiv 0 \text { for }|\gamma|>0,|\xi|>2 r_{j}
$$

Thus, for any choice of $1 \leq r_{1}<r_{2}<\ldots<r_{j}<\ldots$, we have:

$$
\begin{align*}
\mid D_{x}^{\alpha} D_{\xi}^{\beta}\left(\psi\left(r_{j}^{-1}\right) p_{j}(x, \xi) \mid\right. & \leq \sum_{\gamma \leq \beta}\left|\frac{\beta!D_{\xi}^{\gamma} \psi\left(r_{j}^{-1} \xi\right) D_{x}^{\alpha} D_{\xi}^{\beta-\gamma} p(x, \xi)}{(\gamma)!(\beta-\gamma)!}\right| \\
\leq \sum_{\gamma \leq \beta} \beta!A_{\gamma} C_{\alpha, \beta-\gamma}^{j}(1+|\xi|)^{d_{j}-|\beta|+|\gamma|} r_{j}^{-|\gamma|} & \leq \sum_{\gamma \leq \beta} \beta!A_{j} C_{\alpha, \beta-\gamma}^{j}(1+|\xi|)^{d_{j}-|\beta|}\left(1+2 r_{j}\right)^{|\gamma|} r_{j}^{-|\gamma|} \\
& \leq M_{\alpha, \beta}^{j}(1+|\xi|)^{d_{j}-|\beta|} \tag{10}
\end{align*}
$$

where:

$$
M_{\alpha, \beta}^{j}:=\beta!\sum_{\gamma \leq \beta} 3^{|\gamma|} A_{\gamma} C_{\alpha, \beta-\gamma}
$$

is a positive constant independent of any choice of the sequence $1<r_{1}<r_{2} .<r_{j}<\ldots$. This shows that $\psi\left(r_{j}^{-1} \xi\right) p_{j}(x, \xi)$ is also a symbol of order $d_{j}$, and lies in $S^{d_{j}}(U) \subset S^{d_{j}}(V)$.

For each $k \in \mathbb{Z}_{+}$, define:

$$
M_{k}=\sup \left\{M_{\alpha, \beta}^{k}:|\alpha| \leq k,|\beta| \leq k\right\}
$$

Now choose a sequence of numbers $r_{k}>0$ such that $r_{k} \rightarrow \infty$ and:

$$
\begin{equation*}
\sum_{k=2}^{\infty} \frac{M_{k}}{\left(1+r_{k}\right)^{d_{k-1}-d_{k}}}=C<\infty \tag{11}
\end{equation*}
$$

We need to check that $p$ is a symbol of order $d_{1}$. In fact, we make the more general:

## Claim:

$$
q_{j}:=\sum_{k \geq j} \psi\left(r_{k}^{-1} \xi\right) p_{k}(x, \xi)
$$

is a symbol of order $d_{j}$.
Let $\alpha, \beta$ be multi-indices with $|\alpha|,|\beta| \leq m$. It is clearly enough to check the decay condition for $D_{x}^{\alpha} D_{\xi}^{\beta} q_{j}$ on the set $|\xi| \geq r_{m}$. Also, since $\psi\left(r_{i}^{-1} \xi\right) p_{i}(x, \xi)$ is in $S^{d_{i}}(U)$, the finite sum:

$$
\psi\left(r_{j}^{-1} \xi\right) p_{j}(x, \xi)+\ldots .+\psi\left(r_{m-1}^{-1} \xi\right) p_{m-1}(x, \xi)
$$

is clearly a symbol of order $\max \left\{d_{j}, d_{j+1}, . ., d_{m-1}\right\}=d_{j}$. Thus we just need to verify that:

$$
\left|D_{x}^{\alpha} D_{\xi}^{\beta}\left(q_{m}\right)\right| \leq(\text { const })(1+|\xi|)^{d_{j}-|\beta|} \text { for all }|\xi| \geq r_{m}
$$

We have from (10) that:

$$
\begin{equation*}
\left|D_{x}^{\alpha} D_{\xi}^{\beta}\left(q_{m}\right)\right| \leq \sum_{s \geq 0}\left|D_{x}^{\alpha} D_{\xi}^{\beta}\left[\psi\left(r_{m+s}^{-1} \xi\right) p_{m+s}(x, \xi)\right]\right| \leq \sum_{s \in F(\xi)} M_{\alpha, \beta}^{m+s}(1+|\xi|)^{d_{m+s}-|\beta|} \tag{12}
\end{equation*}
$$

where:

$$
F(\xi)=\left\{s: s \geq 0 \text { and } r_{m+s}<|\xi|\right\}
$$

since $\psi\left(r_{m+s}^{-1} \xi\right) \equiv 0$ for $r_{m+s}^{-1}|\xi| \leq 1$, i.e. for all $s$ such that $r_{m+s} \geq|\xi|$.
Since $|\alpha|,|\beta| \leq m \leq m+s$, we have $M_{\alpha, \beta}^{m+s} \leq M_{m+s}$ for all $s \geq 0$. Also, for an $s \in F(\xi)$, because $d_{m+s}-d_{m}<0$, and $|\xi|>r_{m+s}$, we have the inequality:

$$
(1+|\xi|)^{d_{m+s}-|\beta|}=(1+|\xi|)^{d_{m+s}-d_{m}}(1+|\xi|)^{d_{m}-|\beta|} \leq\left(1+r_{m+s}\right)^{d_{m+s}-d_{m}}(1+|\xi|)^{d_{m}-|\beta|}
$$

Plugging these two facts into the inequality (12), and noting that $d_{m}-d_{m+s} \geq d_{m+s-1}-d_{m+s}$ for $s \geq 1$, we have:

$$
\begin{aligned}
\left|D_{x}^{\alpha} D_{\xi}^{\beta}\left(q_{m}\right)\right| & \leq\left[\sum_{s \in F(\xi)} \frac{M_{m+s}}{\left(1+r_{m+s}\right)^{d_{m}-d_{m+s}}}\right](1+|\xi|)^{d_{m}-|\beta|} \\
& \leq\left[M_{m}+\sum_{s \in F(\xi), s \geq 1} \frac{M_{m+s}}{\left(1+r_{m+s}\right)^{d_{m+s-1}-d_{m+s}}}\right](1+|\xi|)^{d_{m}-|\beta|} \\
& \leq\left[M_{m}+\sum_{k=2}^{\infty} \frac{M_{k}}{\left(1+r_{k}\right)^{d_{k-1}-d_{k}}}\right](1+|\xi|)^{d_{m}-|\beta|} \\
& \leq\left(M_{m}+C\right)(1+|\xi|)^{d_{j}-|\beta|}
\end{aligned}
$$

by the equation (11) and the fact that $d_{m} \leq d_{j}$. This proves the Claim that $q_{j} \in S^{d_{j}}(V)$, and in particular $p=q_{1} \in S^{d_{1}}(V)$.

Also note that for each $j, p_{j}(x, \xi)-\psi\left(r_{j}^{-1} \xi\right) p_{j}(x, \xi)$ has compact support in both $x$ and $\xi$, so is a symbol in $S^{-\infty}(U) \subset S^{-\infty}(V)$. Hence $p_{j}(x, \xi) \sim \psi\left(r_{j}^{-1} \xi\right) p_{j}(x, \xi)$ in $S^{d_{j}}(V)$, so that:

$$
p(x, \xi)-\sum_{j=1}^{k-1} p_{j}(x, \xi) \sim p(x, \xi)-\sum_{j=1}^{k-1} \psi\left(r_{j}^{-1} \xi\right) p_{j}(x, \xi)=q_{j}(x, \xi)
$$

and since $q_{j} \in S^{d_{j}}(V)$, it follows that $p \sim \sum_{j=1}^{\infty} p_{j}$ in $S^{d_{1}}(V)$ and the proposition follows.

The other technical lemma one needs stems from the following observation. Let $P$ be a $\psi D O$ given by the symbol $p(x, \xi)$. Suppose $f \in \mathcal{D}$, and we write the formula for $P f$, viz.,

$$
\begin{aligned}
P f & =\int e^{i x \cdot \xi} p(x, \xi) \widehat{f}(\xi) d \xi \\
& =\int e^{i x \cdot \xi} p(x, \xi) \int e^{-i y \cdot \xi} f(y) d y d \xi
\end{aligned}
$$

which can be viewed (by interchanging the orders of integration) as a special case of:

$$
\begin{equation*}
K f:=\int e^{i(x-y) \cdot \xi} a(x, y, \xi) f(y) d \xi d y \tag{13}
\end{equation*}
$$

where $a(x, y, \xi)=a(x, x, \xi)=p(x, \xi)$ for all $y$. The natural question is: do we enlarge the class of $\psi D O$ 's by using the formula (13) instead of the formula for $\operatorname{Pf}$ in terms of $p(x, \xi)$ in the first line above ?

This is answered by the following definition and lemma.

Definition 5.3.4. A bi-symbol $a(x, y, \xi)$ of order $d$ is a smooth function on $\mathbb{R}^{n} \times \mathbb{R}^{n} \times \mathbb{R}^{n} \rightarrow \operatorname{hom}_{\mathbb{C}}\left(\mathbb{C}^{k}, \mathbb{C}^{m}\right)$ which satisfies:
(i): The $x$-support of $a$ is compact.
(ii): $\left|D_{y}^{\alpha} D_{x}^{\beta} D_{\xi}^{\gamma} a(x, y, \xi)\right| \leq C_{\alpha, \beta, \gamma}(1+|\xi|)^{d-|\gamma|}$, where $|\mid$ on the left hand side denotes Hilbert-Schmidt norm, as usual.
By this definition, a symbol $p(x, \xi) \in S^{d}$ with compact $x$-support is a bi-symbol of order $d$, with $a(x, y, \xi):=$ $a(x, x, \xi):=p(x, \xi)$ for all $y$.

Now we have the answer to our earlier question, in the following:
Lemma 5.3.5. Let $a(x, y, \xi)$ be a bi-symbol of order $d$, and define the operator $K$ by:

$$
K f(x)=\int e^{i(x-y) \cdot \xi} a(x, y, \xi) f(y) d y d \xi \quad \text { for } f \in \mathcal{D}
$$

Then $K$ is a $\psi D O$ of order $d$ whose symbol $k$ has the asymptotic expansion (i.e. upto a symbol in $S^{-\infty}$ ) given by:

$$
\begin{equation*}
k(x, \xi) \sim \sum_{\alpha} \frac{d_{\xi}^{\alpha} D_{y}^{\alpha} a(x, y, \xi)_{\mid y=x}}{\alpha!} \tag{14}
\end{equation*}
$$

Note that in the special case when $a(x, y, \xi)=a(x, x, \xi)=p(x, \xi)$ for all $y$, i.e. the bi-symbol is actually a symbol in disguise, we have $D_{y}^{\alpha} \equiv 0$ for all $|\alpha|>0$, and the expansion above reduces to just its first $\alpha=0$ term, viz. $a(x, x, \xi)$, and this is as it should be.

Proof: We will as usual simplify by assuming that $k=m=1$, because the proof is the same. Since the formula for $K f$ in the statement of this lemma is being defined on $f \in \mathcal{D}=C_{c}^{\infty}\left(\mathbb{R}^{n}\right)$, we can write $f=\psi(y) f(y)$ where $\psi \equiv 1$ on $\operatorname{supp} f$, so that in the formula above, $a(x, y, \xi)$ is replaced by $a(x, y, \xi) \psi(y)$, and we lose no generality in assuming that the $y$-support of $a(x, y, \xi)$ is also compact.

Define the function:

$$
\begin{equation*}
q(x, \rho, \eta):=\int e^{-i y \cdot \rho} a(x, y, \eta) d y \tag{15}
\end{equation*}
$$

which is the Fourier transform of $a(x, y, \eta)$ in the $y$-direction. From this, it follows that:

$$
\begin{equation*}
D_{y}^{\alpha} a(x, y, \eta)_{\mid y=x}=\int e^{i y \cdot \rho} \rho^{\alpha} q(x, \rho, \eta) d \rho_{\mid y=x}=\int e^{i x \cdot \rho} q(x, \rho, \eta) d \rho \tag{16}
\end{equation*}
$$

Now we do some formal manipulations to express $K f$ in the form of a pseudodifferential operator with some symbol, and then check that the alleged symbol is actually a symbol. First note that by the Fourier inversion formula:

$$
\begin{aligned}
\int e^{-i y \cdot \xi} a(x, y, \xi) f(y) d y & =\int e^{-i y \cdot \xi} a(x, y, \xi) \int e^{i y \cdot \eta} \widehat{f}(\eta) d \eta d y \\
=\int\left(\int e^{-i y \cdot(\xi-\eta)} a(x, y, \xi) d y\right) \widehat{f}(\eta) d \eta & =\int q(x, \xi-\eta, \xi) \widehat{f}(\eta) d \eta
\end{aligned}
$$

where the interchange of integrals is allowed since $a(x, y, \xi)$ has compact $y$-support, and $f \in \mathcal{D}$ implies $\widehat{f}(\eta)$ has rapid decay in $\eta$. We also need a precise estimate on the decay of $q(x, \xi-\eta, \xi) \widehat{f}(\eta)$. Since $q(x, \eta, \xi)$ has rapid decay in $\eta$ as stated above, and the same decay as $a(x, y, \xi)$ in $\xi$, we have, for each $k \geq 0$

$$
\begin{equation*}
|q(x, \eta, \xi)| \leq C_{k}(1+|\xi|)^{d}(1+|\eta|)^{-k} \tag{17}
\end{equation*}
$$

Since $\widehat{f}(\eta)$ is rapidly decreasing, we also have, for the same $k$ :

$$
|\widehat{f}(\eta)| \leq C_{k}(1+|\eta|)^{-k}
$$

where $C_{k}$ above (and below) is a generic constant depending on $k$. Hence:

$$
\begin{aligned}
|q(x, \xi-\eta, \xi) \widehat{f}(\eta)| & \leq C_{k}(1+|\xi|)^{d}(1+|\xi-\eta|)^{-k}(1+|\eta|)^{-k} \\
& \leq C_{k}(1+|\xi|)^{d}(1+|\xi|)^{-k}=C_{k}(1+|\xi|)^{d-k}
\end{aligned}
$$

by using the Peetre inequality (see the proof of Proposition 3.3.2) for $k / 2>0$ and the fact that the ratio of $\left(1+r^{2}\right)^{k / 2}$ and $(1+r)^{k}$ is bounded above and below by strictly positive constants independent of $r \geq 0$. By choosing $k$ large enough, we see that $|q(x, \xi-\eta, \xi) \widehat{f}(\eta)|$ is integrable over $\mathbb{R}^{n}$ in $\xi$, as well as $\eta$ (since it is rapidly decreasing in the middle variable).

Now,

$$
\begin{aligned}
K f(x)=\int e^{i(x-y) \cdot \xi} a(x, y, \xi) f(y) d y d \xi & \left.=\int e^{i x \cdot \xi}\left(\int e^{-i y \cdot \xi} a(x, y, \xi) f(y) d y\right)\right) d \xi \\
=\int e^{i x \cdot \xi}\left(\int q(x, \xi-\eta, \xi) \widehat{f}(\eta) d \eta\right) d \xi & =\int e^{i x \cdot \eta}\left(\int e^{i x \cdot(\xi-\eta)} q(x, \xi-\eta, \xi) d \xi\right) \widehat{f}(\eta) d \eta \\
& =\int e^{i x \cdot \eta} p(x, \eta) \widehat{f}(\eta) d \eta
\end{aligned}
$$

where the interchange of $\xi$ and $\eta$ variables is allowed because of the last paragraph, and where we have introduced the function:

$$
p(x, \eta):=\int e^{i x \cdot(\xi-\eta)} q(x, \xi-\eta, \xi) d \xi
$$

Now we check the decay of the derivatives of $p(x, \eta)$ in both variables. This is easily done by changing variables $\rho:=\xi-\eta$, so that:

$$
p(x, \eta)=\int e^{i x \cdot \rho} q(x, \rho, \eta+\rho) d \rho
$$

Applying the estimate (17) above for $q$, we have:

$$
\begin{aligned}
|p(x, \eta)| & \leq C_{k} \int(1+|\eta+\rho|)^{d}(1+|\rho|)^{-k} d \rho \leq C_{k} \int(1+|\rho|)^{|d|}(1+|\eta|)^{d}(1+|\rho|)^{-k} d \rho \\
& \leq C_{k}(1+|\eta|)^{d}\left[\int(1+|\rho|)^{|d|-k} d \rho\right] \leq C_{k}(1+|\eta|)^{d}
\end{aligned}
$$

where we have used Peetre's inequality in the first line above, and chosen $k>|d|+n$. Similarly, by writing down the corresponding estimates for $D_{x}^{\alpha} D_{\xi}^{\beta} q(x, \eta, \xi)$ analogous to (17), one can deduce the estimates for $D_{x}^{\alpha} D_{\eta}^{\beta} p(x, \eta)$ using exactly the same arguments.

To get the asymptotic formula for $p(x, \eta)$, first expand the function $q(x, \rho, \eta+\mu)$ by Taylor's theorem in the third variable, to obtain:

$$
\begin{equation*}
q(x, \rho, \eta+\mu)=\sum_{|\alpha| \leq k} \frac{d_{\eta}^{\alpha} q(x, \rho, \eta)}{\alpha!} \mu^{\alpha}+q_{k}(x, \rho, \eta ; \mu) \tag{18}
\end{equation*}
$$

where $q_{k}(x, \rho, \eta ; \mu)$ is a constant times integral of the derivative $d_{\eta}^{k+1} q(x, \rho, \eta+t \mu)$ over $0 \leq t \leq 1$. Analogous to the inequality (17), since $a(x, y, \eta)$ is compactly supported (hence rapidly decreasing) in the middle variable, that, for all $p \geq 0$ :

$$
\begin{aligned}
\left|q_{k}(x, \rho, \eta ; \mu)\right| & \leq C_{k} \sup _{0 \leq t \leq 1}\left|d_{\eta}^{k+1} q(x, \rho, \eta+t \mu)\right| \leq C_{k} \sup _{0 \leq t \leq 1}(1+|\eta+t \mu|)^{d-k-1}(1+|\rho|)^{-p} \\
& \leq \sup _{0 \leq t \leq 1} C_{k}(1+|t \mu|)^{k+1-d}(1+|\eta|)^{d-k-1}(1+|\rho|)^{-p} \\
& \leq C_{k}(1+|\mu|)^{k+1-d}(1+|\eta|)^{d-k-1}(1+|\rho|)^{-p}
\end{aligned}
$$

by Peetre's inequality, for if $k \gg 0$, we have $|d-k-1|=k+1-d \geq 0$. Hence:

$$
\left|q_{k}(x, \rho, \eta ; \rho)\right| \leq C_{k}(1+|\rho|)^{k+1-d-p}(1+|\eta|)^{d-k-1}
$$

Thus, by choosing $p>k+1-d+n$, we see that:

$$
\begin{equation*}
\left|p_{k}(x, \eta)\right|:=\left|\int e^{i x . \rho} q_{k}(x, \rho, \eta ; \rho) d \rho\right| \leq C_{k}(1+|\eta|)^{d-k-1} \tag{19}
\end{equation*}
$$

and is a symbol in $S^{d-k-1}$. Hence, putting together the equations (16) and (18) we have:

$$
\begin{aligned}
p(x, \eta) & =\int e^{i x \cdot \rho} q(x, \rho, \eta+\rho) d \rho \\
& =\int e^{i x \cdot \rho}\left(\sum_{|\alpha| \leq k} \frac{d_{\eta}^{\alpha} q(x, \rho, \eta)}{\alpha!} \rho^{\alpha}\right) d \rho+p_{k}(x, \eta) \\
& =\sum_{|\alpha| \leq k} \frac{d_{\eta}^{\alpha} D_{y}^{\alpha} a(x, y, \eta)_{y=x}}{\alpha!}+p_{k}(x, \eta)
\end{aligned}
$$

which proves the proposition, in view of the fact that $p_{k}(x, \eta) \in S^{d-k-1}$ for all $k$.

Corollary 5.3.6. Let $a$ and $K$ be as in the previous Proposition 5.3.5. If $a(x, y, \xi)$ vanishes in a neighbourhood of the diagonal $\Delta:=\{(x, x, \xi)\}$, then the $\psi D O$ is infinitely smoothing.

Proof: By hypothesis, $D_{y}^{\alpha} a(x, y, \xi)_{\mid x=y} \equiv 0$, and the asymptotic series of the previous proposition implies the symbol $\sigma(K)=k(x, \xi)$ is equivalent to 0 , i.e. is a symbol in $S^{-\infty}$.

The next corollary is the key to many patching arguments for $\Psi D O$ 's that are going to be used on compact manifolds.

Corollary 5.3.7. Let $\psi:=\left(\psi_{1}, \psi_{2}\right) \in C_{c}^{\infty}\left(\mathbb{R}^{n}\right) \times C_{c}^{\infty}\left(\mathbb{R}^{n}\right)$ be a pair of compactly supported smooth functions, and let $P \in \Psi^{d}$ be a pseudodifferential operator of order $d$. Then the operator defined by:

$$
\left(P^{\psi} f\right)(x):=\psi_{1}(x) P\left(\psi_{2} f\right) \quad \text { for } \quad f \in \mathcal{S}
$$

is also a $\psi D O$ in $\Psi^{d}$.

Proof: By definition, for $f \in \mathcal{S}$, we have:

$$
\left(P^{\psi} f\right)(x)=\psi_{1}(x) \int e^{i x \cdot \xi} p(x, \xi)\left(\psi_{2} f\right)^{\wedge}(\xi) d \xi=\int e^{i(x-y) \cdot \xi} \psi_{1}(x) p(x, \xi) \psi_{2}(y) f(y) d y d \xi
$$

which, by Proposition 5.3.5, implies that it is a $\psi D O$ of order $d$, because the bisymbol

$$
a(x, y, \xi)=\psi_{1}(x) p(x, \xi) \psi_{2}(y)
$$

is a bi-symbol of order $d$, with compact $x$ and $y$ support.

If $L=\sum_{|\alpha| \leq d} a_{\alpha}(x) D_{x}^{\alpha}$ is a linear differential operator, then we have the obvious fact that $L f$ vanishes identically on any neighbourhood on which $f$ vanishes identically. i.e.

$$
\operatorname{supp}(L f) \subset \operatorname{supp} f
$$

for $f \in C^{\infty}$. This property is expressed by saying that linear differential operators are local, they read only the local behaviour of $f$. This is clearly false for pseudodifferential operators, because for example we can take $f \in C_{c}^{\infty}$, which is everywhere $\geq 0$, and convolve it with an everywhere $>0$ Schwartz class function like $e^{-|x|^{2}}$ (which is an infinitely smoothing $\psi D O$ by the example 5.2 .3 , and note that $g * f$ will be strictly positive at all points. However, $\psi D O$ 's have the property that they diminish singular support, i.e.

Proposition 5.3.8 ( $\psi D O$ 's are pseudolocal). If $f \in H_{s}$ for some $s \in \mathbb{R}$, and if $f_{\mid U}$ is a smooth function on some open set $U \subset \mathbb{R}^{n}$, then for every $P \in \Psi^{\infty}$, we have $P f$ is smooth on $U$.

Proof: Let $x \in U$, and let $\psi_{1} \in C_{c}^{\infty}(U)$ with $\psi_{1} \equiv 1$ on a neighbourhood $V \subset U$ of $x$. Let $\psi_{2} \in C_{c}^{\infty}(U)$ with $\psi_{2} \equiv 1$ on a neighbourhood $W \subset \bar{W} \subset U$ of the support of $\psi_{1}$. Clearly, $\psi_{2} f \in C_{c}^{\infty}(U)$, and hence $\psi_{2} f \in \mathcal{S}$. By the Proposition 5.2.5, we have $P \psi_{2} f \in \mathcal{S}$.

On the other hand, since $\psi_{1} P\left(1-\psi_{2}\right)$ is defined by the bi- symbol:

$$
a(x, y, \xi)=\psi_{1}(x) p(x, \xi)\left(1-\psi_{2}(y)\right)
$$

where $p=\sigma(P)$, it is easily checked to be of the same order as $p$. Also since $\left(1-\psi_{2}(y)\right)$ vanishes identically for $y$ contained in the neighbourhood $W$ of $\operatorname{supp} \psi_{1}$, it follows that $a(x, y, \xi)$ vanishes identically on a neighbourhood of the diagonal. Thus the pseudodifferential operator $\psi_{1} P\left(1-\psi_{2}\right)$ is infinitely smoothing, by the Corollary 5.3.6 above. Hence, in the neighbourhood $V$ of $x$, since $\psi_{1} \equiv 1$, we have

$$
P f=\psi_{1} P f=\psi_{1} P \psi_{2} f+\psi_{1} P\left(1-\psi_{2}\right) f
$$

and both the terms on the right are smooth on $V$. Hence the proposition.
5.4. The algebra of $\psi D O$ 's. When $\mathcal{H}$ is a separable Hilbert space, there is the (non- commutative) algebra $L(\mathcal{H})$ of bounded linear operators on $\mathcal{H}$, with multiplication given by composition, and a star operarion given by adjoints. Inside $L(\mathcal{H})$, there is the closed two-sided ideal of compact operators, denoted $K(\mathcal{H})$. Finally, we pass to the quotient, and obtain the Calkin algebra $C(\mathcal{H}):=L(\mathcal{H}) / K(\mathcal{H})$. The so-called Fredholm operators are defined to be the invertible elements in $C \mathcal{H}$ ), i.e. they are invertible modulo compact operators. (These matters will be delved in a future section).

We would like to mimic all this for pseudodifferential operators, with the role of compact operators being played by infinitely smoothing operators. The first task is to define composition and adjoints of $\psi D O$ 's.

Definition 5.4.1 (Adjoints). Let $P$ be a $\psi D O$. For $f \in \mathcal{S}$, define the adjoint $P^{*}$ of $P$ by the formula:

$$
\left(P^{*} f, g\right)=\int\left\langle P^{*} f(x), g(x)\right\rangle d x=(f, P g)=\int\langle f(x), P g(x)\rangle d x \quad \text { for all } g \in \mathcal{S}
$$

This certainly defines $P^{*} f$ as a tempered distribution, for each $f \in \mathcal{S}$. We will eventually check that $P^{*}$ is also a $\psi D O$ of the same order as $P$.

For $P, Q$, two $\psi D O^{\prime}$ s, one defines the composite $P Q$ by $(P Q) f=P(Q f)$ for all $f \in \mathcal{S}$, which makes sense since $P f \in \mathcal{S}$ for $f \in \mathcal{S}$ by the Proposition 5.2.5.

Definition 5.4.2 (Support of a $\psi D O)$. We will say that a $\psi D O P$ is supported in a compact set $K$ if:
(i): supp $P f \subset K$ for all $f \in C_{c}^{\infty}\left(\mathbb{R}^{n}\right)$.
(ii): $P f \equiv 0$ if $f \in C_{c}^{\infty}\left(\mathbb{R}^{n}\right)$ and supp $f \cap K=\phi$

In this event we will say $\operatorname{supp} P=K$.

Exercise 5.4.3. If $P \in \Psi^{d}$, and supp $P \subset K$, then the $x$-support of $p(x, \xi)=\sigma(P)$ is contained in $K$. The converse is false in general, but clearly true for differential operators.

Now we can state the main proposition.

Proposition 5.4.4. Let $P \in \Psi^{d}$ with symbol $\sigma(P)=p$ and $Q \in \Psi^{e}$ with symbol $\sigma(Q)=q$ be two $\psi D O$ 's, with $\operatorname{supp} P, \operatorname{supp} Q$ in some compact set $K \subset \mathbb{R}^{n}$. Then:
(i): $P^{*}$ is a $\psi D O$ of order $d$, supported in $K$, and its symbol is given by the asymptotic formula:

$$
\sigma\left(P^{*}\right) \sim \sum_{\alpha} \frac{d_{x}^{\alpha} D_{\xi}^{\alpha} p^{*}(x, \xi)}{\alpha!}
$$

where $p^{*}(x, \xi)=\bar{p}^{t}(x, \xi)$, the matrix adjoint of $p$.
(ii): The composite $P Q$ is a $\psi D O$ of order $d+e$, supported in $K$, and its symbol is given by the asymptotic expansion:

$$
\sigma(P Q) \sim \sum_{\alpha} \frac{d_{\xi}^{\alpha} p D_{x}^{\alpha} q}{\alpha!}
$$

Proof: We have to just write down a suitable bi-symbol for $P^{*}$ and $P Q$, and appeal to the Proposition 5.3.5. First, for the adjoint we have for, $f, g \in \mathcal{S}$ and $\langle-,-\rangle$ denoting the Hermitian inner product on $\mathbb{C}^{m}$, that:

$$
\begin{aligned}
(f, P g) & =\int\langle f(y), P g(y)\rangle d y=\int e^{-i \xi \cdot y}\langle f(y), p(y, \xi) \widehat{g}(\xi)\rangle d \xi d y \\
& =\int e^{-i \xi \cdot y}\left\langle p^{*}(y, \xi) f(y), \widehat{g}(\xi)\right\rangle d \xi d y=\iint e^{i(x-y) \cdot \xi}\left\langle p^{*}(y, \xi) f(y), g(x)\right\rangle d x d \xi d y \\
& =\left(P^{*} f, g\right)
\end{aligned}
$$

where all changes of integrals are allowed by the rapid decay of $f$ and $g$ and compact $x$-support and rapid $\xi$-decay of $p(x, \xi) \widehat{g}(\xi)$ :

$$
P^{*} f:=\int e^{i(x-y) \cdot \xi} p^{*}(y, \xi) f(y) d \xi d y
$$

which is the $\psi D O$ corresponding to the bisymbol:

$$
a(x, y, \xi)=p^{*}(y, \xi)
$$

It is easy to check from the definition $\left(P^{*} f, g\right)=(f, P g)$ that the support $\operatorname{supp} P^{*} \subset K$ if $\operatorname{supp} P \subset K$. Also the $y$-support of $a(x, y, \xi)$ is contained in $K$, by the previous Exercise 5.4.3, and the $\xi$-decay is the same as that of $p^{*}$, which is the same as that of $p$. So, by the Proposition 5.3.5, we have that $P^{*}$ is a $\psi D O$ of order $d$, and its symbol has the asymptotic expansion:

$$
\sigma\left(P^{*}\right) \sim \sum_{\alpha} \frac{d_{\xi}^{\alpha} D_{x}^{\alpha} p^{*}(x, \xi)}{\alpha!}
$$

which proves (i) of our proposition.
To see (ii), let us first note that if $r(y, \xi):=\sigma\left(Q^{*}\right)$, the symbol of $Q^{*}$, then by definition we have for $f \in \mathcal{S}$ that:

$$
Q^{*} g(y)=\int e^{i y \cdot \xi} r(y, \xi) \widehat{g}(\xi)(y) d y
$$

Now, for $f \in \mathcal{S}$, we have:

$$
\begin{aligned}
(\widehat{Q f}, \widehat{g}) & =(Q f, g)=\left(f, Q^{*} g\right)=\int\left\langle f(y),\left(Q^{*} g\right)(y)\right\rangle d y \\
& =\int\left\langle f(y), e^{i y \cdot \xi} r(y, \xi) \widehat{g}(\xi)\right\rangle d \xi d y \\
& =\int\left\langle\int e^{-i y \cdot \xi} r^{*}(y, \xi) f(y) d y, \widehat{g}(\xi)\right\rangle d \xi
\end{aligned}
$$

which implies that:

$$
\begin{equation*}
\widehat{Q f}(\xi)=\int e^{-i y \cdot \xi} r^{*}(y, \xi) f(y) d y \tag{20}
\end{equation*}
$$

Now, letting $p(x, \xi)=\sigma(P)$, we have by definition, and substitution from (20) above:

$$
P Q f(x)=\int e^{i x . \xi} p(x, \xi)(\widehat{Q f})(\xi) d \xi=\int e^{i(x-y) \cdot \xi} p(x, y) r^{*}(y, \xi) f(y) d y d \xi
$$

which is in the form required by the Lemma 5.3.5, with the bi-symbol:

$$
a(x, y, \xi)=p(x, \xi) r^{*}(y, \xi)
$$

which, by the self-same lemma shows that $P Q$ is a pseudodifferential operator whose symbol has the asympttotic expansion:

$$
\begin{aligned}
\sigma(P Q) & \sim \sum_{\alpha} \frac{d_{\xi}^{\alpha} D_{y}^{\alpha}\left(p(x, \xi) r^{*}(y, \xi)\right)_{\mid y=x}}{\alpha!}=\sum_{\gamma, \alpha} \frac{d_{\xi}^{\alpha-\gamma} p(x, \xi) d_{\xi}^{\gamma} D_{y}^{\alpha} r^{*}(y, \xi)_{\mid y=x}}{\gamma!(\alpha-\gamma)!} \\
& =\sum_{\rho, \delta} \frac{d_{\xi}^{\rho} p(x, \xi) d_{\xi}^{\delta} D_{y}^{\rho+\delta} r^{*}(y, \xi)_{\mid y=x}}{\rho!\delta!}=\sum_{\rho} \frac{d_{\xi}^{\rho} p(x, \xi) D_{x}^{\rho}}{\rho!}\left(\sum_{\delta} \frac{d_{\xi}^{\delta} D_{x}^{\delta} r^{*}(x, \xi)}{\delta!}\right)
\end{aligned}
$$

Now, since $Q=\left(Q^{*}\right)^{*}$, and the symbol of $Q^{*}$ is $r(x, \xi)$, we have by the part (i) above:

$$
\sigma(Q)=q(x, \xi) \sim \sum_{\delta} \frac{d_{\xi}^{\delta} D_{x}^{\delta} r^{*}(x, \xi)}{\delta!}
$$

which on substitution into the last equation above yields:

$$
\sigma(P Q) \sim \sum_{\rho} \frac{d_{\xi}^{\rho} p(x, \xi) D_{x}^{\rho} q(x, \xi)}{\rho!}
$$

and proves (ii) of our proposition. The statements about the supports are readily verified, and left as an exercise.

Corollary 5.4.5. Denote by $\Psi_{K}^{d}$ the space of $\psi D O$ 's with support in $K$, and let $\Psi_{K}^{-\infty}:=\cap_{d} \Psi_{K}^{d}$, and $\Psi_{K}^{\infty}:=$ $\cup_{d} \Psi_{K}^{d}$. Then, by the previous proposition, $\Psi_{K}^{\infty}$ is a (non-commutative) algebra with adjoints.

### 5.5. Ellipticity.

Notation : 5.5.1. From this point onwards, the letter " $P$ " will always denote a linear differential operator or order $d$, so that its symbol $p(x, \xi)$ will always be a polynomial in $\xi$, with coefficients as smooth matrix-valued functions in $x$. In this situation, supp $P$ is contained in $K$ iff the $x$-support of $p(x, \xi)$ is contained in $K$.

Definition 5.5.2. A differential operator $P$ is said to be elliptic over an open set $U \subset \mathbb{R}^{n}$ if:
(i): There exists a constant $C>0$ such that for some $V \supset \bar{U}, p(x, \xi)$ is an invertible linear transformation for all $x \in V$ and all $|\xi| \geq C$, and furthermore,
(ii): The Hilbert-Schmidt norm of the matrix $p(x, \xi)^{-1}$ for $|\xi| \geq C$ satisfies:

$$
\left|p(x, \xi)^{-1}\right| \leq A(1+|\xi|)^{-d} \quad \text { for } \quad x \in V, \quad|\xi| \geq C
$$

In this event, we say that $p(x, \xi)$ is an elliptic symbol of order $d$ over $U$.

Example 5.5.3. It is trivial to check that for any positive integer $d$, the symbol:

$$
p(x, \xi)=\left(1+|\xi|^{2}\right)^{d}
$$

is an elliptic symbol of order $2 d$. If we take $a(x) \in C_{c}^{\infty}$ with support a compact set $K$, then the symbol:

$$
p(x, \xi)=a(x)\left(1+|\xi|^{2}\right)^{d}
$$

will be elliptic over any open set $U$ whose closure is contained in $K$. Thus elliptic symbols of all even orders exist.

Definition 5.5.4 (Leading symbol). For a differential operator $P=\sum_{|\alpha| \leq d} a_{\alpha}(x) D_{x}^{\alpha}$ of order $d$, we define its leading symbol as:

$$
\sigma_{L}(P):=\sum_{|\alpha|=d} a_{\alpha}(x) \xi^{\alpha}
$$

Here is a simple criterion for checking ellipticity of a linear differential operator.
Lemma 5.5.5. $P$ is elliptic over $U$ iff $\sigma_{L}(P)$ is elliptic over $U$.

Proof: Let $P$ be elliptic over $U$, of order $d$, with symbol $p(x, \xi)$. By definition, for $|\xi| \geq C, p(x, \xi)$ is invertible for $x \in V \supset \bar{U}$. Let $q(x, \xi):=(p(x, \xi))^{-1}$ for $x \in V$, and $|\xi| \geq C$. For $t>1$, we have by (ii) of the Definition 5.5.2 that for $x \in V$ and $|\xi| \geq C$.

$$
\mathrm{Id}=p(x, t \xi) q(x, t \xi)=t^{-d} p(x, t \xi) \cdot t^{d} q(x, t \xi)
$$

On taking limits, we find that $\lim _{t \rightarrow \infty} t^{-d} p(x, t \xi)=\sigma_{L}(P)(x, \xi)$, for all $x, \xi$. This implies that

$$
\lim _{t \rightarrow \infty} t^{d} q(x, t \xi)
$$

exists and is finite for $x \in V$ and $|\xi| \geq C$. Call this limit $r(x, \xi)$. It follows that $r(x, \xi)$ is the inverse of $\sigma_{L}(P)(x, \xi)$.

Since

$$
|q(x, t \xi)| \leq A(1+t|\xi|)^{-d} \quad \text { for } x \in V,|\xi| \geq C
$$

we clearly have:

$$
|r(x, \xi)| \leq B(1+|\xi|)^{-d} \text { for } x \in V,|\xi| \geq C
$$

Thus it follows that $r(x, \xi)=\left(\sigma_{L}(P)(x, \xi)\right)^{-1}$ for $x \in V$ and $|\xi| \geq C$, and that $\sigma_{L}(P)$ fulfils both (i) and (ii) of 5.5.2, and hence is an elliptic symbol.

To check the converse, one merely writes:

$$
p(x, \xi)=\sigma_{L}(P)(x, \xi)(I-k(x, \xi))
$$

where $|k(x, \xi)|<1$ for $|\xi|$ large enough. Then one uses the geometric series expansion to get

$$
p(x, \xi)^{-1}=\left(\sigma_{L}(P)(x, \xi)^{-1}\left(I+k(x, \xi)+k(x, \xi)^{2}+\ldots .+\ldots\right)\right.
$$

for $|\xi|$ large enough. We leave the estimate for $\left|p(x, \xi)^{-1}\right|$ as an exercise, it follows from the corresponding estimate for $\sigma_{L}(P)^{-1}$.

Example 5.5.6. If $M$ is a Riemannian manifold, then in a local coordinate chart $U$, we can write the Laplacian of $M$ as:

$$
\Delta=-\sum_{i, j} g^{i j} \partial_{i} \partial_{j}+(\text { lower order terms })
$$

so that on the coordinate chart $U$, its leading symbol is $-\sum_{i, j} g^{i j}(x) \xi_{i} \xi_{j}$, which is certainly elliptic of order 2 all over $U$, since $\left[g^{i j}(x)\right]$ is a positive definite quadratic form for each $x$.

Definition 5.5.7. Let $P \in \Psi^{d}$. We say that the $\psi D O Q$ is a parametrix for $P$ if $Q \in \Psi^{-d}$, and $P Q-I$ and $Q P-I$ are infinitely smoothing operators (i.e. are elements of $\Psi^{-\infty}$ ).

Remark 5.5.8. Note that if $P$ is elliptic of order $d$ over $U, \sigma(P)=p(x, \xi)^{-1}$ exists for all $x \in V$ and $|\xi| \geq C$. It follows that $p(x, \xi)$ is everywhere non-vanishing for $x \in V \supset \bar{U}$ and $|\xi| \geq 2 C$. Thus, if the support of $p$ is a compact set $K(\Leftrightarrow \operatorname{supp} P=K$, since $P$ is a differential operator) we must have $K \supset \bar{V}$.

Definition 5.5.9. Let us say a symbol $s(x, \xi)$ is infinitely smoothing over $V$ if $\psi(x) s(x, \xi) \in S^{-d}(V)$ for all $\psi \in C_{c}^{\infty}(V)$, and all $d$. (See the Definition 5.3.1). A $\psi D O P$ is said to be infinitely smoothing over $V$ if its symbol $p(x, \xi)$ is infinitely smoothing over $V$.

Clearly, since $C_{c}^{\infty}(U) \subset C_{c}^{\infty}(V)$ for $U \subset V$, we have $s$ is infinitely smoothing over $U$ if it is infinitely smoothing over $V \supset U$.

Lemma 5.5.10. Let $p(x, \xi)$ be an elliptic symbol over $U$, of order $d$, and let $V, C$ be as in the Definition 5.5.2, with $\bar{U} \subset V$. Then there exists a symbol $q_{0} \in S^{-d}$ such that:
(i): $p q_{0}-I$ and $q_{0} p-I$ are infinitely smoothing over $V_{1}$, where $V_{1}$ is any open set satisfying $\bar{U} \subset V_{1} \subset \bar{V}_{1} \subset V$.
(ii): If $p$ has compact $x$-support, with $\operatorname{supp}_{x} p=K$, then the $x$-support of $q_{0}$ satisfies:

$$
\operatorname{supp}_{x} q_{0}(x, \xi) \subset V \subset \bar{V} \subset K
$$

Proof: By hypothesis,

$$
\left|p(x, \xi)^{-1}\right| \leq A(1+|\xi|)^{-d} \text { for } x \in V,|\xi| \geq C
$$

Let $\phi(t) \in C_{c}^{\infty}(\mathbb{R})$ such that $\phi \equiv 0$ for $t \leq C$ and $\phi \equiv 1$ for $t \geq 2 C$. Define:

$$
q_{0}(x, \xi)=\phi(|\xi|) p(x, \xi)^{-1} \text { for } \quad x \in V
$$

Multiplying $q_{0}$ above with a function $\psi \in C_{c}^{\infty}(V)$ which is $\equiv 1$ on the subset $V_{1} \subset V$, we can assume that $q_{0}(x, \xi)$ is defined for all $x \in \mathbb{R}^{n}$, and the above equation defining $q_{0}$ holds good for all $x \in V_{1}$.

Thus $p q_{0}-I$ and $q_{0} p-I$ are equal to $(\phi(|\xi|)-1) I$ for all $x \in V_{1}$ and all $\xi$. Since $\left.\phi(|\xi|)-1\right) \equiv 0$ for $|\xi| \geq 2 C$, the operator $(\phi(|\xi|)-1) I$ is infinitely smoothing over $V_{1}$. The proof that $q_{0}$ obeys the decay conditions for a $\psi D O$ of order $(-d)$ follows from the decay condition for $\left|p(x, \xi)^{-1}\right|$ in (ii) of the Definition 5.5.2, and formulas like:

$$
D_{x_{j}}\left(p^{-1}\right)=-p^{-1}\left(D_{x_{j}} p\right) p^{-1}, \quad D_{\xi_{j}}\left(p^{-1}\right)=-p^{-1}\left(D_{\xi_{j}} p\right) p^{-1}
$$

combined with Leibnitz's rule. This proves (i).
For the statement about $x$-supports, note that if $x$-support of $p$ is $K$, then by the Remark 5.5 .8 , we have $\bar{V} \subset K$, the $x$-support of $p$. Since we have multiplied by the compactly supported function $\psi \in C_{c}^{\infty}(V)$ right after the definition of $q_{0}$, we have that the support of $q_{0}$ is a compact subset of $V$, and (ii) follows.

Proposition 5.5.11 (Parametrices for elliptic operators). Let $P$ be an elliptic differential operator of order $d$, elliptic over $U$. Assume that $\operatorname{supp} P \subset K$. Let $V \supset \bar{U}$ as in the Definition 5.5.2. Then, there exists a $\psi D O$ $Q$ of order $-d$ such that for any open set $V_{1}$ satisfying:

$$
\bar{U} \subset V_{1} \subset \bar{V}_{1} \subset V
$$

$P Q-I$ and $Q P-I$ are infinitely smoothing over $V_{1}$.

Proof: Note that by the Remark 5.5.8 above, we must have $K \supset \bar{V} \supset U$.
Define $q_{0} \in S^{-d}$ by Lemma 5.5.10 above, so that $p q_{0}-I$ and $q_{0} p-I$ are infinitely smoothing over $V_{1}$. For $k>0$, we would like to satisfy the formula $\sigma(P Q-I) \sim 0$ and $\sigma(Q P-I) \sim 0$, and we would like $q=\sigma(Q)$ to be a sum:

$$
q \sim q_{0}+q_{1}+\ldots q_{j}+\ldots
$$

with $q_{j} \in S^{-d-j}$, in accordance with the Lemma 5.3.3. From (ii) of the Proposition 5.4.4, we see that $\sigma(P Q-I) \sim 0$ results in the requirements:

$$
p q_{0}-I \sim 0 ; \quad \text { and } \quad \sum_{0 \leq|\alpha| \leq k} \frac{d_{\xi}^{\alpha} p D_{x}^{\alpha} q_{k-|\alpha|}}{\alpha!} \sim 0 \text { for } k>0
$$

where the sum on the right is the homogeneous component of $\sigma(P Q-I)$ which lies in $S^{-k}$ for $k>0$ (Note that $D_{\xi}^{\alpha} p \in S^{d-|\alpha|}$ and $\left.D_{x}^{\alpha} q_{k-|\alpha|} \in S^{-d-k+|\alpha|}\right)$. The first is already satisfied by the definition of $q_{0}$ and the Lemma 5.5.10, and the second may be rewritten as:

$$
p q_{k} \sim-\sum_{0<|\alpha| \leq k} \frac{d_{\xi} p^{\alpha} D_{x}^{\alpha} q_{k-|\alpha|}}{\alpha!} \text { for } k>0
$$

where the right hand side involves only $q_{0}, \ldots, q_{k-1}$. Since $q_{0} p \sim I$, we might as well multiply both sides on the left by $q_{0}$, and define $q_{k}$ by the inductive formula:

$$
q_{k}=-q_{0} \sum_{0<|\alpha| \leq k} \frac{d_{\xi}^{\alpha} p D_{x}^{\alpha} q_{k-|\alpha|}}{\alpha!} \text { for } k>0
$$

Indeed, from this inductive definition, it inductively follows that $q_{k} \in S^{-d-k}$ for all $k$.
If $\operatorname{supp}_{x} p=K$, then by (ii) of the Lemma 5.5 .10 above, we have $\operatorname{supp}_{x} q_{0} \subset V \subset \bar{V} \subset K$, and by the definition of $q_{k}$, we also have $\operatorname{supp}_{x} q_{k} \subset V \subset \bar{V} \subset K$ for all $k$. Then, if one defines $q \sim \sum_{j} q_{j}$ by the Lemma 5.3.3, $q$ will be supported in a subset of $\bar{V}$. At any rate, since the inductive definition forces $P Q-I \sim 0$ on $V_{1}$, we have that $P Q-I$ is infinitely smoothing on $V_{1}$.

By a similar procedure, one may define $Q^{\prime} \in S^{-d}$ such that $Q^{\prime} P-I$ is infinitely smoothing on $V_{1}$. But then since pre or post-composing an infinitely smoothing operator with any $\psi D O$ leads to an infinitely smoothing operator (by (ii) of Proposition 5.4.4), we have:

$$
Q^{\prime} \sim Q^{\prime} . I \sim Q^{\prime} P Q \sim I . Q \sim Q
$$

on $V_{1}$. The proposition follows.

## 6. $\psi D O$ 's and Elliptic Operators on Compact Manifolds

We revert to the setup of $\S 4$. Let $E, F$ be smooth complex vector bundles on a compact manifold $M$, and let $\left\{U_{i}\right\}_{i=1}^{N}$ be an open covering of $M$ such that $U_{i}$ is diffeomorphic to $\mathbb{R}^{n}$ for each $i$, and the restricted bundles $E_{\mid U_{i}}$ and $F_{\mid U_{i}}$ are both trivial (of ranks $k$ and $m$ respectively). $\left\{\lambda_{i}\right\}$ is a smooth partition of unity subordinate to $\left\{U_{i}\right\}$.

### 6.1. Basic definitions and lemmas.

Definition 6.1.1. Let $P: C^{\infty}(M, E) \rightarrow C^{\infty}(M, F)$ be a $\mathbb{C}$-linear operator. We say $P$ is a $\psi D O$ or pseudodifferential operator on $M$ of order $d$ if for all $i, j \in 1,2, \ldots, N$, and all $\psi \in C_{c}^{\infty}\left(U_{j}\right)$ and $\phi \in C_{c}^{\infty}\left(U_{i}\right)$, the "localised operators"

$$
\psi P \phi: C^{\infty}\left(U_{i}, E_{\mid U_{i}}\right) \rightarrow C^{\infty}\left(U_{j}, F_{\mid U_{j}}\right)
$$

are $\psi D O^{\prime} s$ of order $d$, where (the domain and target are identified with $\mathcal{E}^{k}$ and $\mathcal{E}^{m}$ respectively). That this definition makes sense follows from the Corollary 5.3.7. The $\mathbb{C}$-vector space of these $\psi D O$ 's of order $d$ is denoted $\Psi^{d}(M)$, where we have suppressed $E, F$ from the notation for brevity.

Furthermore, we will call $P$ as above a linear differential operator of order $d$ if all the localisations above are differential operators of order $d$. We will call it an elliptic differential operator if each of these localisations $\psi P \phi$ are elliptic over each open set $U$ satisfying $\bar{U} \subset\{x: \phi(x) \psi(x) \neq 0\} \subset U_{i} \cap U_{j}$.

We now have an analogue of the Proposition 5.2.5.

Proposition 6.1.2. Let $P: C^{\infty}(M, E) \rightarrow C^{\infty}(M, F)$ be a $\psi D O$ of order $d$. Then $P$ extends to a continuous (=bounded) linear operator of Hilbert Spaces:

$$
P: H_{s+d}(M, E) \rightarrow H_{s}(M, F)
$$

where the Sobolev spaces $H_{s+d}(M, E)$ and $H_{s}(M, F)$ are as defined in Definition 4.2.1.

Proof: Let $\left\{\lambda_{i}\right\}$ be the partition of unity as described above at the beginning of this section (i.e. as in $\S 4.2$ ), subordinate to the open covering $\left\{U_{i}\right\}_{i=1}^{N}$. By the foregoing definition, we have $\lambda_{i} P \lambda_{j}$ a $\psi D O$, with symbol of compact support. Now, for $f \in C^{\infty}(M, E)$, we compute, using $f=\sum_{j=1}^{N} \lambda_{j} f_{\mid U_{j}}$, that:

$$
\begin{aligned}
\|P f\|_{s}^{2} & =\left\|\sum_{j=1}^{N} P \lambda_{j} f_{\mid U_{j}}\right\|_{s}^{2} \leq C \sum_{j=1}^{N}\left\|P \lambda_{j} f_{\mid U_{j}}\right\|_{s}^{2} \\
& =C \sum_{i, j=1}^{N}\left\|\lambda_{i} P \lambda_{j} f_{\mid U_{j}}\right\|_{s}^{2} \leq C \sum_{i, j=1}^{N} C_{i j}\left\|f_{\mid U_{j}}\right\|_{s+d}^{2} \\
& \leq C\|f\|_{s+d}^{2}
\end{aligned}
$$

where we have used the Definition 4.2.1, Proposition 5.2.5 applied to $\lambda_{i} P \lambda_{j}$ and $\left\|f_{\mid U_{j}}\right\|_{s+d}^{2} \leq\|f\|_{s+d}^{2}$ to arrive at the last line. The proposition follows.

Similarly, one can deduce the pseudolocal property of $\psi D O$ 's on $M$ by appealing to the Proposition 5.3.8, whose statement and proof we leave as an exercise.

Proposition 6.1.3. Let $P: C^{\infty}(M, E) \rightarrow C^{\infty}(M, F)$ be a $\psi D O$ of order $d$. Using Hermitian metrics on $E$ and $F$, gives global $L_{2}$-inner products on $C^{\infty}(M, E)$ and $C^{\infty}(M, F)$ (which we called (,-- ) in (iii) of Proposition 4.2.2), call them $(-,-)_{E}$ and $(-,-)_{F}$ respectively. Define the $L_{2}$-adjoint of $P$ by the formula:

$$
\left(P^{*} f, g\right)_{E}=(f, P g)_{F} \quad \text { for } f \in C^{\infty}(M, F), \quad g \in C^{\infty}(M, E)
$$

Then $P^{*}$ is a $\psi D O$ of order $d$.
If $P: C^{\infty}(M, E) \rightarrow C^{\infty}(M, F)$ is a $\psi D O$ of order $d$, and $Q: C^{\infty}(M, F) \rightarrow C^{\infty}(M, G)$ is a $\psi D O$ of order $e$, the composite $Q P: C^{\infty}(M, E) \rightarrow C^{\infty}(M, G)$ is a $\psi D O$ of order $d+e$.

Proof: Let $\phi, \psi$ be as in Definition 6.1.1. Then, by the definition of $P^{*}$, we have:

$$
\left(\phi P^{*} \psi f, g\right)_{E}=\left(P^{*} \psi f, \bar{\phi} g\right)_{E}=(\psi f, P \bar{\phi} g)_{F}=(f, \bar{\psi} P \bar{\phi} g)_{F}
$$

which implies that $\phi P^{*} \psi=(\bar{\psi} P \bar{\phi})^{*}$. Because the right hand expression is a $\psi D O$ of order $d$ by definition 6.1.1 and (i) of Proposition 5.4.4, it follows that $P^{*}$ is a $\psi D O$ of order $d$.

For the composite, note that if $P$ and $Q$ are $\psi D O$ 's of orders $d$ and $e$, and $\phi$ and $\psi$ are as in the last paragraph, we may write:

$$
\phi P Q \psi=\sum_{i=1}^{N} \phi P \tau_{i} \lambda_{i} Q \psi
$$

where $\tau_{i} \in C_{c}^{\infty}\left(U_{i}\right)$ is a function which is $\equiv 1$ on the support of $\lambda_{i}$, and therefore satisfies $\tau_{i} \lambda_{i} \equiv \lambda_{i}$ for all $i$. Now we can appeal to (ii) of the Proposition 5.4.4 to conclude that each term $\left(\phi P \tau_{i}\right)\left(\lambda_{i} Q \psi\right)$ on the right is a $\psi D O$ of order $d+e$, and hence so is their sum.
6.2. Elliptic operators on manifolds and parametrices. Now we come to the most crucial proposition about elliptic differential operators on compact manifolds.

Proposition 6.2.1 (Parametrices for elliptic operators on manifolds). Let $P: C^{\infty}(M, E) \rightarrow C^{\infty}(M, F)$ be an elliptic differential operator on the compact manifold $M$. Then there exists a $\psi D O Q: C^{\infty}(M, F) \rightarrow$ $C^{\infty}(M, E)$ of order $-d$ such that $P Q-I$ and $Q P-I$ are infinitely smoothing operators.

Proof: It is enough to construct the "left" parametrix satisfying $Q P-I \in \Psi^{-\infty}(M)$, for by the same argument as the last paragraph of Proposition 5.5.11, it serves as the "right" parametrix too.

So let $\lambda_{i}, U_{i}$ be as at the outset of this section. Let us denote:

$$
W_{i}:=\left\{x: \lambda_{i}(x) \neq 0\right\} \subset U_{i}
$$

By the choices and definitions made in the past, the closure $\bar{W}_{i}$ is a compact subset of $U_{i}$ for all $i=1,2 .,, N$. Let $\psi_{i} \in C_{c}^{\infty}\left(U_{i}\right)$ with $\psi_{i} \equiv 1$ on $\bar{W}_{i}$. Let $\rho_{i} \in C_{c}^{\infty}\left(U_{i}\right)$ with $\rho_{i} \equiv 1$ on the support $\operatorname{supp} \psi_{i}$, for $i=1,2, . ., N$.

Consider the localisation $\psi_{i} P \rho_{i}$. It is easy to check that $W_{i}=\left\{x: \lambda_{i}(x) \neq 0\right\}$ is an open subset of

$$
\left\{x: \psi_{i}(x) \neq 0\right\} \cap\left\{x: \rho_{i}(x) \neq 0\right\}
$$

and indeed $\bar{W}_{i}$ is contained in the intersection above. Thus, by the Definition 6.1.1, $\psi_{i} P \rho_{i}$ is elliptic over $W_{i}$.
Since $P$ is a differential operator, and $\rho_{i} \equiv 1$ on $\operatorname{supp} \psi_{i}$, we have $\psi_{i} P \rho_{i}=\psi_{i} P$ for all $i=1,2, . ., N$. Thus $\psi_{i} P$ is elliptic over $W_{i}$. Also $\psi_{i} P$ has support contained in the compact set $K_{i}=\operatorname{supp} \psi_{i}$.

Thus, by the Proposition 5.5.11, there exists an open set $V_{i} \supset W_{i}$ and a parametrix $Q_{i}$ which is a $\psi D O$ of order $-d$ such that $Q_{i}\left(\psi_{i} P\right)-I$ is infinitely smoothing over $V_{i}$. That is, $\lambda\left(Q_{i}\left(\psi_{i} P-I\right)\right.$ is infinitely smoothing on $M$ for all $\lambda \in C_{c}^{\infty}\left(V_{i}\right)$. In particular, since $\operatorname{supp} \lambda_{i}=\bar{W}_{i}$ is a compact subset of $V_{i}$, we have $\lambda_{i}\left(Q_{i}\left(\psi_{i} P\right)-I\right)$ is in $\Psi^{-\infty}(M)$. Hence so is the sum:

$$
\sum_{i} \lambda_{i}\left(Q_{i}\left(\psi_{i} P\right)-I\right)=\sum_{i}\left(\lambda_{i} Q \psi_{i}\right) P-I
$$

since $\sum_{i} \lambda_{i} \equiv 1$. But this means that $Q:=\sum_{i} \lambda_{i} Q \psi_{i}$ is the desired left parametrix. It is of order $-d$ because each term in this sum is of order $-d$.

One of the deepest consequences of the existence of a parametrix for an elliptic differential operator is the following proposition.

Proposition 6.2.2 (Garding-Friedrichs Inequality). Let $P: C^{\infty}(M, E) \rightarrow C^{\infty}(M, F)$ be an elliptic differential operator of order $d$. Then there exists a constant $C>0$ (depending only on $P, M, E$ and $F$ ) such that:

$$
\|f\|_{s+d} \leq C\left(\|P f\|_{s}+\|f\|_{s}\right) \quad \text { for all } \quad f \in H_{s+d}(M, E)
$$

Proof: Let $Q$ be the parametrix for $P$ from the previous Proposition 6.2.1. Then, by definition:

$$
f=Q P f+S f
$$

where $S: C^{\infty}(M, E) \rightarrow C^{\infty}(M, E)$ is infinitely smoothing. Thus

$$
\|f\|_{s+d} \leq\|Q P f\|_{s+d}+\|S f\|_{s+d}
$$

Since $S$ is in $\Psi^{-\infty}(M)$, it is in $\Psi^{-d}(M)$, so by the Proposition 6.1.2, we have:

$$
\|S f\|_{s+d} \leq C\|f\|_{s}
$$

By the same proposition, since $Q \in \Psi^{-d}(M)$, we have:

$$
\|Q P f\|_{s+d} \leq C\|P f\|_{s}
$$

Thus the desired inequality follows.

Corollary 6.2.3 (An equivalent Sobolev norm). Let $P: C^{\infty}(M, E) \rightarrow C^{\infty}(M, E)$ be an elliptic differential operator of order $d$. Let $(-,-)$ denote the global $L_{2}$ inner product on $C^{\infty}(M, E)$ as before. Then the norm associated to the inner product:

$$
\langle f, g\rangle:=(f, g)+(P f, P g) \quad f, g \in C^{\infty}(M, E)
$$

is equivalent to the Sobolev norm $\left\|\|_{d}\right.$ on $C^{\infty}(M, E)$ defined in the Definition 4.2.1. Hence completing $C^{\infty}(M, E)$ with respect to the norm defined by $\langle-,-\rangle$ gives exactly the Sobolev space $H_{d}(M, E)$.

Proof: Let us denote:

$$
\|f\|^{\prime}:=\langle f, f\rangle^{\frac{1}{2}}
$$

for $f \in C^{\infty}(M, E)$. Then, noting that $(-,-)=(-,-)_{0}$, the Sobolev 0-norm, we have

$$
\begin{aligned}
\|f\|^{2} & =\|P f\|_{0}^{2}+\|f\|_{0}^{2} \\
& \leq C\|f\|_{d}^{2}+\|f\|_{d}^{2} \leq C\|f\|_{d}^{2}
\end{aligned}
$$

where we have used the Proposition 6.1.2, and the fact that $\|f\|_{0} \leq\|f\|_{d}$ for $d \geq 0$ in the last line above.
On the other hand, by the Garding-Friedrichs inequality of 6.2 .2 , we have:

$$
\begin{aligned}
\|f\|_{d} & \leq C\left(\|P f\|_{0}+\|f\|_{0}\right) \\
& \leq C\left(\|f\|^{\prime}+\|f\|^{\prime}\right)=2 C\|f\|^{\prime}
\end{aligned}
$$

Thus our proposition follows. Since $H_{d}(M, E)$ is the completion of $C^{\infty}(M, E)$ with respect to $\left\|\|_{d}\right.$, and the last norm is equivalent to $\left\|\|^{\prime}\right.$, the second statement of the proposition follows.

Proposition 6.2.4 (Elliptic Regularity Theorem). Let $P=\sum_{\alpha} a_{\alpha}(x) D_{x}^{\alpha}: C^{\infty}(M, E) \rightarrow C^{\infty}(M, E)$ be an elliptic differential operator of degree $d \geq 1$, so that $P: \mathcal{D}^{\prime}(M, E) \rightarrow \mathcal{D}^{\prime}(M, E)$ gets defined on distributional sections (see the Definition 4.1.4) by the formula:

$$
P f(g)=f\left(\sum_{\alpha}(-1)^{|\alpha|} D_{x}^{\alpha} a_{\alpha}(x)^{t} g\right) \quad \text { for } \quad f \in \mathcal{D}^{\prime}(M, E), g \in C^{\infty}(M, E)=\mathcal{D}(M, E)
$$

Let $f \in \mathcal{D}^{\prime}(M, E)$ be a distributional solution to:

$$
P f=g
$$

where $g \in H_{s}(M, E)$. Then $f \in H_{d+s}(M, E)$. In particular, if $g$ is smooth, then $f$ is also smooth.
Proof: Since $M$ is compact, we have from (v) of Proposition 4.2.2 that $\mathcal{D}^{\prime}(M, E)=\cup_{k} H_{k}(M, E)$. Thus $f \in H_{k}(M, E)$ for some $k$. Let $Q \in \Psi^{-d}(M)$ be a parametrix for $P$, by the Proposition 6.2.1. Then, by definition, the operator $S:=Q P-I \in \Psi^{-\infty}(M)$ is infinitely smoothing, and we have:

$$
f=Q P f+S f=Q g+S f
$$

But since $g \in H_{s}(M, E)$ and $Q \in \Psi^{-d}(M)$, we have $Q g \in H_{s+d}(M, E)$, by Proposition 6.1.2. Also $f \in$ $H_{k}(M, E)$ and $S \in \Psi^{-\infty}(M)$ implies $S \in \Psi^{k-d-s}(M, E)$, so that again by 6.1.2, we have $S f \in H_{d+s}(M, E)$. Thus $f \in H_{d+s}(M, E)$.

If $g \in C^{\infty}(M, E)$, we have $g \in H_{s}(M, E)$ for all $s$ by the Sobolev Embedding Theorem (iv) of Proposition 4.2.2. The last paragraph implies that $f \in H_{s+d}(M, E)$ for all $s$, i.e. $f \in H_{\infty}(M, E)=C^{\infty}(M, E)$ by the same proposition.

## 7. Elliptic operators on $\mathbb{R}^{n}$

7.1. Parametrices on $\mathbb{R}^{n}$. It is quite natural to ask what the analogues of the results obtained in the last section are in the setting of $\mathbb{R}^{n}$.

Definition 7.1.1. Let $P=\sum_{\alpha} a_{\alpha}(x) D_{x}^{\alpha}$ be a linear differential operator of order $d$. Then say that $P$ is elliptic if it is elliptic over a neighbourhood $U$ of each point $x \in \mathbb{R}^{n}$ (in the sense of Definition 5.5.2. (Note that this is weaker than saying that it is elliptic over $\mathbb{R}^{n}$, because we are not demanding one single constant $C$ for all $x \in \mathbb{R}^{n}$ )

Proposition 7.1.2 (Existence of parametrices). Let $P$ be an elliptic linear differential operator on $\mathbb{R}^{n}$ of order $d$. Then there exists a $q(x, \xi)$ of such that:
(i): $\rho(x) q(x, \xi) \in S^{-d}$ for all $\rho \in C_{c}^{\infty}\left(\mathbb{R}^{n}\right)$.
(ii): For a relatively compact subset $W \subset \mathbb{R}^{n}$, let $\rho \in C_{c}^{\infty}\left(\mathbb{R}^{n}\right)$ with $\rho(x) \equiv 1$ for all $x \in W$. Then for the $\psi D O Q$ corresponding to $\rho(x) q(x, \xi)$, the $\psi D O^{\prime} s P Q-I$ and $Q P-I$ are infinitely smoothing over $W$.

Proof: By definition, we have $P$ elliptic over $U_{\alpha}$, for $\left\{U_{\alpha}\right\}$ an open covering of $\mathbb{R}^{n}$. By appealing to paracompactness and second countability of $\mathbb{R}^{n}$, we have a countable locally finite open covering $\left\{U_{i}\right\}_{i=1}^{\infty}$ of $\mathbb{R}^{n}$ such that $P$ is elliptic over $U_{i}$. Let $\left\{\lambda_{i}\right\}$ be a partition of unity subordinate to $\left\{U_{i}\right\}$.

By the Proposition 5.5.11, there are $\psi D O$ 's $Q_{i}$ which satisfy $P Q_{i}-I$ is infinitely smoothing over $V_{1, i}$ where $V_{1, i} \supset \overline{U_{i}}$. That is, $\rho_{i}\left(P Q_{i}-I\right)=S_{i}$, where $S_{i} \in \Psi^{-\infty}$ for all $\rho_{i} \in C_{c}^{\infty}\left(V_{1, i}\right)$. If we take $\rho_{i} \equiv 1$ on $U_{i}$, we have:

$$
P Q_{i}-I=S_{i} ; \quad Q_{i} P-I=T_{i} \quad \text { on } \quad U_{i}
$$

where $S_{i}, T_{i}$ are the restrictions to $U_{i}$ of some infinitely smoothing operators in $\Psi^{-\infty}$. Since we can replace $P$ above by $\rho_{i} P$ on $U_{i}$, we can also assume by the last para of the proof of Proposition 5.5 .11 that the $x$-supports of $q_{i}$ are compact sets for all $i$, as are the $x$-supports of $t_{i}=\sigma\left(T_{i}\right)$ and $s_{i}=\sigma\left(S_{i}\right)$.

The trouble is that $Q_{i}$ and $Q_{j}$ won't generally agree on the overlaps $U_{i} \cap U_{j}$. However, we do know that for $x \in U_{i} \cap U_{j}$, we have:

$$
\begin{aligned}
q_{i}(x, \xi) & =\sigma\left(Q_{i}\right)=\sigma\left(Q_{i} . I\right)=\sigma\left(Q_{i}\left(P Q_{j}-S_{j}\right)\right)=\sigma\left(Q_{i} P Q_{j}-Q_{i} S_{j}\right)=\sigma\left(Q_{j}+T_{i} Q_{j}-Q_{j} S_{i}\right) \\
& =q_{j}(x, \xi)+r_{i j}(x, \xi)
\end{aligned}
$$

where $r_{i j}=\sigma\left(R_{i j}\right):=\sigma\left(T_{i} Q_{j}-Q_{i} S_{j}\right)$. By the formula in (ii) of 5.4.4, the symbols $\sigma\left(T_{i} Q_{j}\right)$ and $\sigma\left(Q_{i} S_{j}\right)$ are also compactly supported, and we may as well assume that the $\operatorname{support}^{\operatorname{supp}} \mathrm{su}_{x}(x, \xi)$ is compact for all $i, j$.

Finally, by 5.5.11, each $R_{i j}$ is the restriction of an infinitely smoothing operator (the pre and post composition of an infinitely smoothing operator with any $\psi D O$ is infinitely smoothing), call it $R_{i j}$ again, to $U_{i} \cap U_{j}$. Thus $r_{i j} \in S^{-\infty}$.

Also note that for $x \in U_{i} \cap U_{j}$ we have $r_{i j}(x, \xi)=-r_{j i}(x, \xi)$, and on the triple intersection $U_{i} \cap U_{j} \cap U_{k}$ we have the cocycle condition on the $r_{i j}$ 's:

$$
r_{i j}(x, \xi)+r_{j k}(x, \xi)+r_{k i}(x, \xi)=\left(q_{i}-q_{j}\right)+\left(q_{j}-q_{k}\right)+\left(q_{k}-q_{i}\right)=0 \quad \text { for } x \in U_{i} \cap U_{j} \cap U_{k}
$$

Now we borrow a trick from sheaf theory and define:

$$
k_{i}(x, \xi)=\sum_{l} \lambda_{l} r_{i l}(x, \xi)
$$

since $\lambda_{j}$ are a partition of unity, the sum on the right makes sense. Unfortunately, $k_{i}$ are no longer compactly supported, and hence the decay conditions on $r_{i j}(x, \xi)$ will no longer translate into global decay conditions for $k_{i}$. However, for any relatively compact subset $W \subset \mathbb{R}^{n}$, $W$ will meet only finitely many of the locally finite collection $U_{i}$, say for $i \in F$. Then, since we have conditions:

$$
\left|D_{x}^{\alpha} D_{\xi}^{\beta} r_{i j}(x, \xi)\right| \leq C_{\alpha, \beta}^{i j}(1+|\xi|)^{-k} \quad \text { for all } i, j, \alpha, \beta, k, x
$$

we will get a corresponding condition:

$$
\left|D_{x}^{\alpha} D_{\xi}^{\beta} k_{i}(x, \xi)\right| \leq C^{i}(W)_{\alpha, \beta}(1+|\xi|)^{-k} \quad \text { for all } i, j, \alpha, \beta, k, x \in \bar{W}
$$

by majorising all the derivatives of $\left\{\lambda_{l}\right\}_{l \in F}$ upto order $\alpha$ and the $C_{\alpha, \beta}^{i j}$ over $\bar{W}$. This implies that that $k_{i}$ is infinitely smoothing over every $W$ which is relatively compact.

Also we have:

$$
k_{i}(x, \xi)-k_{j}(x, \xi)=\sum_{l}\left(\lambda_{l} r_{i l}-\lambda_{l} r_{j l}\right)=\sum_{l} \lambda_{l}\left(-r_{l i}-r_{j l}\right)=\sum_{l} \lambda_{l} r_{i j}=r_{i j}(x, \xi) \text { for } x \in U_{i} \cap U_{j}
$$

This implies:

$$
q_{i}(x, \xi)-q_{j}(x, \xi)=k_{i}(x, \xi)-k_{j}(x, \xi) \text { for } x \in U_{i} \cap U_{j}
$$

which implies that $q_{i}(x, \xi)-k_{i}(x, \xi)=q_{j}(x, \xi)-k_{j}(x, \xi)$ for $x \in U_{i} \cap U_{j}$. Let us define a global function:

$$
q(x, \xi):=q_{i}(x, \xi)-k_{i}(x, \xi) \quad \text { for } x \in U_{i}
$$

Then $q$ makes sense all over $\mathbb{R}^{n}$. It may not be a symbol for the simple reason that $k_{i}$ are no longer globally defined symbols. However, from the decay properties above for $k_{i}$ on a relatively compact open set $W \subset \mathbb{R}^{n}$, it is trivial to check that $\rho Q \in S^{-d}$ for all $\rho \in C_{c}^{\infty}\left(\mathbb{R}^{n}\right)$. It is also readily verified that if $W$ is a relatively compact subset of $\mathbb{R}^{n}$ with $\rho \equiv 1$ on $W$, and $Q$ is the $\psi D O$ corresponding to $\rho(x) q(x, \xi)$, we have $\sigma(P Q-I)$ is infinitely smoothing over $W$. Likewise for $Q P-I$. This proves the proposition.

Definition 7.1.3. Let $W$ be a relatively compact (=bounded) open subset of $\mathbb{R}^{n}$. Define the Sobolev space $H_{s}^{0}(W)$ to be the closure of $C_{c}^{\infty}(W)$ with respect to the Sobolev $s$-norm $\left\|\|_{s}\right.$. Note that it is a closed subspace of $H_{s}\left(\mathbb{R}^{n}\right)$ by definition.

Proposition 7.1.4. Let $W \subset \mathbb{R}^{n}$ be a relatively compact open set, and let $P \in \Psi^{d}$ be a $\psi D O$ or order $d$ with symbol $p(x, \xi) \in S^{d}$. Assume that the $\operatorname{support}_{\operatorname{supp}}^{x} \boldsymbol{p}(x, \xi)$ is a compact. Then

$$
P: H_{s+d}^{0}(W) \rightarrow H_{s}\left(\mathbb{R}^{n}\right)
$$

is a bounded operator. If further the compact subset $K=\operatorname{supp}_{x} p(x, \xi)$ is contained in $W$, then $P$ is a bounded operator from $H_{s+d}^{0}(W) \rightarrow H_{s}^{0}(W)$

Proof: The first statement is clear from the Proposition 5.2.5, because with the compact $x$-support hypothesis imposed on $p$, we have $P: H_{s+d}\left(\mathbb{R}^{n}\right) \rightarrow H_{s}\left(\mathbb{R}^{n}\right)$ is a bounded operator, and $H_{s+d}^{0}(W)$ is a closed subspace of $H_{s+d}\left(\mathbb{R}^{n}\right)$, so the restriction to this subspace is also bounded.

For the second statement, let $f \in H_{s+d}^{0}(W)$, and let $f_{n} \in C_{c}^{\infty}(W)$ be a sequence of smooth functions with $\left\|f_{n}-f\right\|_{s+d} \rightarrow 0$. Since $p$ is compactly supported, and $f_{n}$ are clearly Schwartz class, the Proposition 5.2.5 implies that $P f_{n}$ are smooth Schwartz class functions on $\mathbb{R}^{n}$. Also, the formula:

$$
P f_{n}(x)=\int e^{i x . \xi} p(x, \xi) \widehat{f}_{n}(\xi) d \xi
$$

shows that $\operatorname{supp}_{x} P f_{n} \subset \operatorname{supp}_{x} p(x, \xi)=K \subset W$. Thus $P f_{n} \in C_{c}^{\infty}(W)$. Also, since the $x$-support of $p$ is compact, we have by 5.2.5 that:

$$
\left\|P f_{n}-P f_{m}\right\|_{s} \leq C\left\|f_{n}-f_{m}\right\|_{s+d}
$$

Thus $\left\{P f_{n}\right\}$ is a Cauchy sequence in $H_{s}^{0}(W)$, and since it converges to $P f \in H_{s}\left(\mathbb{R}^{n}\right)$, and the subspace $H_{s}^{0}\left(\mathbb{R}^{n}\right)$ is a closed subspace of $H_{s}\left(\mathbb{R}^{n}\right)$, it follows that $P f \in H_{s}^{0}(W)$. By the first part, the restricted operator:

$$
P: H_{s+d}^{0}(W) \rightarrow H_{s}^{0}(W)
$$

is also a bounded operator. The proposition follows.

Proposition 7.1.5 (Garding-Friedrichs Inequality II). Let $W$ be a relatively compact open subset of $\mathbb{R}^{n}$, and let $P$ be a linear differential operator elliptic over $W$. Then there exists a constant depending only on $W$ and $P$ such that:

$$
\|f\|_{s+d} \leq C\left(\|P f\|_{s}+\|f\|_{s}\right) \quad \text { for } \quad f \in H_{s+d}^{0}(W)
$$

Proof: By hypothesis, there is a open set $V \supset \bar{W}$ and a constant $C$ such that that $p(x, \xi)$ is invertible for $x \in V$ and $|\xi| \geq C$, and the following estimate holds:

$$
\left|p(x, \xi)^{-1}\right| \leq C(1+|\xi|)^{-d} \quad \text { for } \quad x \in V,|\xi| \geq C
$$

Let $\rho \in C_{c}^{\infty}$ be a smooth function which is identically 1 on $V$, and hence identically 1 on $W$. Then since $P$ is a differential operator, we have $\rho P f=P f$ for all $f \in C_{c}^{\infty}(W)$. Also $\rho P$ is clearly elliptic over $W$ by the above criterion, so without loss of generality, we may assume that $p(x, \xi):=\sigma(P)$ has compact $x$-support.

By the Proposition 5.5.11, there exists a $\psi D O Q$ which is of order $(-d)$, also having compact $x$-support for its symbol, and satisfying

$$
Q P-I=S
$$

where $S$ is infinitely smoothing over $V_{1} \supset \bar{V}$. This means $\tau S$ is in $\Psi^{-\infty}$ for every $\tau \in C_{c}^{\infty}\left(V_{1}\right)$. Let us choose a $\tau$ which is identically 1 on $W$. Then we have:

$$
f=\tau f=(\tau I) f=\tau Q P f-\tau S f \quad \text { for } f \in C_{c}^{\infty}(W)
$$

Thus

$$
\|f\|_{s+d} \leq\|\tau Q P f\|_{s+d}+\|(\tau S) f\|_{s+d} \quad \text { for } \quad f \in C_{c}^{\infty}(W)
$$

Since $P$ is a differential operator, $P f \in C_{c}^{\infty}(W)$ as well, and since $\tau Q$ is a compactly supported $\psi D O$ of order $-d$, we have by the first part of the last Proposition 7.1.4 that:

$$
\|\tau Q P f\|_{s+d} \leq C\|P f\|_{s}
$$

Because $\tau S$ is also a compactly supported $\psi D O$ in $\Psi^{-\infty} \subset \Psi^{-d}$, and $f \in C_{c}^{\infty}(W)$, we have similarly:

$$
\|\tau S f\|_{s+d} \leq C\|f\|_{s}
$$

by the same Proposition 7.1.4. Thus we have the desired inequality for all $f \in C_{c}^{\infty}(W)$.
Now let $f \in H_{s+d}^{0}(W)$. Choose a sequence $f_{n} \in C_{c}^{\infty}(W)$ with $f_{n} \rightarrow f$ in $H_{s+d}^{0}(W)$. Since $P$ has compact $x$-support, it follows by the first part of 7.1.4 that $P f_{n} \rightarrow P f$ in $H_{s}\left(\mathbb{R}^{n}\right)$. Since the inclusion $H_{s+d}^{0}(W) \rightarrow$ $H_{s+d}\left(\mathbb{R}^{n}\right) \rightarrow H_{s}\left(R^{n}\right)$ is continuous, we also have $f_{n} \rightarrow f$ in $H_{s}\left(\mathbb{R}^{n}\right)$. Thus the norms $\left\|P f_{n}\right\|_{s} \rightarrow\|P f\|_{s}$ and $\left\|f_{n}\right\|_{s} \rightarrow\|f\|_{s}$. Thus we have:

$$
\|f\|_{s+d}=\lim _{n \rightarrow \infty}\left\|f_{n}\right\|_{s+d} \leq C \lim _{n \rightarrow \infty}\left(\left\|P f_{n}\right\|_{s}+\left\|f_{n}\right\|_{s}\right)=C\left(\|P f\|_{s}+\|f\|_{s}\right)
$$

which proves the proposition.

## 8. Operators on Hilbert Spaces and Fredholm Theory

$\mathcal{H}$ will always denote a separable complex Hilbert space, with inner product denoted $\langle-,-\rangle$, which is $\mathbb{C}$ linear in the first argument and $\mathbb{C}$-antilinear in the second. $\mathcal{B}(\mathcal{H})$ will denote the algebra of bounded operators on $\mathcal{H}$. For $T \in \mathcal{B}(\mathcal{H})$, its adjoint is the operator $T^{*} \in \mathcal{B}(\mathcal{H})$, and is the operator defined by $\langle T x, y\rangle=\left\langle x, T^{*} y\right\rangle$. This defines an involution on $\mathcal{B}(\mathcal{H})$ and makes it $C^{*}$-algebra. More generally, for $T \in \mathcal{B}\left(\mathcal{H}_{1}, \mathcal{H}_{2}\right)$ (=the space of bounded operators from $\mathcal{H}_{1}$ to $\left.\mathcal{H}_{2}\right)$, the adjoint $T^{*} \in \mathcal{B}\left(\mathcal{H}_{2}, \mathcal{H}_{1}\right)$ is defined by the formula $\langle T x, y\rangle_{2}=\left\langle x, T^{*} y\right\rangle_{1}$, where $\langle-,-\rangle_{i}$ are the inner products in the Hilbert spaces $\mathcal{H}_{i}$.

### 8.1. Compact Operators.

Definition 8.1.1 (Compact operator). Let $\mathcal{H}_{1}, \mathcal{H}_{2}$ be Hilbert spaces. Then $T \in \mathcal{B}\left(\mathcal{H}_{1}, \mathcal{H}_{2}\right)$ is said to be a compact operator if for every bounded sequence $x_{n}$ in $\mathcal{H}_{1}$, the sequence $T x_{n}$ in $\mathcal{H}_{2}$ contains a convergent subsequence. Because our Hilbert spaces are always assumed separable, this condition is equivalent to saying that the image $T(B)$ of each bounded set $B \subset \mathcal{H}_{1}$ has compact closure in $\mathcal{H}_{2}$. The subset of compact operators in $\mathcal{B}(\mathcal{H})$ is denoted $\mathcal{K}(\mathcal{H})$.

Example 8.1.2. Clearly, the identity operator $I \in \mathcal{B}(\mathcal{H})$ is a compact operator if and only if $\mathcal{H}$ is finite dimensional.

Example 8.1.3 (Linear maps of finite rank). If $T: \mathcal{H} \rightarrow \mathcal{H}$ is a bounded linear map such that $\operatorname{dim} \operatorname{Im} T<\infty$, then $T$ is a compact operator. For compactness, note that the Heine-Borel theorem for $V=\mathbb{C}^{n}$ implies that every bounded subset of $V$ has compact closure. Thus if $T$ has finite dimensional image, the image $T(B)$ of every bounded subset $B \subset \mathcal{H}$ would be a bounded subset of the finite dimensional space $V=\operatorname{Im} T \subset \mathcal{H}$, and hence have compact closure.

Next, if $T \in \mathcal{B}(\mathcal{H})$ is such that ker $T$ has finite codimension, i.e. $\operatorname{dim}(\operatorname{ker} T)^{\perp}<\infty$, then again $T$ is compact. For then, $T$ would induce a linear embedding:

$$
\widetilde{T}:(\operatorname{ker} T)^{\perp} \rightarrow \mathcal{H}
$$

whose image is the same as $\operatorname{Im} T$. But since $\operatorname{Im} \widetilde{T}$ is finite dimensional, we have $\operatorname{dim} \operatorname{Im} T<\infty$ as well, so $T$ is a bounded operator of finite rank, and a compact operator by the above discussion. Finally, if $\mathcal{H}$ is itself finite dimensional, then $\operatorname{End}_{\mathbb{C}}(\mathcal{H})=\mathcal{B}(\mathcal{H})=\mathcal{K}(\mathcal{H})$.

Example 8.1.4 (Diagonal operators). Let $T \in \mathcal{B}(\mathcal{H})$, and $\left\{e_{n}\right\}$ be an orthonormal basis of $\mathcal{H}$ such that $T e_{n}=\lambda_{n} e_{n}$ for every $n$, where $\lambda_{n} \in \mathbb{C}$. Then (exercise) $T$ is compact iff $\lim _{n \rightarrow \infty} \lambda_{n} \rightarrow 0$.

Example 8.1.5 (The Green Operator on $\left.S^{1}\right)$. The Hilbert space $\mathcal{H}:=L_{2}\left(S^{1}\right)$ has an orthonormal basis $\left\{e_{n}:=e^{i n t}\right\}_{n \in \mathbb{Z}}$ where $0 \leq t<2 \pi$ is the angle parameter on $S^{1}$. The Green operator on $S^{1}$ is the operator defined by:

$$
\begin{aligned}
G: \mathcal{H} & \rightarrow \mathcal{H} \\
e_{n} & \mapsto \frac{e_{n}}{n^{2}} \text { for } n \neq 0 \\
& \mapsto 0 \text { for } n=0
\end{aligned}
$$

In view of the previous example 8.1.4, this operator $G$ is compact. It has the following significance. For the Laplace operator $\Delta: C^{\infty}\left(S^{1}\right) \rightarrow C^{\infty}\left(S^{1}\right)$ on $S^{1}$, defined by $\Delta=\frac{-d^{2}}{d t^{2}}$ on the circle, we have an extension to the domain of $\Delta$, call it $\mathcal{D}:=\operatorname{dom} \Delta \subset \mathcal{H}$. We note that $e_{n}$ satisfy $\Delta e_{n}=n^{2} e_{n}$ for $n \in \mathbb{Z}$. Thus $\mathcal{D}$ consists of all $f=\sum_{n} \widehat{f}(n) e_{n} \in \mathcal{H}$ such that the series $\sum_{n \in \mathbb{Z}} n^{4}|\widehat{f}(n)|^{2}$ is convergent. That is, the sequence $\left\{n^{2} \widehat{f}(n)\right\}_{n \in \mathbb{Z}}$ should be in $l_{2}(\mathbb{Z})$. Note that $\mathcal{D}$ is a proper $L_{2}$-dense linear subspace of $\mathcal{H}$ (it contains each $e_{n}!$ ).

In fact, we see that $\mathcal{D}$ is set-theoretically the Sobolev space $H_{2}\left(S^{1}, E\right)$ for the trivial bundle $E=S^{1} \times \mathbb{C}$. This is because $\Delta$ is clearly an elliptic operator (its leading symbol is $\equiv-1$ in any chart with coordinate $t$ ), and for $f \in C^{\infty}\left(S^{1}\right)$, we have $\Delta f$ is given by convergent Fourier series $\sum_{n \in \mathbb{Z}} n^{2} \widehat{f}(n) e_{n}$, so that the $L_{2}\left(S^{1}\right)$ norm ( $=$ Sobolev 0 -norm $\left\|\|_{0}\right.$ ) of $\Delta f$ is given by:

$$
\|\Delta f\|^{2}=\sum_{n \in \mathbb{Z}} n^{4}|\widehat{f}(n)|^{2}
$$

By the Garding-Friedrichs inequality and its Corollary 6.2.3, we have:

$$
\|f\|_{2}^{2}=\|f\|^{2}+\|\Delta f\|^{2}
$$

and this is finite iff $f \in \mathcal{H}$ and $\Delta f \in \mathcal{H}$, i.e. iff $f \in \mathcal{D}$.
Note that since we are putting the $L_{2}$-norm (and not the Sobolev 2-norm) on $\mathcal{D}$, the operator
$\Delta: \mathcal{D} \rightarrow \mathcal{H}$ is an unbounded operator. Indeed, $\left\|e_{n}\right\|=1$ but $\left\|\Delta e_{n}\right\|=n^{2}$. However, we claim that $\Delta$ has closed range in $\mathcal{H}$, and $\operatorname{Im} \Delta=\left(\mathbb{C} e_{0}\right)^{\perp}$, the closed subspace of all functions in $\mathcal{H}$ which are orthogonal to $e_{0}$, i.e.

$$
\operatorname{Im} \Delta=\left\{f \in \mathcal{H}: \int_{0}^{2 \pi} f(t) d t=0\right\}
$$

This is seen as follows. Note that the Green operator $G$ defined above satisfies the identity:

$$
I_{\mathcal{H}}=\pi_{0}+\Delta G
$$

where $\pi_{0}$ is orthogonal projection onto the space $\mathbb{C} e_{0}=$ ker $\Delta$, and defined by $\pi_{0} f=\left\langle f, e_{0}\right\rangle=\frac{1}{2 \pi} \int_{0}^{2 \pi} f(t) d t$. The above identity makes sense since $G(\mathcal{H}) \subset \mathcal{D}$. It is true on all of $\mathcal{H}$ because it is trivially checked to be true for all $e_{n}, n \in \mathbb{Z}$. It follows that the image $\operatorname{Im} \Delta$ is nothing but $\operatorname{Im}\left(I d-\pi_{0}\right)=\operatorname{Im} \pi_{1}$ where $\pi_{1}: \mathcal{H} \rightarrow\left(\mathbb{C} e_{0}\right)^{\perp}$ is the complementary orthogonal projection to $\pi_{0}$. Thus $\operatorname{Im} \Delta$ is $\left(\mathbb{C} e_{0}\right)^{\perp}$, which is closed. Thus $G$ is an 'inverse' to $\Delta$ on $\operatorname{Im} \Delta$, and gives a Hilbert space isomorphism between $\left(\mathbb{C} e_{0}\right)^{\perp}$ and $\operatorname{Im} \Delta$. Note that ker $G=\operatorname{ker} \Delta=\mathbb{C} e_{0}$.

Similarly, we have the other identity:

$$
\pi_{1 \mid \mathcal{D}}=I_{\mathcal{D}}-\pi_{0 \mid \mathcal{D}}=G \Delta
$$

which holds on $\mathcal{D}$.

Example 8.1.6. Let $M$ be a compact Riemannian manifold. Then, by Rellich's Lemma in (vi) of the Proposition 4.2.2, the inclusion:

$$
i: H_{s}(M, E) \hookrightarrow H_{t}(M, E)
$$

is a compact operator.

Example 8.1.7. If we take a non-compact manifold, say $M=\mathbb{R}$. Then as pointed out in the Exercise 3.3.3, take a fixed function $\phi \in H_{1}(\mathbb{R})$ of $\|\phi\|_{1}=1$, with compact support in say $\left(-\frac{1}{2}, \frac{1}{2}\right)$, and consider its translates $\phi_{n}=\phi(x+n)$. Clearly, by (ii) of the Proposition 3.1.4,

$$
\left\|\phi_{n}\right\|_{1}^{2}=\left\|\phi_{n}\right\|_{0}^{2}+\left\|D_{x} \phi_{n}\right\|_{0}^{2}=\|\phi\|_{1}^{2} \text { for all } n
$$

so that $\left\{\phi_{n}\right\}$ is a bounded sequence in $H_{1}(\mathbb{R})$. But $\left\{\phi_{n}\right\}$ can have no convergent subsequence in $H_{0}(\mathbb{R})$. Indeed, since $\phi_{n}$ and $\phi_{m}$ have disjoint supports for $n \neq m$, we have $\left\langle\phi_{n}, \phi_{m}\right\rangle=0$ for $n \neq m$, which implies $\left\|\phi_{n}-\phi_{m}\right\|_{0}=\sqrt{2}\|\phi\|_{0}$ for all $n \neq m$. Hence $\left\{\phi_{n}\right\}$ cannot have a Cauchy subsequence in $H_{0}(\mathbb{R})$. Thus the inclusion $H_{1}(\mathbb{R}) \hookrightarrow H_{0}(\mathbb{R})$ is not compact.

Example 8.1.8. Let $M$ be a compact Riemannian manifold, and let:

$$
P: H_{s}(M, E) \rightarrow H_{s+d}(M, E)
$$

be any pseudo-differential operator of order $-d<0$ (See the Proposition 6.1.2). Then, the composite:

$$
H_{s}(M, E) \xrightarrow{P} H_{s+d}(M, E) \hookrightarrow H_{s}(M, E)
$$

is a compact operator. This is because $P$ is a bounded operator $H_{s}(M, E) \rightarrow H_{s+d}(M, E)$, and $i: H_{s+d}(M, E) \rightarrow$ $H_{s}(M, E)$ is a compact operator by 8.1.6 above, and it is easy to check that pre or post composing a bounded operator with a compact operator results in a compact operator (See Proposition 8.2.1 below).

In particular, if $Q$ is a parametrix for an elliptic differential operator $P$ on $M$ of order $d>0$, then the composite

$$
H_{s}(M, E) \xrightarrow{Q} H_{s+d}(M, E) \hookrightarrow H_{s}(M, E)
$$

is a compact operator for each $s$, since $Q$ is of order $-d$.
The Green Operator cited in the Example 8.1.5 above is a particular case for $M=S^{1}$ and $E=M \times \mathbb{C}$, the trivial bundle. For $\Delta$ is clearly an elliptic operator of order 2 on $S^{1}$, and by the Proposition 6.2.1, has a parametrix $Q$, which is precisely the operator $G$, because as we remarked above $I-\Delta G$ and $I-G \Delta$ give projection to $e_{0}$, which is the constant function 1 on $S^{1}$, and hence infinitely smoothing. Since $G$ is an operator of order -2 , if we view $G$ as the composite operator:

$$
\mathcal{H}=L_{2}\left(S^{1}\right)=H_{0}\left(S^{1}\right) \rightarrow H_{2}\left(S^{1}\right) \hookrightarrow L_{2}\left(S^{1}\right)=\mathcal{H}
$$

then by the last paragraph, $G$ is a compact operator.
Finally, if $S$ is an infinitely smoothing operator, then for any $s, t \in \mathbb{R}$, we choose $d$ so that $d>t-s$, and since $S \in \Psi^{d}$ for each $d$, we see that the composite:

$$
H_{s}(M, E) \xrightarrow{S} H_{s+d} \rightarrow H_{t}(M, E)
$$

is compact for all $s, t$.

Example 8.1.9. It is natural to wonder what happens for the Laplacian $\Delta=-\sum_{i=1}^{n} \frac{\partial^{2}}{\partial x_{i}^{2}}$ on $\mathbb{R}^{n}$, which is an elliptic differential operator of order 2 on $\mathbb{R}^{n}$. To simplify things, let us take the case of $n=1$, because the sharp contrast with the compact manifold $S^{1}$ considered above are already visible for $n=1$. Indeed, we saw in the Example 8.1.7 above how the inclusion $H_{s}(\mathbb{R}) \hookrightarrow H_{t}(\mathbb{R})$ fails to be compact for $s>t$. This affects everything, as we shall soon see.

The first thing to note is that if $f \in H_{-\infty}$ is a tempered distribution, then $\Delta f=0$ implies that $f$ is smooth. (This is a version of elliptic regularity for $\mathbb{R}^{n}$, which can be deduced from the existence of local parametrices from 5.5.11 applied to $\Delta$ and noting that $f$ is smooth over $U$ iff $\rho f$ is smooth for all $\left.\rho \in C_{c}^{\infty}(U)\right)$.

Thus, for every $s$, the space of harmonic distributions inside the Sobolev space $H_{s}$ is given by:

$$
\{a x+b: a, b \in \mathbb{C}\} \cap H_{s}(\mathbb{R})
$$

from which it follows that $(\operatorname{ker} \Delta) \cap \mathcal{H}=\{0\}$, where we define $\mathcal{H}:=L_{2}(\mathbb{R})=H_{0}(\mathbb{R})$.
The natural domain $\mathcal{D} \subset \mathcal{H}$ for the operator $\Delta$ can also be described. Let $\mathcal{D} \subset \mathbb{R}=\{f \in \mathcal{H}: \Delta f \in \mathcal{H}\}$, which makes sense because for $f \in \mathcal{H}=H_{0}(\mathbb{R}), \Delta f$ is a tempered distribution in $H_{-2}(\mathbb{R})$. We use Plancherel's Theorem (iv) of the Proposition 1.2.5 on $L_{2}(\mathbb{R})$ and the fact (ii) of the same proposition that $(\Delta f)^{\wedge}=\xi^{2} \widehat{f}(\xi)$ to get the commutative diagram:

where $\mathcal{D}_{1}:=\mathcal{D}^{\wedge}$, and the lower horizontal arrow is multiplication by $\xi^{2}$. Note that since $f \mapsto \widehat{f}$ is an isometry, and $\operatorname{ker} \Delta \cap \mathcal{H}=\{0\}$ as noted above, both horizontal maps are injective linear isomorphisms, though not bounded operators.

Since $g \in \mathcal{H}$ iff $\widehat{g} \in \mathcal{H}$, it follows from the diagram above $\mathcal{D}_{1}=\left\{g \in \mathcal{H}: \xi^{2} g \in \mathcal{H}\right\}$, and hence:

$$
\mathcal{D}_{1}=L_{2}(\mathbb{R}, d \mu)
$$

where the measure $d \mu=\left(1+|\xi|^{2}\right)^{2} d \xi$. Hence also the space $\mathcal{D}$ is given by:

$$
\mathcal{D}:=\operatorname{dom} \Delta=\left\{f \in \mathcal{H}: \xi^{2} \widehat{f}(\xi) \in \mathcal{H}\right\}
$$

$\mathcal{D}$ is again an $L_{2}$-dense linear subspace of $\mathcal{H}$, for it contains all Schwartz class (rapidly decreasing) functions. This is similar with with the case of $S^{1}$ discussed in 8.1.5 above, where the condition for $f \in \mathcal{D}$ was that $\left\{n^{2} \widehat{f}(n)\right\}$ should be a square-summable sequence, which again included all $f \in C^{\infty}\left(S^{1}\right)$, a dense subspace. In fact, exactly as in the $S^{1}$-case, one immediately checks by using (ii) of the Proposition 3.1.4 that the conditions $f \in \mathcal{H}$ and $\xi^{2} \widehat{f} \in \mathcal{H}$ imply that $\mathcal{D}$ is set-theoretically the Sobolev space $H_{(2)}(\mathbb{R}) \subset \mathcal{H}=L_{2}(\mathbb{R})$.

But here the analogy ends. It is clear that for a Fourier series $\sum_{n \in \mathbb{Z}} \widehat{f}(n) e_{n}$ on $S^{1}$, the finiteness of $\sum_{n \in \mathbb{Z}}\left|n^{2} \widehat{f}(n)\right|^{2}$ implies the finiteness of $\sum_{n \in \mathbb{Z}}|\widehat{f}(n)|^{2}$. On the other hand we have:
Claim 1: $\xi^{2} \mathcal{D}_{1} \neq \mathcal{H}$, or equivalently, $\Delta(\mathcal{D}) \neq \mathcal{H}$.
Take any $g \in C_{c}^{\infty}(\mathbb{R})$ such that $g(0) \neq 0$, then we claim that the function:

$$
\begin{aligned}
\rho(\xi) & :=\xi^{-2} g(\xi) \text { for } \xi \neq 0 \\
& =0 \text { for } \xi=0
\end{aligned}
$$

is not in $\mathcal{H}$. For, since $g(0) \neq 0$, we have $|g(\xi)|^{2} \geq C>0$ for $\xi \in(0, a)$ and some $a>0$ so that

$$
\|\rho\|^{2} \geq \int_{0}^{a} C \xi^{-4} d \xi=\infty
$$

so that $\rho \notin \mathcal{H}$, so $\rho \notin \mathcal{D}_{1}$, but $\xi^{2} \rho(\xi)=g(\xi)$ is in $\mathcal{H}$. Thus the image of $\mathcal{D}_{1}$ under $\xi^{2}$ is not all of $\mathcal{H}$ and excludes, for example, all compactly supported $g \in C_{c}^{\infty}(\mathbb{R})$ with $g(0) \neq 0$. Hence, for any such $g, g^{\vee} \notin \Delta(\mathcal{D})$, and so $\Delta(\mathcal{D})$ is a proper subspace of $\mathcal{H}$ by the commutative diagram above.

However, we have:
Claim 2: $\xi^{2} \mathcal{D}_{1}$ is dense in $\mathcal{H}$, or equivalently, $\Delta(\mathcal{D})$ is dense in $\mathcal{H}$. (Contrast with $S^{1}$, where $\Delta(\mathcal{D})$ was of codimension 1 in $\mathcal{H}$ )

Let $g \in C_{c}^{\infty}(\mathbb{R})$ be a compactly supported function, then the function:

$$
\begin{aligned}
g_{n}(\xi) & :=g(\xi) \text { for }|\xi| \geq \frac{1}{n} \\
& =0 \text { for }|x| \leq \frac{1}{n}
\end{aligned}
$$

is in $\mathcal{H}$ for each $n$. Again, one computes:

$$
\left\|g_{n}-g\right\|^{2}=\int_{-1 / n}^{1 / n}|g(\xi)|^{2} d \xi \leq \frac{2}{\sqrt{2 \pi} n} \cdot\|g\|_{\infty}^{2}
$$

so that $g_{n} \rightarrow g$ in $\mathcal{H}$. Now $g_{n}=\xi^{2}\left(\xi^{-2} g_{n}\right)$ and $\xi^{-2} g_{n} \in \mathcal{D}_{1}$ since it is bounded and compactly supported, so $g_{n} \in \xi^{2}\left(\mathcal{D}_{1}\right)$. Thus $\xi^{2} \mathcal{D}_{1}$ is dense in $C_{c}^{\infty}(\mathbb{R})$, and since $C_{c}^{\infty}(\mathbb{R})$ is dense in $\mathcal{H}$, we have $\xi^{2} \mathcal{D}_{1}$ is dense in $\mathcal{H}$. The commutative diagram above implies $\Delta(\mathcal{D})$ is dense in $\mathcal{H}$.

Claim 3: $\xi^{2} \mathcal{D}_{1}$ is not closed in $\mathcal{H}$, or equivalently $\Delta(\mathcal{D})$ is not closed in $\mathcal{H}$.
For, if $\xi^{2} \mathcal{D}_{1}$ were closed in $\mathcal{H}$, then Claim 2 above would imply $\xi^{2} \mathcal{D}_{1}=\mathcal{H}$, which would contradict Claim 1. The commutative diagram implies that $\Delta(\mathcal{D}) \neq \mathcal{H}$.

An immediate consequence of Claims 2 and 3 above is that the cokernel Coker $\Delta$ in $\mathcal{H}$ is infinite dimensional. Contrast with $S^{1}$, where the cokernel was the 1 -dimensional space $\mathbb{C} e_{0}$.

Also, in sharp contrast to the case of the circle in 8.1.5, if one formally defines the Green operator on the subspace $\Delta(\mathcal{D})$ to be $\Delta^{-1}$, it would fail to be a compact operator. In fact,

Claim 4: $\xi^{-2}: \xi^{2} \mathcal{D}_{1} \rightarrow \mathcal{D}_{1}$ is an unbounded operator, or equivalently, $G=\Delta^{-1}: \Delta(\mathcal{D}) \rightarrow \mathcal{D}$ is an unbounded operator.

For, let $g \in C_{c}^{\infty}(\mathbb{R})$ with $g(0) \neq 0$, as in the proof of Claim 1 above. Define $\rho(\xi)=\xi^{-2} g(\xi)$ and $\rho_{n}(\xi)=\xi^{-2} g_{n}$, where $g_{n}$ are as in the proof of Claim 2 above. We saw that $g_{n} \rightarrow g$ in $\mathcal{H}$, so that we have:

$$
\left\|\xi^{2} \rho_{n}\right\|=\left\|g_{n}\right\| \rightarrow\|g\|
$$

and hence $\left\|\xi^{2} \rho_{n}\right\|$ is a bounded sequence. However, letting $n>1 / a, a$ as in the proof of Claim 1 above, we have

$$
\left\|\rho_{n}\right\|^{2} \geq \int_{1 / n}^{a}\left|\rho_{n}(\xi)\right|^{2} d \xi=\int_{1 / n}^{a}|\rho(\xi)|^{2} d \xi=\int_{1 / n}^{a} \xi^{-4}|g(\xi)|^{2} d \xi \geq C \int_{1 / n}^{a} \xi^{-4} d \xi \geq A n^{3}
$$

for some $A>0$, which means $\left\|\rho_{n}\right\|$ is an unbounded sequence. Since $\rho_{n}=\xi^{-2} g_{n} \in \mathcal{D}_{1}$ from the proof of Claim 2, it follows that the operator $\xi^{-2}: \xi^{2}\left(\mathcal{D}_{1}\right) \rightarrow \mathcal{D}_{1}$ cannot be a bounded operator. From the commutative diagram above, $G:=\Delta^{-1}: \Delta(\mathcal{D}) \rightarrow \mathcal{D}$ is also not bounded.

Later, we will see how discreteness of the spectrum of $\Delta$ has to do with the compactness of the Green operator, which in turn has to do with the compactness of $M$. Meanwhile, we state a proposition which is the key to many of the results on spectra of the Laplacian, and more generally any self-adjoint elliptic differential operator.

Proposition 8.1.10 (Spectra of self-adjoint compact operators). Let $\mathcal{H}$ be a separable Hilbert space, and let $G \in \mathcal{B}(\mathcal{H})$ be a compact self-adjoint operator. Then there is an orthonormal basis $\left\{e_{n}\right\}_{n=1}^{\infty}$ of $\mathcal{H}$ consisting of eigenvectors of $G$, viz.

$$
G e_{n}=\mu_{n} e_{n} \quad \text { for } n=1,2, \ldots .
$$

with $\mu_{n} \in \mathbb{R}$. Indeed $\left\{\mu_{n}\right\}_{n=1}^{\infty}$ is a bounded sequence, and satisfies $\lim _{n \rightarrow \infty} \mu_{n}=0$.
Proof: That there is an orthonormal basis $\left\{e_{n}\right\}_{n=1}^{\infty}$ of eigenvectors for $G$ is a consequence of the well-known spectral theorem for a bounded self-adjoint operator. That the set of eigenvalues $\left\{\mu_{n}\right\}_{n=1}^{\infty}$ is a bounded subset of $\mathbb{R}$ follows from the boundedness and self-adjointness of $G$.

If $\mu \neq 0$ is a cluster point of $\left\{\mu_{n}\right\}_{n=1}^{\infty}$, then we can find a subsequence $\mu_{n_{k}}$ satisfying $\left|\mu_{n_{k}}\right|>|\mu| / 2$, say. Then, if $B$ is the unit ball in the infinite dimensional subspace $W \subset \mathcal{H}$ spanned by $\left\{e_{n_{k}}\right\}_{k=1}^{\infty}$, the image $G(B)$ will contain the ball $\frac{|\mu|}{2} B$, which is non-compact. Thus $G(B)$ cannot have compact closure, contradicting that $G$ is a compact operator. Thus $\mu=0$, and $\lim _{n \rightarrow \infty} \mu_{n}=0$.
8.2. The Calkin Algebra. Let $\mathcal{H}$ be a complex Hilbert space as above, with inner product $\langle-,-\rangle$. Let $\mathcal{B}(\mathcal{H})$ denote the $C^{*}$-algebra of bounded linear operators on $\mathcal{H}$, and let $\mathcal{K}(\mathcal{H})$ denote the complex linear subspace of compact operators (verify that it is a complex subspace). We have the following easy lemma:

Proposition 8.2.1. $\mathcal{K}(\mathcal{H})$ is a two-sided $*$-ideal in $\mathcal{B}(\mathcal{H})$. Finally $\mathcal{K}(\mathcal{H})$ is closed with respect to the operator norm topology on $\mathcal{B}(\mathcal{H})$.

Proof: Let $\left\{x_{n}\right\}$ be a bounded sequence in $\mathcal{H}, T \in \mathcal{K}(\mathcal{H})$, and $S \in \mathcal{B}(\mathcal{H})$. Then, since there is a convergent subsequence $\left\{T x_{n_{k}}\right\}$, and since $S$ is bounded and hence continuous, the sequence $\left\{S T x_{n_{k}}\right\}$ is also convergent, so $S T$ is a compact operator.

Similarly, since $S$ is bounded, $\left\{S x_{n}\right\}$ is also a bounded sequence in $\mathcal{H}$. By the compactness of $T$, there exists a convergent subsequence $\left\{T S x_{n_{l}}\right\}$ of $\left\{T S x_{n}\right\}$. Thus $T S$ is also a compact operator.

To show $\mathcal{K}(\mathcal{H})$ is a star ideal, we need to show that $T^{*}$ is compact if $T$ is compact. Let $\left\{x_{n}\right\}$ be a bounded sequence in $\mathcal{H}$, with $\left\|x_{n}\right\| \leq A$ for all $n$. Since $T^{*}$ is a bounded operator, we have from the fact that $\mathcal{K}(\mathcal{H})$ is a right ideal that $T T^{*}$ is a compact operator, if $T$ is a compact operator. Thus there exists a subsequence $\left\{T T^{*} x_{n_{k}}\right\}$ which converges. That is, for each $\epsilon>0$, there exists a $N(\epsilon)$ such that

$$
\left\|T T^{*} x_{n_{k}}-T T^{*} x_{n_{l}}\right\|<\epsilon \text { for all } k, l \geq N(\epsilon)
$$

This implies, since $\left\|x_{n_{k}}-x_{n_{l}}\right\| \leq 2 A$ for all $k, l$, and Cauchy-Schwartz, that

$$
\left\|T^{*} x_{n_{k}}-T^{*} x_{n_{l}}\right\|^{2}=\left\langle x_{n_{k}}-x_{n_{l}}, T T^{*} x_{n_{k}}-T T^{*} x_{n_{l}}\right\rangle<2 A \epsilon \text { for all } k, l \geq N(\epsilon)
$$

which shows that the subsequence $\left\{T^{*} x_{n_{k}}\right\}$ is a Cauchy sequence, hence convergent. Thus $T^{*}$ is compact, and $\mathcal{K}(\mathcal{H})$ is a $*$-ideal.

To see that $\mathcal{K}(\mathcal{H})$ is a closed ideal, let $T_{n} \in \mathcal{K}(\mathcal{H})$ be a sequence of compact operators, with $T_{n} \rightarrow T$, and $T \in \mathcal{B}(\mathcal{H})$. We need to show that $T$ is a compact operator. Let $\left\{x_{n}\right\}$ be a bounded sequence in $\mathcal{H}$, with $\left\|x_{n}\right\| \leq A$ for all $n$. Let $\epsilon>0$ be given.

Because of the compactness of all $T_{n}$ 's, we can first find a subsequence $\left\{x_{n}^{1}\right\}$ of $x_{n}$ such that $\left\{T_{1} x_{n}^{1}\right\}$ converges, and then a subsequence $\left\{x_{n}^{2}\right\}$ of $\left\{x_{n}^{1}\right\}$ such that $\left\{T_{2} x_{n}^{2}\right\}$ converges. Clearly then, both $\left\{T_{1} x_{n}^{2}\right\}$ and $\left\{T_{2} x_{n}^{2}\right\}$ converge. Proceeding inductively, for each $j \geq 1$ we have the following:
(i): $\left\{x_{n}^{j}\right\}$ is a subsequence of $\left\{x_{n}^{j-1}\right\}$.
(ii): $\left\{T_{m} x_{n}^{j}\right\}$ is a convergent sequence for all $m \leq j$.

Now consider the diagonal subsequence $\left\{x_{n}^{n}\right\}$ by taking the $n$-th element of the $n$-th subsequence among the $\left\{x_{n}^{j}\right\}$. By (i) above, $\left\{x_{n}^{n}\right\}$ is a subsequence of each of the subsequences $\left\{x_{n}^{j}\right\}$, so it is a subsequence of $\left\{x_{n}\right\}$.

Claim: The sequence $\left\{T x_{n}^{n}\right\}$ is convergent.
For, let $\epsilon>0$. Since $\left\{x_{n}^{n}\right\}$ is a subsequence of each $\left\{x_{n}^{j}\right\}$, it follows by (ii) above that $\left\{T_{j} x_{n}^{n}\right\}$ is a convergent sequence for each $j$. Let its limit be $y_{j}$.

Since $T_{n} \rightarrow T$, there exists an $N \geq 0$ such that

$$
\left\|T_{j}-T\right\|<\epsilon \quad \text { for all } j, k \geq N
$$

where the norm is operator norm. This implies that for $j, k \geq N,\left\|T_{j}-T_{k}\right\|<2 \epsilon$, and hence:

$$
\left\|T_{j} x_{n}^{n}-T_{k} x_{n}^{n}\right\| \leq 2 \epsilon\left\|x_{n}^{n}\right\| \leq 2 A \epsilon \text { for all } j, k \geq N \text { and each } n
$$

Taking the limit $\lim _{n \rightarrow \infty}$ of these inequalities, we obtain:

$$
\left\|y_{j}-y_{k}\right\| \leq 2 A \epsilon \text { for } j, k \geq N
$$

which shows that $\left\{y_{j}\right\}$ is a Cauchy sequence, and hence converges to $y \in \mathcal{H}$.
Thus there is an $N_{1} \geq N>0$ such that $\left\|y_{j}-y\right\|<\epsilon$ for $j \geq N_{1}$. Also there is an $N_{2}>0$ such that

$$
\left\|T_{N_{1}} x_{n}^{n}-y_{N_{1}}\right\|<\epsilon \text { for } n \geq N_{2}
$$

Then for $n \geq N_{2}$, we have:

$$
\begin{aligned}
\left\|T x_{n}^{n}-y\right\| & \leq\left\|T x_{n}^{n}-T_{N_{1}} x_{n}^{n}\right\|+\left\|T_{N_{1}} x_{n}^{n}-y_{N_{1}}\right\|+\left\|y_{N_{1}}-y\right\| \\
& \leq\left\|T-T_{N_{1}}\right\|\left\|x_{n}^{n}\right\|+\epsilon+\epsilon \\
& \leq(A+2) \epsilon
\end{aligned}
$$

which proves that $\left\{T x_{n}^{n}\right\}$ converges to $y$, and hence the claim.
Hence $T$ is compact, and $\mathcal{K}(\mathcal{H})$ is a closed ideal.

Definition 8.2.2. The quotient algebra $\mathcal{B}(\mathcal{H}) / \mathcal{K}(\mathcal{H})$ is called the Calkin Algebra of $\mathcal{H}$, and denoted $\mathcal{C}(\mathcal{H})$. By the lemma 8.2.1 above, this algebra is a Banach $*$-algebra. The star operation in $\mathcal{C}(\mathcal{H})$ is the one induced from $\mathcal{B}(\mathcal{H})$, viz. $[T]^{*}:=\left[T^{*}\right]$. The norm of an element $[T] \in \mathcal{C}(\mathcal{H})$ is defined as:

$$
\|[T]\|=\inf \{\|T+K\|: K \in \mathcal{K}(\mathcal{H})\}
$$

which is a bonafide norm because $\mathcal{K}(\mathcal{H})$ is closed. From the fact that $\mathcal{B}(\mathcal{H})$ is a $C^{*}$-algebra, and the lemma above, it follows (not entirely trivially) that $\mathcal{C}(\mathcal{H})$ is also a $C^{*}$-algebra with this norm.

### 8.3. Fredholm Operators.

Definition 8.3.1. We say that $T \in \mathcal{B}(\mathcal{H})$ is a Fredholm operator if its image $[T] \in \mathcal{C}(\mathcal{H})$ is an invertible element of $\mathcal{C}(\mathcal{H})$. Since $\mathcal{K}(\mathcal{H})$ is a two-sided ideal, $T$ is Fredholm if and only if there exist operators $S, S_{1} \in \mathcal{B}(\mathcal{H})$ such that $S T-I_{\mathcal{H}} \in \mathcal{K}(\mathcal{H})$ and $T S_{1}-I_{\mathcal{H}} \in \mathcal{K}(\mathcal{H})$. (Since inverses are unique in $\mathcal{C}(\mathcal{H})$, we see that $[S]=\left[S_{1}\right]=[T]^{-1}$, i.e., $\left.S-S_{1} \in \mathcal{K}(\mathcal{H})\right)$

Remark 8.3.2. Note that $T$ Fredholm implies that $T^{*}$ is also Fredholm, because $S T-I_{\mathcal{H}}$ (resp. $T S_{1}-I_{\mathcal{H}}$ ) compact implies $T^{*} S^{*}-I_{\mathcal{H}}$ (resp. $S_{1}^{*} T^{*}-I_{\mathcal{H}}$ ) are compact, because $\mathcal{K}(\mathcal{H})$ is a $*$-ideal by lemma 8.2.1.

The definition above is often not very practical, since we have to be lucky enough to hit upon the operators $S$ and $S_{1}$, given the operator $T$. Fortunately, there is a criterion for $T$ to be Fredholm which can be stated entirely in terms of $T$. More precisely:

Proposition 8.3.3 (Fredholm Theorem). Let $T \in \mathcal{B}(\mathcal{H})$. Then $T$ is Fredholm if and only if all the following three criteria are satisfied:
(i): The image $\operatorname{Im} T$ of $T$ is a closed subspace of $\mathcal{H}$.
(ii): The kernel ker $T$ is a finite dimensional subspace of $\mathcal{H}$.
(iii): The cokernel Coker $T:=(\operatorname{Im} T)^{\perp}$ is finite dimensional.

Proof: First let us prove the sufficiency (i.e. the if) part. Let us denote $N:=\operatorname{ker} T, R:=\operatorname{Im} T$, both closed subspaces of $\mathcal{H}$ by hypothesis. Let $V:=N^{\perp}$, and $W:=R^{\perp}$. By hypothesis $\operatorname{dim} N<\infty$ and $\operatorname{dim} W<\infty$. Let $i_{N}, i_{V}, i_{W}, i_{R}$ denote the inclusions of $N, V, W, R$ into $\mathcal{H}$, and similarly let $\pi_{N}, \pi_{V}, \pi_{W}, \pi_{R}$ denote the orthogonal projections onto these closed subspaces.

By definition (and the Open Mapping Theorem), there is an induced map:

$$
T_{1}: V=N^{\perp} \rightarrow R
$$

(viz. the restriction of $T$ to $V=N^{\perp}$ ) which is an isomorphism. Note that $T_{1} \pi_{V}=\pi_{R} T$ and $T i_{V}=i_{R} T$.
Let $Q: R \rightarrow V$ be the inverse of $T_{1}$. Then $Q T_{1}=I_{V}$, and $T_{1} Q=I_{R}$. We need to construct maps $S, S_{1} \in \mathcal{B}(\mathcal{H})$ with $S T-I_{\mathcal{H}}$ and $T S_{1}-I_{\mathcal{H}}$ compact.

Set $S=S_{1}:=i_{V} Q \pi_{R}$. Then $S T=i_{V} Q \pi_{R} T=i_{V} Q T_{1} \pi_{V}=i_{V} I_{V} \pi_{V}=i_{V} \pi_{V}=I_{\mathcal{H}}-i_{N} \pi_{N}$. But $i_{N} \pi_{N} \in \mathcal{B}(\mathcal{H})$ has finite dimensional range, viz. $N$, so it is compact by the example 8.1.3. Hence $S T-I_{\mathcal{H}}$ is compact. Similarly, one checks that $T S_{1}=i_{R} \pi_{R}=I_{\mathcal{H}}-i_{W} \pi_{W}$, so that $T S_{1}-I_{\mathcal{H}}$ is the compact operator $i_{W} \pi_{W}$.

To see the necessity part, assume $T$ is Fredholm. To show that $R:=\operatorname{Im} T$ is closed, let $y_{n}=T x_{n}$ be a sequence in $R$, with $\lim _{n \rightarrow \infty} y_{n}=y \in \mathcal{H}$. We need to show that $y \in R$. Without loss of generality, one can assume that $x_{n} \perp N(T)$ for all $n$. We first claim that $\left\{x_{n}\right\}$ must then be bounded. For if not, assume there is a subsequence $\left\{x_{n_{k}}\right\}$ such that $\left\|x_{n_{k}}\right\| \geq k$. Then set $z_{k}=\left\|x_{n_{k}}\right\|^{-1} x_{n_{k}}$. Then

$$
\lim _{k \rightarrow \infty} T z_{k}=\lim _{k \rightarrow \infty}\left\|x_{n_{k}}\right\|^{-1} T x_{n_{k}}=0
$$

since $T x_{n_{k}} \rightarrow y$. Thus $T z_{k} \rightarrow 0$. By the equation $S T-I=K$ a compact operator, it follows that some subsequence of $K z_{k}$ converges (since $\left\|z_{k}\right\|=1$ ) and thus $z_{k}$ contains a convergent subsequence. Let the limit of that subsequence be $z$. Then $T z=0$ by the above. Thus $z \in N(T)$. On the other hand $\left\|z_{k}\right\|=1$, and $z_{k} \in(N(T))^{\perp}$ implies $\|z\|=1$ and $z \in N(T)^{\perp}$. This is a contradiction, and proves the claim.

Since $x_{n}$ is a bounded sequence, $K x_{n}$ contains a convergent subsequence $x_{n_{k}}$. Also $S T x_{n_{k}}$ converges to $S y$. Thus $x_{n_{k}}=S T x_{n_{k}}-K x_{n_{k}}$ is a convergent sequence, converging to $x$ say. Then clearly $T x=y$.

To show that $N=$ ker $T$ is finite dimensional, let $x \in N$ be any vector. Then $S T x=0=I_{\mathcal{H}} x+K x=x+K x$. Thus $x=-K x$ for all $x \in N$. That is, $I_{N}=-\pi_{N} K i_{N}$, so that $I_{N}$ is a compact operator. From Example 8.1.2, this implies that $N$ is finite dimensional.

By remark 8.3.2, $T^{*}$ is also Fredholm, it follows that $\operatorname{dim} N\left(T^{*}\right)<\infty$ as well, by replacing $T$ with $T^{*}$ in the last paragraph. But by the fact that $R$ is closed, it is easy to check that Coker $T=R^{\perp}=N\left(T^{*}\right)=\operatorname{ker} T^{*}$. Thus Coker $T$ is finite dimensional, and the proposition is proved.
8.4. Two Hilbert Spaces for the price of one. All of the above discussion can be generalised to $\mathcal{B}\left(\mathcal{H}_{1}, \mathcal{H}_{2}\right)$. where $\mathcal{H}_{1}$ and $\mathcal{H}_{2}$ are two different separable Hilbert spaces. This is scarcely surprising, since all infinite dimensional separable Hilbert spaces are (non-canonically) isomorphic to one another, but sometimes it helps to see them as distinct objects. For finite dimensional $\mathcal{H}_{1}$ and $\mathcal{H}_{2}, \mathcal{H}_{1}$ may not be isomorphic to $\mathcal{H}_{2}$, but in that case everything is a tautology from elementary linear algebra.

Note that $\mathcal{B}\left(\mathcal{H}_{1}, \mathcal{H}_{2}\right)$ is no longer an algebra, but just a Banach space. (If both $\mathcal{H}_{1}$ and $\mathcal{H}_{2}$ are infinite dimensional separable Hilbert spaces, we can fix an isomorphism $\Psi: \mathcal{H}_{2} \rightarrow \mathcal{H}_{1}$, then the map $T \mapsto \Psi \circ T$ will be an isomorphism of the Banach space $\mathcal{B}\left(\mathcal{H}_{1}, \mathcal{H}_{2}\right)$ with the Banach space $\mathcal{B}\left(\mathcal{H}_{1}\right)$, and we can use this isomorphism of Banach spaces to define an algebra structure on the former. But, of course, this algebra structure will be non-canonical, and depend on $\Psi$.)

We have already seen in definition 8.1.1 what a compact operator $K: \mathcal{H}_{1} \rightarrow \mathcal{H}_{2}$ is. The subset of compact operators in $\mathcal{B}\left(\mathcal{H}_{1}, \mathcal{H}_{2}\right)$ is denoted $\mathcal{K}\left(\mathcal{H}_{1}, \mathcal{H}_{2}\right)$. It is easily seen to be a closed Banach subspace of $\mathcal{B}\left(\mathcal{H}_{1}, \mathcal{H}_{2}\right)$.

The adjoint defines a $\mathbb{C}$-antilinear isomorphism

$$
\begin{aligned}
*: \mathcal{B}\left(\mathcal{H}_{1}, \mathcal{H}_{2}\right) & \rightarrow \mathcal{B}\left(\mathcal{H}_{2}, \mathcal{H}_{1}\right) \\
T & \mapsto T^{*}
\end{aligned}
$$

We also have the following proposition, whose proof is a trivial generalisation of the proofs of the corresponding propositions for $\mathcal{H}_{1}=\mathcal{H}_{2}=\mathcal{H}$ enunciated in the last two subsections.

Proposition 8.4.1. Let $\mathcal{H}_{1}, \mathcal{H}_{2}, \mathcal{B}\left(\mathcal{H}_{1}, \mathcal{H}_{2}\right), \mathcal{K}\left(\mathcal{H}_{1}, \mathcal{H}_{2}\right)$ etc. be as above. Then:
(i): For $T \in \mathcal{K}\left(\mathcal{H}_{1}, \mathcal{H}_{2}\right)$, and $S_{1} \in \mathcal{B}\left(\mathcal{H}_{3}, \mathcal{H}_{1}\right), S_{2} \in \mathcal{B}\left(\mathcal{H}_{2}, \mathcal{H}_{3}\right), \mathcal{H}_{3}$ any separable Hilbert space, $T \circ S_{1}$ and $S_{2} \circ T$ are compact operators.
(ii): $\mathcal{K}\left(\mathcal{H}_{1}, \mathcal{H}_{2}\right)$ is a closed subspace of $\mathcal{B}\left(\mathcal{H}_{1}, \mathcal{H}_{2}\right)$.
(iii): Under the isomorphism $*$ defined above, $\mathcal{K}\left(\mathcal{H}_{1}, \mathcal{H}_{2}\right)$ maps isomorphically onto $\mathcal{K}\left(\mathcal{H}_{2}, \mathcal{H}_{1}\right)$.
(iv): An operator $T \in \mathcal{B}\left(\mathcal{H}_{1}, \mathcal{H}_{2}\right)$ is said to be Fredholm if there exist operators $S, S_{1} \in \mathcal{B}\left(\mathcal{H}_{2}, \mathcal{H}_{1}\right)$ such that $S T-I_{\mathcal{H}_{1}} \in \mathcal{K}\left(\mathcal{H}_{1}\right)$ and $T S_{1}-I_{\mathcal{H}_{2}} \in \mathcal{K}\left(\mathcal{H}_{2}\right) . T$ is Fredholm iff ker $T$ is finite dimensional, $\operatorname{Im} T$ is closed and Coker $T$ is also finite dimensional. The adjoint $T^{*}$ is also a Fredholm operator if $T$ is a Fredholm operator.
(v): If $T \in \mathcal{B}\left(\mathcal{H}_{1}, \mathcal{H}_{2}\right)$ and $S \in \mathcal{B}\left(\mathcal{H}_{2}, \mathcal{H}_{3}\right)$ are Fredholm, then so is $S T \in \mathcal{B}\left(\mathcal{H}_{1}, \mathcal{H}_{3}\right)$.

We now run through some examples of Fredholm operators.

Example 8.4.2. If $T \in \mathcal{B}\left(\mathcal{H}_{1}, \mathcal{H}_{2}\right)$ is invertible, then clearly $T$ is Fredholm. The composite of two Fredholm operators is also clearly Fredholm.

Example 8.4.3. Obviously, any linear map between two finite dimensional Hilbert spaces $\mathcal{H}_{1}, \mathcal{H}_{2}$ is always Fredholm.

Example 8.4.4 (Unilateral shifts). Let $\left\{e_{n}\right\}_{n=1}^{\infty}$ be an orthonormal basis for a separable Hilbert space $\mathcal{H}$. Then define the unilateral right 1-shift operator:

$$
\begin{aligned}
T: \mathcal{H} & \rightarrow \mathcal{H} \\
e_{i} & \mapsto e_{i+1} \quad \text { for all } i \geq 1
\end{aligned}
$$

This is clearly a Fredholm operator by the proposition 8.3 .3 , for ker $T=\{0\}$, and the range $\operatorname{Im} T=\left(\mathbb{C} e_{1}\right)^{\perp}$ is closed, and the cokernel Coker $T=\mathbb{C} e_{1}$. The adjoint of this operator is easily checked to be:

$$
\begin{aligned}
T^{*}: \mathcal{H} & \rightarrow \mathcal{H} \\
e_{i} & \mapsto e_{i-1} \quad \text { for all } i \geq 2 \\
e_{1} & \mapsto 0
\end{aligned}
$$

As remarked before, $T^{*}$ is also Fredholm, and is called the unilateral ( -1 )-shift operator. Now ker $T^{*}=\mathbb{C} e_{1}$, and Coker $T=\{0\}$. By (v) of the proposition 8.4.1 above, the unilateral $k$-shift $T^{k}$ and the unilateral ( $-k$ )-shift $\left(T^{*}\right)^{k}$ are also Fredholm, and their kernels (resp. cokernels) are $\{0\}$ and $\oplus_{i=1}^{k} \mathbb{C} e_{i}$ (resp. $\oplus_{i=1}^{k} \mathbb{C} e_{i}$ and $\{0\}$ ) respectively.

Example 8.4.5 (Parametrices). Let $M$ be a compact Riemannian manifold, $E$ and $F$ two complex vector bundles on $M$, and let:

$$
P: C^{\infty}(M, E) \rightarrow C^{\infty}(M, F)
$$

be an elliptic differential operator of order $d \geq 1$ (See Definition 6.1.1). Then $P: H_{s+d}(M, E) \rightarrow H_{s}(M, F)$ is a Fredholm operator.

For, by the Proposition 6.2.1, we have a parametrix

$$
Q: H_{s}(M, F) \rightarrow H_{s+d}(M, E)
$$

such that $S:=P Q-I$ is an infinitely smoothing operator on $H_{s}(M, F)$ and $T:=Q P-I$ is an infinitely smoothing operator on $H_{s}(M, E)$. By the Example 8.1.8, it follows that both $S$ and $T$ are compact operators. Thus, by definition, both $P$ and $Q$ are Fredholm operators. Hence, by the Fredholm Theorem 8.3.3, $P\left(H_{s+d}(M, E)\right)$ is closed in $H_{s}(M, F)$, and $\operatorname{ker} P$ and Coker $P$ are finite dimensional.

As a particular case, let us look at the Laplacian on $S^{1}$ again.
Example 8.4.6 (Green operator on $\left.S^{1}\right)$. We recall the example 8.1.5. Let $\mathcal{H}=L_{2}\left(S^{1}\right)$ as before, and recall

$$
\mathcal{D}=\operatorname{dom} \Delta=\left\{f \in L_{2}\left(S^{1}\right): \sum_{n=-\infty}^{\infty} n^{4}|\widehat{f}(n)|^{2}<\infty\right\}
$$

We also recall that $\Delta e_{n}=n^{2} e_{n}$, (where $e_{n}=e^{i n \theta}$ ) so that $\Delta$ became an unbounded linear operator from $\mathcal{D} \rightarrow \mathcal{H}$. Then consider the space:

$$
\mathcal{H}_{2}:=H_{2}\left(S^{1}\right)=\left\{f \in \mathcal{H}: \sum_{n=-\infty}^{\infty}\left(1+n^{4}\right)|\widehat{f}(n)|^{2}<\infty\right\}
$$

Clearly, $\mathcal{H}_{2}=\mathcal{D}$ as a vector space. However, on $\mathcal{H}_{2}$ we have the Sobolev inner product $\langle-,-\rangle_{2}$, which by the Corollary 6.2.3, can also be defined as:

$$
\langle f, g\rangle_{2}:=\langle f, g\rangle_{0}+\langle\Delta f, \Delta g\rangle_{0}=\sum_{n=-\infty}^{\infty}\left(1+n^{4}\right) \widehat{f}(n) \overline{\widehat{g}(n)}
$$

which explains the notation $H_{2}\left(S^{1}\right)$ for the space above, and by earlier considerations makes it into a Hilbert Space. It is clear that $e_{n}=e^{i n \theta}$ continue to be orthogonal, but not orthonormal with respect to $\langle-,-\rangle_{2}$. Indeed, $\left\|e_{n}\right\|_{2}=\left(1+n^{4}\right)^{\frac{1}{2}}$.

Clearly, by definition, we have:

$$
\|\Delta f\|^{2}=\sum_{n=-\infty}^{\infty} n^{4}|\widehat{f}(n)|^{2} \leq\|f\|_{2}^{2}
$$

which makes $\Delta: \mathcal{H}_{2} \rightarrow \mathcal{H}$ a bounded operator, an element of $\mathcal{B}\left(\mathcal{H}_{2}, \mathcal{H}\right)$ (a fact we already know from Proposition 6.1.2) an element of $\mathcal{B}\left(\mathcal{H}_{2}, \mathcal{H}\right)$.

Similarly, for the Green operator $G$ introduced in 8.1.5,

$$
\|G f\|_{2}^{2}=\sum_{n=-\infty, n \neq 0}^{\infty} \frac{\left(1+n^{4}\right)}{n^{4}}|\widehat{f}(n)|^{2} \leq 2\|f\|^{2}
$$

so $G$ is also a bounded operator, and lies in $\mathcal{B}\left(\mathcal{H}, \mathcal{H}_{2}\right)$. The relations $I_{\mathcal{H}_{2}}-\pi_{0}=G \Delta$ and $I_{\mathcal{H}}-\pi_{0}=\Delta G$ found in 8.1.5 show that both $\Delta: \mathcal{H}_{2} \rightarrow \mathcal{H}$ and $G: \mathcal{H} \rightarrow \mathcal{H}_{2}$ are Fredholm operators. Note that ker $\Delta=\mathbb{C} e_{0}$, and ker $G=\mathbb{C} e_{0}$ as well.

Exercise 8.4.7. The Green operator can be written explicitly as a convolution with an $L_{2}$ function on $S^{1}$. Define the function $g \in L_{2}\left(S^{1}\right)=\mathcal{H}$ by the formula:

$$
g=\frac{1}{2 \pi} \sum_{n \neq 0, n \in \mathbb{Z}} \frac{1}{n^{2}} e_{n}
$$

where $e_{n}\left(e^{i t}\right)=e^{i n t}$ for $z=e^{i t} \in S^{1}$. Verify that:

$$
(G f)(z)=\int_{S^{1}} g\left(z w^{-1}\right) f(w) d w \text { for } z \in S^{1}
$$

where $w=e^{i s}$, and $d w:=d s$. Calculate the distribution $\Delta g$.

We might as well record a direct consequence of the last few sections in the following:
Proposition 8.4.8 (Green Operator for a Self-adjoint Elliptic Differential Operator). Let $M$ be a compact Riemannian manifold, with a smooth complex vector bundle $E$ on it. Let $d V(x)$ be the Riemannian volume form on $M$ and let $\langle-,-\rangle$ be a Hermitian inner product on $E$. Let $P: C^{\infty}(M, E) \rightarrow C^{\infty}(M, E)$ be an elliptic differential operator of order $d>0$. Assume that $P$ is formally self-adjoint, viz.

$$
(P f, g)=\int_{M}\langle P f(x), g(x)\rangle_{x} d V(x)=(f, P g) \text { for all } f, g \in C^{\infty}(M, E)
$$

Consider the bounded operator:

$$
P: H_{d}(M, E) \rightarrow H_{0}(M, E)=L_{2}(M, E)
$$

Then, for this last operator, we have:
(i): $\operatorname{dim}$ ker $P<\infty$, and this kernel is contained in $C^{\infty}(M, E)$, and in particular $H_{s}(M, E)$ for all $s \in \mathbb{R}$.
(ii): $\operatorname{Im} P \subset L_{2}(M, E)$ is closed, and Coker $P:=(\operatorname{Im} P)^{\perp}=\operatorname{ker} P$.
(iii): There exists a bounded self-adjoint operator called the Green Operator

$$
G: L_{2}(M, E) \rightarrow L_{2}(M, E)
$$

for $P$ which satisfies:
(a): $G \equiv 0$ on $\operatorname{ker} P \subset L_{2}(M, E)$, and $G=P^{-1}$ on $\operatorname{ker} P^{\perp}=\operatorname{Coker} P^{\perp}=\operatorname{Im} P \subset L_{2}(M, E) . G$ is a compact operator.
(b): $G\left(C^{\infty}(M, E)\right) \subset C^{\infty}(M, E)$, and $G P=P G$ on $C^{\infty}(M, E)$.
(c): $G: L_{2}(M, E) \rightarrow L_{2}(M, E)$ is a compact self adjoint operator. There is an orthonormal basis $\left\{e_{i}\right\}_{i=1}^{\infty}$ of $L_{2}(M, E)$ of eigensections of $G$, which satisfy

$$
G e_{i}=\mu_{i} e_{i} \quad \text { for all } i
$$

where $\mu_{i} \in \mathbb{R}$ for all $i$. 0 is the only cluster point of the set $\left\{\mu_{i}\right\}_{i=1}^{\infty}$, and $\lim _{i \rightarrow \infty} \mu_{i}=0$.
(d): The eigensections $\left\{e_{i}\right\}$ of (c) above are all smooth, and are also eigensections for $P$, satisfying:

$$
P e_{i}=\lambda_{i} e_{i} \text { for all } i
$$

where $\lambda_{i} \in \mathbb{R}$ is a discrete subset of $\mathbb{R}$, and $\lim _{i \rightarrow \infty}\left|\lambda_{i}\right|=\infty$

Proof: The operator:

$$
P: H_{d}(M, E) \rightarrow L_{2}(M, E)=H_{0}(M, E)
$$

is bounded by the Proposition 6.1.2. Its kernel ker $P$ is finite dimensional by the Example 8.4.5 above, where $P$ was found to be Fredholm, and (ii) of the Fredholm Theorem 8.3.3. That ker $P \subset C^{\infty}(M, E)$ is a consequence of the elliptic regularity theorem Proposition 6.2.4. Since $C^{\infty}(M, E) \subset H_{s}(M, E)$ for all $s$, it follows that ker $P \subset H_{s}(M, E)$ for all $s$. This proves (i).

That $\operatorname{Im} P$ is closed in $L_{2}(M, E)$ follows from (i) of the Fredholm Theorem 8.3.3, and Example 8.4.5. Since $P: H_{d} \rightarrow H_{0}$ is bounded, $P\left(C^{\infty}(M, E)\right)$ is dense in $\operatorname{Im} P$. Hence $f \in L_{2}(M, E)$ is orthogonal to $\operatorname{Im} P$ iff $(f, P g)=0$ for all $g \in H_{d}(M, E)$. By the formal self-adjointness of $P$, and the natural duality of $H_{d}$ and $H_{-d}$ in (iii) of 4.2.2, $(f, P g)=(P f, g)$ for $f \in H_{0}$ and $g \in H_{d}$. Thus we have $(P f, g)=0$ for all $g \in H_{d}(M, E)$. This is equivalent to $P f=0$. Thus Coker $P=(\operatorname{Im} P)^{\perp}=\operatorname{ker} P$, and (ii) follows.

By (ii), we have an $L_{2}$-orthogonal decomposition:

$$
L_{2}(M, E)=\operatorname{Im} P \oplus \operatorname{Coker} P=\operatorname{Im} P \oplus \operatorname{ker} P
$$

We now define $G$ by setting $G \equiv 0$ on $\operatorname{ker} P$, and $G$ to be equal $\theta$ which is the composite:

$$
\operatorname{Im} P \xrightarrow{P^{-1}}(\operatorname{ker} P)^{\perp} \hookrightarrow H_{d}(M, E) \hookrightarrow H_{0}(M, E)=L_{2}(M, E)
$$

By the Open Mapping Theorem, $P^{-1}: \operatorname{Im} P \rightarrow \operatorname{ker} P^{\perp}$ is a bounded operator, as is the inclusion $(\operatorname{ker} P)^{\perp} \rightarrow$ $H_{d}(M, E)$. The last inclusion $H_{d}(M, E) \rightarrow L_{2}(M, E)$ is a compact operator by Rellich's Lemma 4.2.2. Thus by the Proposition 8.2.1, the map $\theta$ is a compact operator. Since $G=\theta \circ \pi$, where $\pi: L_{2} \rightarrow \operatorname{Im} P$ is orthogonal projection onto the closed subspace $\operatorname{Im} P, G$ is also a compact operator by 8.2.1. This proves (a) of (iii).

Since $P$ is a differential operator, $P\left(C^{\infty}(M, E)\right) \subset C^{\infty}(M, E)$, and if $g \in C^{\infty}(M, E) \subset L_{2}(M, E)$, then its projection to the closed subspace $\operatorname{Im} P=(\operatorname{ker} P)^{\perp}$ is given by:

$$
\pi(g)=g-\sum_{i=1}^{k}\left(g, f_{i}\right) f_{i}
$$

where $\left\{f_{i}\right\}_{i=1}^{k}$ is an $L_{2}$-orthonormal basis for $\operatorname{ker} P$. By (i) above, all the $f_{i}$ are smooth, thus the scalar combination $\sum_{i}\left(g, f_{i}\right) f_{i}$ is smooth, and hence $\pi(g)$ above is smooth. On the other hand, for a smooth $g=P f$ in $\operatorname{Im} P$, it follows by elliptic regularity of Proposition 6.2 .4 that $f$ is also smooth. Thus the map $\theta$ above also maps smooth sections in $\operatorname{Im} P$ into smooth sections. Since $G=\theta \circ \pi$, it maps smooth sections to smooth sections. The fact that $G P=P G$ on smooth sections follows immediately from the definitions. This proves (b).

That $G$ is a compact operator was seen in (a). That it is self-adjoint follows from the definition $G=\theta \circ \pi$, and $\theta$ is the inverse of the formally self-adjoint $P$, and $C^{\infty}(M, E)$ is dense in $L_{2}(M, E)$. The statement about its eigenvalues and the orthonormal decomposition of $L_{2}(M, E)$ into eigenspaces of $G$ is the content of the Proposition 8.1.10. The eigenvalues are real since $G$ is self-adjoint. This proves (c).

To see (d), note that $G e_{i}=\mu_{i} e_{i}$, and $\mu_{i} \neq 0$ implies that $e_{i}$ are orthogonal to ker $P$, and hence so are $G e_{i}$, so that:

$$
\mu_{i} P e_{i}=P G e_{i}=P \theta e_{i}=P P^{-1} e_{i}=e_{i}
$$

so that $P e_{i}=\mu_{i}^{-1} e_{i}$ for all $\mu_{i} \neq 0$, and $e_{i}$ become eigensections of $P$, corresponding to the real non-zero eigenvalues $\lambda_{i}=\mu_{i}^{-1}$. Since $\left(P-\lambda_{i}\right) e_{i}=0$, and $P-\lambda_{i}$ is also elliptic of order $d$ (it has the same leading symbol as $P$ ), it follows that $e_{i} \in C^{\infty}(M, E)$ for all $i$ such that $\mu_{i} \neq 0$. For those $i$ 's which have $\mu_{i}=0$, we have $e_{i} \in \operatorname{ker} P$, and we already know those are smooth by (b). Hence $e_{i}$ are all smooth, and the rest of (d) follows from (c) above.

Actually, we can refine (iii) (d) of the previous proposition. To be precise, we have the following proposition.
Proposition 8.4.9. Let $M, E$ and $P$ be as above in Proposition 8.4.8. Then let us arrange the absolute values of the eigenvalues $\lambda_{i}$ of $P$ as in (iii) (d) of the previous proposition in non-decreasing order as:

$$
\left|\lambda_{1}\right| \leq\left|\lambda_{2}\right| \ldots \leq\left|\lambda_{k}\right| \leq \ldots
$$

Then there exists constants $C, \delta>0$ and $N \in \mathbb{N}$ such that $\left|\lambda_{n}\right| \geq C n^{\delta}$ for all $n \geq N$.

Proof: First we note that the eigenvalues of $P^{k}$ will be $\lambda_{n}^{k}$, and obtaining the assertion for $\lambda_{n}^{k}$ is sufficient to imply the same assertion for $\lambda_{n}$ (with $\delta$ replaced by $\delta / k$ ). So we may assume without loss of generality that $P$ is of degree $d>n / 2$ where $n=\operatorname{dim} M$.

Since we are assuming $d>n / 2$, by (iv) of the Proposition 4.2.2 (viz. the Sobolev embedding theorem), we have for $f \in C^{\infty}(M, E)$ the inequality:

$$
\|f\|_{\infty}=\sup _{x \in M}|f(x)| \leq C\|f\|_{d} \quad \text { for all } f \in C^{\infty}(M, E)
$$

and combining this with the Garding-Friedrichs inequality Proposition 6.2 .2 we have:

$$
\begin{equation*}
\|f\|_{\infty} \leq C\left(\|P f\|_{0}+\|f\|_{0}\right) \quad \text { for all } \quad f \in C^{\infty}(M, E) \tag{21}
\end{equation*}
$$

We note that by elliptic regularity, all the eigensections $\phi_{k}$ of $P$ are smooth sections. We assume they are orthonormal with respect to $L_{2}$-norm $\left\|\|_{0}\right.$. Define:

$$
F(a):=\operatorname{span}_{\mathbb{C}}\left\{\phi_{k}: P \phi_{k}=\lambda_{k} \phi_{k} \text { and }\left|\lambda_{k}\right| \leq a\right\}
$$

Let $m=\operatorname{dim} F(a)$. This dimension is finite by the fact that $\lambda_{k}^{-1}$ have no cluster point except 0 from (iii) (c) of 8.4 .8 . We will make an estimate for $m$ in terms of $a$, which will imply our assertion.

Note that for $f=\sum_{j=1}^{m} \alpha_{j} \phi_{j} \in F(a)$, we have $P f=\sum_{j=1}^{m} \alpha_{j} \lambda_{j} \phi_{j}$, which shows that

$$
\|P f\|_{0} \leq a\|f\|_{0} \quad \text { for all } f \in F(a)
$$

Plugging this into the inequality (21) above, we have for all choices of complex constants $\alpha_{j}$, the inequality:

$$
\begin{equation*}
\left\|\sum_{j=1}^{m} \alpha_{j} \phi_{j}\right\|_{\infty} \leq C(1+a)\left\|\sum_{j=1}^{m} \alpha_{j} \phi_{j}\right\|_{0} \tag{22}
\end{equation*}
$$

In a local frame $\left\{e_{i}(x)\right\}_{i=1}^{k}$ orthonormal frame of $E$ over $U \subset M$, where $k=\operatorname{rk}_{\mathbb{C}} E$, write:

$$
\phi_{j}(x)=\sum_{i=1}^{k} \phi_{j}^{i}(x) e_{i}(x)
$$

So that for any choice of constants $\alpha_{j}$, we have for $x \in U$ that

$$
\sum_{j=1}^{m} \alpha_{j} \phi_{j}(x)=\sum_{i=1}^{k}\left(\sum_{j} \alpha_{j} \phi_{j}^{i}(x)\right) e_{i}(x)
$$

so that for any choice of constants $\alpha_{j} \in \mathbb{C}$, the inequality (22) implies:

$$
\left|\sum_{j} \alpha_{j} \phi_{j}^{i}(x)\right| \leq\left\|\sum_{j=1}^{m} \alpha_{j} \phi_{j}\right\|_{\infty} \leq C(1+a)\left(\sum_{j}\left|\alpha_{j}\right|^{2}\right)^{\frac{1}{2}} \quad \text { for each } i=1,2, . . k
$$

For $x \in U$, choose $\alpha_{j}=\bar{\phi}_{j}^{i}(x)$. Then the last inequality reads:

$$
\sum_{j}\left|\phi_{j}^{i}(x)\right|^{2} \leq C(1+a)\left(\sum_{j}\left|\phi_{j}^{i}(x)\right|^{2}\right)^{\frac{1}{2}}
$$

that is:

$$
\left(\sum_{j}\left|\phi_{j}^{i}(x)\right|^{2}\right)^{\frac{1}{2}} \leq C(1+a)
$$

Squaring both sides and summing over $i=1, . ., k$, we have:

$$
\sum_{j=1}^{m}\left\|\phi_{j}(x)\right\|_{x}^{2}=\sum_{j}\left(\sum_{i}\left|\phi_{j}^{i}(x)\right|^{2}\right)=\sum_{i}\left(\sum_{j}\left|\phi_{j}^{i}(x)\right|^{2}\right) \leq k C^{2}(1+a)^{2}=C^{2}(1+a)^{2}
$$

where $C$ is a generic constant independent of $a$. This inequality is true for each $x \in M$, so we may integrate both sides over all of $M$ to obtain

$$
m=\int_{M} \sum_{j=1}^{m}\left(\left\|\phi_{j}(x)\right\|_{x}^{2}\right) d V(x) \leq C^{2}(1+a)^{2}
$$

which shows that $\frac{1}{C} m^{\frac{1}{2}}-1 \leq a$. Since $\left|\lambda_{j}\right| \leq a$ for $j=1,2, . ., m$, we can take $a=\max _{j=1}^{m}\left|\lambda_{j}\right|=\left|\lambda_{m}\right|$, so that we have:

$$
\left|\lambda_{m}\right| \geq C m^{\frac{1}{2}} \text { for } m \geq N
$$

and the proposition follows.

### 8.5. The Fredholm Index.

Definition 8.5.1 (Fredholm Index). Let $T: \mathcal{H} \rightarrow \mathcal{H}$ be a Fredholm operator. The Fredholm index of $T$ is defined by:

$$
\operatorname{ind} T=\operatorname{dim} \operatorname{ker} T-\operatorname{dim} \operatorname{Coker} T
$$

It makes sense, and is an integer, because of proposition 8.3.3. Similarly for $T \in \mathcal{B}\left(\mathcal{H}_{1}, \mathcal{H}_{2}\right)$, one again defines the index ind $T$ by the same formula.

Example 8.5.2 (Index of examples above). We can easily compute the indices of the various examples of Fredholm operators listed above. For an invertible operator $T: \mathcal{H}_{1} \rightarrow \mathcal{H}_{2}$, the index is clearly 0 . For a linear map $T: V \rightarrow W$ of finite dimensional vector spaces, the index is easily seen to be $\operatorname{dim} V-\operatorname{dim} W$ by elementary linear algebra. (Thus in the finite dimensional situation, the index depends only on the domain $V$ and range $W$, and is not a very interesting invariant of $T$ ). For the unilateral right (resp. left) $k$-shift, the index is $(-k)$ (resp. $k$ ).

For an elliptic self-adjoint differential operator $P: C^{\infty}(M, E) \rightarrow C^{\infty}(M, E)$ of order $d, M$ a compact Riemannian manifold, we have that $P: H_{d}(M, E) \rightarrow L_{2}(M, E)$ is Fredholm, by 8.4.5. Also by (ii) of 8.4.8, ker $P=\operatorname{Coker} P$, and hence ind $P=0$.

Proposition 8.5.3. Let $T \in \mathcal{B}\left(\mathcal{H}_{1}, \mathcal{H}_{2}\right)$ and $S \in \mathcal{B}\left(\mathcal{H}_{2}, \mathcal{H}_{3}\right)$ be Fredholm operators. Then $T S \in \mathcal{B}\left(\mathcal{H}_{1}, \mathcal{H}_{3}\right)$ is Fredholm, and

$$
\operatorname{ind} S T=\operatorname{ind} S+\operatorname{ind} T
$$

In particular, if $\mathcal{H}_{1}=\mathcal{H}_{2}=\mathcal{H}$, the index is a group homomorphism from the group of units (=set of invertible elements) in the Calkin algebra $\mathcal{C}(\mathcal{H})$ to $\mathbb{Z}$.

Proof: In the sequel, we will denote the kernel of a linear operator $T$ by $N(T)$. The fact that $S T$ is Fredholm follows from the fact that post and precomposing compact operators with bounded operators again yields compact operators (see example 8.4.2).) Note that for any linear operator $T$, we have the following identity for a closed subspace $W$ :

$$
\begin{equation*}
T^{-1}\left(W^{\perp}\right)=\left(T^{*}(W)\right)^{\perp} \tag{23}
\end{equation*}
$$

where the left side is the inverse image of $W^{\perp}$. From this (by taking $W^{\perp}=\{0\}$ ), one sees that Coker $T^{*}=$ $\operatorname{Im} T^{* \perp}=N(T)$, and Coker $T=(\operatorname{Im} T)^{\perp}=N\left(T^{*}\right)$. Now one may do orthogonal decompositions of $\mathcal{H}_{1}, \mathcal{H}_{2}$ and $\mathcal{H}_{3}$ as follows:

$$
\begin{aligned}
\mathcal{H}_{1} & =N(T) \oplus F \\
\mathcal{H}_{2} & =N\left(T^{*}\right) \oplus G_{1}=N(S) \oplus G_{2} \\
\mathcal{H}_{3} & =N\left(S^{*}\right) \oplus H
\end{aligned}
$$

where $T: F \rightarrow G_{1}$ and $S: G_{2} \rightarrow H$ are isomorphisms.
The kernel of $S T$ is given by (using the identity (23), and noting that $T_{\mid F}: F \rightarrow \operatorname{Im} T$ is an isomorphism) above):

$$
\begin{aligned}
N(S T) & =T^{-1} S^{-1}(0)=T^{-1}(N(S)) \\
& =N(T)+T_{\mid F}^{-1}(N(S) \cap \operatorname{Im} T) \\
& =N(T)+T_{\mid F}^{-1}\left(N(S) \ominus\left(N(S) \cap N\left(T^{*}\right)\right)\right.
\end{aligned}
$$

so that

$$
\operatorname{dim} N(S T)=\operatorname{dim} N(T)+\operatorname{dim} N(S)-\operatorname{dim}\left(N(S) \cap N\left(T^{*}\right)\right)
$$

Similarly,

$$
\begin{aligned}
\operatorname{dim} N\left((S T)^{*}\right) & =\operatorname{dim} N\left(T^{*} S^{*}\right)=\operatorname{dim} N\left(S^{*}\right)+\operatorname{dim} N\left(T^{*}\right)-\operatorname{dim}\left(N\left(T^{*}\right) \cap N\left(S^{* *}\right)\right) \\
& =\operatorname{dim} N\left(S^{*}\right)+\operatorname{dim} N\left(T^{*}\right)-\operatorname{dim}\left(N\left(T^{*}\right) \cap N(S)\right)
\end{aligned}
$$

Combining the two identities above, we get:
$\operatorname{ind} S T=\operatorname{dim} N(S T)-\operatorname{dim} N\left((S T)^{*}\right)=\operatorname{dim} N(T)+\operatorname{dim} N(S)-\operatorname{dim} N\left(S^{*}\right)-\operatorname{dim} N\left(T^{*}\right)=\operatorname{ind} T+\operatorname{ind} S$ proving the proposition.
8.6. Path components and Fredholm index. We have the following crucial fact about Fredholm operators.

Proposition 8.6.1 (Invariance of index). Let $t \mapsto T_{t}$ be a continuous map of an interval $I$ to $\mathcal{B}(\mathcal{H})$, with $T_{t}$ a Fredholm operator for each $t \in I$. Then:

$$
\operatorname{ind} T_{t}=\operatorname{ind} T_{s} \quad \text { for all } t, s \in I
$$

Thus the index remains a constant integer on each path component of the set of units (=invertible elements) in the Calkin Algebra.

Proof: We will show that the index is locally constant on $I$, and that will make $t \mapsto \operatorname{ind} T_{t}$ a continuous, and hence constant map. For a point $t \in I$, denote the kernel of $T_{t}$ by $K_{t}$. Since $T_{t}$ is Fredholm for each $t$, $K_{t}$ is finite dimensional for all $t$. Let $V_{t}:=K_{t}^{\perp}$, and $W_{t}:=\operatorname{Im} T_{t}$. By Fredholmness of $T_{t}, W_{t}^{\perp}$ is also finite dimensional for each $t$.

Fix any $a \in I$. We claim that for a small enough $\epsilon>0$, and for $|t-a|<\epsilon$, the index ind $T_{t}=\operatorname{ind} T_{a}$. For simplicity of notation, denote $K_{a}$ by $K, V_{a}$ by $V$ and $W_{a}$ by $W$. Let $\pi: \mathcal{H} \rightarrow W$ denote the orthogonal projection, and $i: V \rightarrow \mathcal{H}$ denote the inclusion. Then $\pi T_{a} i: V \rightarrow W$ is an isomorphism, by definition. Since isomorphisms from $V$ to $W$ form an open set in $\mathcal{B}(V, W)$, it follows that there is an $\epsilon>0$ such that $\pi T_{t} i: V \rightarrow W$ is an isomorphism for all $t$ such that $|t-a|<\epsilon$.

Thus, for $|t-a|<\epsilon$, the index

$$
\operatorname{ind}\left(\pi T_{t} i\right)=0
$$

By the proposition 8.5.3, and the facts that ker $\pi=W^{\perp}=\operatorname{Coker} T_{a}$, Coker $\pi=0$, ker $i=0$, Coker $i=V^{\perp}=$ $\operatorname{ker} T_{a}$, it follows that

$$
\operatorname{ind}\left(\pi T_{t} i\right)=\operatorname{ind} \pi+\operatorname{ind} T_{t}+\operatorname{ind} i=\operatorname{dim} W^{\perp}+\operatorname{ind} T_{t}+\left(-\operatorname{dim} V^{\perp}\right)=\operatorname{ind} T_{t}-\operatorname{ind} T_{a}
$$

Thus ind $T_{t}=\operatorname{ind} T_{a}$ for $|t-a|<\epsilon$, and the proposition is proved.

## 9. Elliptic Complexes on Compact Riemannian manifolds

9.1. The de Rham complex. Let $M$ be a smooth connected oriented (i.e.the Jacobian of each coordinate change in the atlas being used has positive determinant) Riemannian manifold of dimension $n$, with Riemannian metric $g$. We recall that the volume $n$-form $d V$ associated to this Riemannian metric, is given in a local coordinate system $(\phi, U)$ of an oriented atlas by:

$$
d V_{g}(x)=\sqrt{\operatorname{det} g_{i j}(x)} d x_{1} \wedge d x_{2} \wedge . . \wedge d x_{n}
$$

with coordinate functions $x_{i}$ being the components of $\phi$ on the open set $U$. The expression above is independent of the coordinate system chosen, by the tranformation properties of the coordinate changes on the overlaps $U_{i} \cap U_{j}$, and the orientability of the atlas $\left\{\left(\phi_{i}, U_{i}\right)\right\}$. We will usually write $d V$ instead of $d V_{g}$.

One also has the complex vector space of smooth complex-valued differential $p$-forms on $M$, which is denoted by $\bigwedge^{p}(M)$. Let $\omega$ be a differential $p$-form is given in a coordinate chart $(\phi, U)$ by the local expression:
where $I=\left(i_{1}<i_{2}<. .<i_{p}\right), \quad 1 \leq i_{j} \leq n$ denotes a multi-index of length $p$, and $\omega_{I}$ are all smooth functions on the open set $U$. Note that $\bigwedge^{0}(M)$ is just the vector space of smooth functions on $M$, and when $M$ is oreinted, there is an isomorphism $f \mapsto f d V_{g}$ of $\bigwedge^{0}(M)$ with $\bigwedge^{n}(M)$ upon choosing a Riemannian volume element.

Then one can define the exterior derivative operator

$$
d: \bigwedge^{p}(M) \rightarrow \bigwedge^{p+1}(M)
$$

by $d \omega:=\sum_{I} d \omega_{I} \wedge d x_{I}$, where $d \omega_{I}=\sum_{j} \frac{\partial \omega_{I}}{\partial x_{j}} d x_{j}$ is a 1-form. One easily checks that this definition of $d$ is global, and does not depend on the choice of local coordinate charts. (In the case of $M=\mathbb{R}^{3}$, the exterior
derivatives on $\bigwedge^{0}\left(\mathbb{R}^{3}\right), \bigwedge^{1}\left(\mathbb{R}^{3}\right)$ and $\bigwedge^{2}\left(\mathbb{R}^{3}\right)$ lead to the familiar grad, curl and divergence operators.) It is well known that $d \circ d=0$, and so we have a cochain complex of complex vector spaces:

$$
\bigwedge_{0}^{0}(M) \xrightarrow{d} \bigwedge^{1}(M) \xrightarrow{d} \ldots \bigwedge^{p}(M) \xrightarrow{d} \bigwedge^{p+1}(M) \ldots \xrightarrow{d} \bigwedge^{n}(M)
$$

which is called the de Rham complex of $M$. We also have the skew-derivation formula for the exterior derivative:

$$
d(\omega \wedge \tau)=d \omega \wedge \tau+(-1)^{p} \omega \wedge d \tau
$$

where $\omega \in \bigwedge^{p}(M), \tau \in \bigwedge^{q}(M)$.
The de Rham complex contains much of the algebraic topology of $M$, even though its definition is purely analytical. For example, we can define the $i$-th de Rham cohomology of $M$ as the quotient:

$$
H^{i}(M, \mathbb{C}):=\frac{\operatorname{ker} d: \bigwedge^{i}(M) \rightarrow \bigwedge^{i+1}(M)}{\operatorname{Im} d: \bigwedge^{i-1}(M) \rightarrow \bigwedge^{i}(M)}
$$

It turns out by de Rham's theorem (to be stated below) that the dimension $\operatorname{dim} H^{i}(M, \mathbb{C})$ is the $i$-th Betti number of $M$, and the alternating sum

$$
\sum_{i=0}^{n}(-1)^{i} \operatorname{dim} H^{i}(M, \mathbb{C})
$$

is the topological Euler characteristic $\chi(M)$ of $M$.
Now one brings in the Riemannian metric to introduce pointwise and global hermitian inner products on differential forms. The Riemannian metric $g$ defines inner products for all real tangent vectors, and gives an identification of the real cotangent space with the tangent space by the identification $X \mapsto X^{*}$ where $X^{*}$ is defined by the formula $X^{*}(Y)=g(X, Y)$ for all tangent vectors $Y$ to $M$ at $x$. Declaring the vector space isomorphism above to be an isometry puts a real positive definite inner product $\langle-,-\rangle$ on real cotangent vectors. More explicitly, in a coordinate chart $(\phi, U)$ with coordinates $x_{i}$ we have:

$$
\left\langle d x_{i}, d x_{j}\right\rangle=g^{i j}
$$

where $g^{i j}$ is the inverse of the $n \times n$ positive definite matrix $\left[g_{k l}:=g\left(\frac{\partial}{\partial x_{k}}, \frac{\partial}{\partial x_{l}}\right)\right]$. Each of these $g_{i j}$ 's is a smooth function of $x \in U$. Now we get inner products on all real $p$-covectors by the formula:

$$
\left\langle d x_{i_{1}} \wedge d x_{i_{2}} \ldots \wedge d x_{i_{p}}, d x_{j_{1}} \wedge d x_{j_{2}} \ldots \wedge d x_{j_{p}}\right\rangle=\operatorname{det}\left[g^{i_{l} j_{m}}\right]
$$

Thus we can talk of $\langle\omega(x), \tau(x)\rangle$ for two real $p$-forms $\omega, \tau \in \bigwedge^{p}(M, \mathbb{R})$. We do the canonical Hermitian extension of this real inner product on $\bigwedge^{p}(M, \mathbb{R})$ to a Hermitian inner product on its complexification $\bigwedge^{p}(M)=$ $\bigwedge^{p}(M, \mathbb{R}) \otimes \mathbb{C}$. We continue to denote it by $\langle-,-\rangle$. By definition, for $\omega, \tau \in \bigwedge^{p}(M)$, the pointwise inner product $\langle\omega(x), \tau(x)\rangle$ is a smooth function of $x \in M$. We can then define the global inner product:

$$
(\omega, \tau):=\int_{M}\langle\omega(x), \tau(x)\rangle d V(x)
$$

of the smooth $p$-forms $\omega, \tau \in \bigwedge^{p}(M)$, and if $M$ is compact, $(\omega, \tau)$ will be finite for all $\omega, \tau \in \bigwedge^{p}(M)$.
The Hodge star operator is an operator:

$$
*: \bigwedge^{p}(M) \rightarrow \bigwedge^{n-p}(M)
$$

which is the unique operator obeying the identity:

$$
\omega \wedge(* \tau)=\langle\omega, \tau\rangle d V=\tau \wedge * \omega
$$

for $\omega, \tau \in \bigwedge^{p}(M)$. That is, $\langle-,-\rangle$ being a non-degenerate pairing gives an identification of $\bigwedge^{p}(M)$ with its dual vector space $\bigwedge^{p *}(M)$, and $\wedge$ being a non-degenerate pairing of $\bigwedge^{p}(M)$ with $\bigwedge^{n-p}(M)$ provides an identification of $\bigwedge^{n-p}(M)$ with $\bigwedge^{p *}(M)$, so the Hodge $*$-operator is the resulting identification of $\bigwedge^{p}(M)$ with $\bigwedge^{n-p}(M)$. Using the fact that $\omega \wedge * \tau=(-1)^{p(n-p)}(* \tau) \wedge \omega$ and that $\langle\omega, \tau\rangle d V=\langle\tau, \omega\rangle d V$, it easily follows from the definition of $*$ above that

$$
* \circ *=(-1)^{p(n-p)} \quad \text { on } \bigwedge^{p}(M)
$$

As expected, $*: \bigwedge^{0}(M) \rightarrow \bigwedge^{n}(M)$ is the isomorphism $f \mapsto f d V$ discussed earlier.
Using the Hodge $*$-operator, one can define the differential operator:

$$
\left.\begin{array}{rl}
\delta: \bigwedge_{\bigwedge}^{p}(M) & \rightarrow \bigwedge_{1}^{p-1}(M) \\
\omega & \mapsto
\end{array}\right)(-1)^{n p+n-1} * d(* \omega) \text {. }
$$

From this definition it follows that $* \delta \omega=(-1)^{n p+n-1} * * d * \omega=(-1)^{n p+n-1+(n-p+1)(p-1)} d * \omega=(-1)^{p} d(* \omega)$.

We note that if $\omega$ is a $p$-form and $\tau$ a $(p+1)$-form, then:

$$
d(\omega \wedge * \tau)=d \omega \wedge * \tau+(-1)^{p} \omega \wedge d(* \tau)=d \omega \wedge * \tau-\omega \wedge(* \delta \tau)
$$

Now $d(\omega \wedge * \tau)$ is an $n$-form on $M$, and if $M$ is compact, or if one of $\tau, \omega$ are of compact support, then by Stokes theorem we have (since $\partial M=\phi$ ) that:

$$
\begin{align*}
(d \omega, \tau) & =\int_{M} d \omega \wedge * \tau=\int_{M} \omega \wedge * \delta \tau+\int_{M} d(\omega \wedge * \tau)=\int_{M} \omega \wedge * \delta \tau+\int_{\partial M}(\omega \wedge * \tau) \\
& =(\omega, \delta \tau)+0=(\omega, \delta \tau) \tag{24}
\end{align*}
$$

That is, the operators $d$ and $\delta$ are formal adjoints to each other on the spaces of smooth compactly supported forms, with respect to the global inner product (, ) defined above.

### 9.2. The Laplacian on differential forms.

Definition 9.2.1. The Laplace-Beltrami operator, or Laplacian on $\bigwedge^{p}(M)$ is defined as:

$$
\begin{aligned}
\Delta: \bigwedge^{p}(M) & \rightarrow \bigwedge^{p}(M) \\
\omega & \mapsto(d \delta+\delta d) \omega
\end{aligned}
$$

Since $d$ and $\delta$ are both first-order differential operators, $\Delta$ is a second order differential operator.

One can also write down expressions for $\Delta$ in local coordinates, which are messy. For $\Delta: \bigwedge^{0}(M) \rightarrow \bigwedge^{0}(M)$, the expression is:

$$
\Delta f=-\frac{1}{\sqrt{g}} \sum_{i, j=1}^{n} \partial_{i}\left(\sqrt{g} g^{i j} \partial_{j} f\right) \quad \text { for } f \in \bigwedge^{0}(M)
$$

where $\sqrt{g}:=\sqrt{\operatorname{det} g_{i j}}$ and $\partial_{j}:=\frac{\partial}{\partial x_{j}}$.

Remark 9.2.2 (Formal self-adjointness and positivity of $\Delta$ ). By (24) above, we also have for $M$ compact, or one of $\omega, \tau \in \bigwedge^{p}(M)$ of compact support that:

$$
(\Delta \omega, \tau)=((d \delta+\delta d) \omega, \tau)=(\delta \omega, \delta \tau)+(d \omega, d \tau)=(\omega, \Delta \tau)
$$

that is, $\Delta$ is formally self-adjoint with respect to the global inner product $(-,-)$ on $\bigwedge^{p}(M)$.
Further, by the above, if $\omega$ is of compact support, or $M$ is compact,

$$
(\Delta \omega, \omega)=(d \omega, d \omega)+(\delta \omega, \delta \omega)
$$

Hence for $M$ compact, $(\Delta \omega, \omega) \geq 0$ for all $\omega \in \bigwedge^{i}(M)$, and $\Delta \omega=0$ for $\omega \in \bigwedge^{i}(M)$ if and only if $d \omega$ and $\delta \omega=0$.

Instead of proving ellipticity of the Laplace operator separately, we will set up the general notion of an elliptic complex, and the Laplacian above will follow as a special case.
9.3. Elliptic operators on compact manifolds. In the sequel, $T M$ will always denote the complexified tangent bundle $T_{\mathbb{C}}(M):=T_{\mathbb{R}}(M) \otimes_{\mathbb{R}} \mathbb{C}$. Likewise, for the cotangent bundle $T^{*} M:=T_{\mathbb{C}}^{*} M=\operatorname{hom}\left(T_{\mathbb{R}} M, \mathbb{C}\right)$.

Definition 9.3.1 (Algebra of differential operators on $M$ ). Define the space $\mathcal{D}^{0}(M)$ of linear differential operators of order 0 to be $C^{\infty}(M)$.

Let $\chi(M)$ denote the space of vector fields on $M$. That is, $\chi(M)$ is the space $C^{\infty}(M, T M)$ of smooth sections of $T M$ on $M$. Note that $\chi(M)$ has a natural left-module structure over the ring $C^{\infty}(M)$.

Define the space $\mathcal{D}^{1}(M)$ of linear differential operators of order 1 by

$$
\mathcal{D}^{1}(M):=C^{\infty}(M) \oplus \chi(M)
$$

This $\mathbb{C}$-vector space inherits the left $C^{\infty}$-module structure from both its summands. In addtion, it also the structure of a right $C^{\infty}(M)$ module, defined by:

$$
(P g)=g P+[P, g]=g P+[\alpha X, g]=g P+\alpha X(g) \text { for } P=\beta+\alpha X, \quad X \in \chi(M), \alpha, \beta \in C^{\infty}(M)
$$

This formula arises from the fact that the vector field $X$ is a derivation on $C^{\infty}(M)$, or more simply because $P g$ is naturally defined by the formula:

$$
P g(f)=P(g f) \text { for all } f \in C^{\infty}(M)
$$

Note that the commutators satisfy:
(i): $\left[\mathcal{D}^{1}(M), \mathcal{D}^{1}(M)\right] \subset \mathcal{D}^{1}(M)$.
(ii): $\left[\mathcal{D}^{0}(M), \mathcal{D}^{1}(M)\right]=\left[\mathcal{D}^{1}(M), \mathcal{D}^{0}(M)\right] \subset \mathcal{D}^{0}(M)$

The $k-t h$ tensor power of $\mathcal{D}^{1}(M)$ is defined as

$$
\mathcal{T}^{k}:=\mathcal{D}^{1}(M) \otimes_{C^{\infty}(M)} \mathcal{D}^{1}(M) \otimes \ldots \otimes_{C^{\infty}(M)} \mathcal{D}^{1}(M) \quad(k \text { times })
$$

uses the right $C^{\infty}(M)$-module structure of the $i$-th factor and the left $C^{\infty}(M)$-module structure of the $(i+1)$-th factor. Thus it has a natural left and right $C^{\infty}(M)$-module structure. Note that

$$
\mathcal{T}^{k}=\otimes^{k} \mathcal{D}^{1}(M)=\left(\otimes^{k} \mathcal{D}^{1}(M)\right) \otimes C^{\infty}(M) \subset\left(\otimes^{k} \mathcal{D}^{1}(M)\right) \otimes \mathcal{D}^{1}(M)=\mathcal{T}^{k+1}
$$

so that we can define:

$$
\mathcal{T}:=\cup_{k=0}^{\infty} \mathcal{T}^{k}=\cup_{k=0}^{\infty}\left(\otimes^{k} \mathcal{D}^{1}(M)\right)
$$

as an associative, non-commutative $C^{\infty}(M)$ algebra, fitered by $\mathcal{T}^{k}$. Let $\mathcal{I}$ be the left-ideal in $\mathcal{T}$ generated by all elements of the form

$$
P_{1} \otimes P_{2}-P_{2} \otimes P_{1}-\left[P_{1}, P_{2}\right] \otimes 1, \quad P_{1}, P_{2} \in \mathcal{D}^{1}(M)
$$

A simple calculation shows that for $g \in C^{\infty}(M), P_{1}, P_{2} \in \mathcal{D}^{1}$, we have:

$$
\begin{aligned}
\left(P_{1} \otimes P_{2}-P_{2} \otimes P_{1}-\left[P_{1}, P_{2}\right] \otimes 1\right) g & =g\left(P_{1} \otimes P_{2}-P_{2} \otimes P_{1}-\left[P_{1}, P_{2}\right] \otimes 1\right) \\
& +\left(\left[P_{1}, g\right] \otimes P_{2}-P_{2} \otimes\left[P_{1}, g\right]-\left[\left[P_{1}, g\right], P_{2}\right] \otimes 1\right) \\
& +\left(P_{1} \otimes\left[P_{2}, g\right]-\left[P_{2}, g\right] \otimes P_{1}-\left[P_{1},\left[P_{2}, g\right]\right] \otimes 1\right)
\end{aligned}
$$

(where one uses the Jacobi identity $\left(\left[\left[P_{1}, P_{2}\right], g\right]+\left[\left[P_{2}, g\right], P_{1}\right]+\left[\left[g, P_{1}\right], P_{2}\right]=0\right)$. Thus the left $C^{\infty}(M)$-ideal generated by the elements $P_{1} \otimes P_{2}-P_{2} \otimes P_{1}-\left[P_{1}, P_{2}\right] \otimes 1$ automatically becomes a right $C^{\infty}(M)$-ideal as well. Now we can go modulo this ideal $\mathcal{I}$.

Hence we define the algebra of differential operators on $M$ to be the associative algebra:

$$
\mathcal{D}^{\infty}(M)=\mathcal{T} / \mathcal{I}
$$

The image of $D^{d}(M)$ of $\mathcal{T}^{d}$ is the left $C^{\infty}(M)$-module of linear differential operators of order $d$. Since $\mathcal{T}^{d} \subset$ $\mathcal{T}^{d+1}$, we have $\mathcal{D}^{d}(M) \subset \mathcal{D}^{d+1}(M)$ for all $d$, and $\mathcal{D}^{0}(M)=C^{\infty}(M) . \mathcal{D}^{\infty}$ is a non-commutative, associative algebra over $\mathcal{C}^{\infty}(M)$, filtered by $\mathcal{D}^{d}(M)$. From the corresponding property of $\mathcal{T}^{k}$, , it follows that:

$$
\mathcal{D}^{i}(M) \cdot \mathcal{D}^{j}(M) \subset \mathcal{D}^{i+j}(M)
$$

Finally, if $E$ and $F$ are two smooth complex vector bundles over $M$, the space of smooth sections of the bundle $\operatorname{hom}(E, F)$, namely $C^{\infty}(M, \operatorname{hom}(E, F))$, is a right (=same as left) $C^{\infty}(M)$-module in a natural way. Hence we may define the $C^{\infty}(M)$-modules:
$\mathcal{D}^{d}(M ; E, F):=C^{\infty}(M, \operatorname{hom}(E, F)) \otimes_{C^{\infty}(M)} \mathcal{D}^{d}(M), \quad \mathcal{D}^{\infty}(M ; E, F):=C^{\infty}(M, \operatorname{hom}(E, F)) \otimes_{C^{\infty}(M)} \mathcal{D}^{\infty}(M)$
Note that these left $C^{\infty}(M)$-modules are algebras if $E=F$.

Exercise 9.3.2. Verify (by using local coordinates) that $P: C^{\infty}(M, E) \rightarrow C^{\infty}(M, F)$ is a linear differential operator of order $d$ in the sense of 6.1 .1 iff it is an element of $\mathcal{D}^{d}(M ; E, F)$.

We will denote $\mathcal{D}^{d}(M ; E, F)$ simply as $\mathcal{D}^{d}(E, F)$, and sometimes even $\mathcal{D}^{d}$ when no confusion is likely, for notational convenience.

Lemma 9.3.3 (Leading symbols again). Let $\pi:\left(T^{*} M\right) \rightarrow M$ denote the natural projection, where $T^{*} M$ is the real cotangent bundle of $M$. Note that there is the scaling map $T^{*} M \rightarrow T^{*} M$ given by $\xi \mapsto t \xi$, which preserves each fibre $T_{x}^{*} M$. Define the space of symbols of order $d$ by:

$$
\operatorname{Sym}^{d}(E, F):=\left\{\sigma \in C^{\infty}\left(T^{*} M, \pi^{*} \operatorname{hom}(E, F)\right): \sigma(t \xi)=t^{d} \sigma(\xi), \quad \xi \in T^{*}(M)\right\}
$$

(In other words, those smooth sections $\sigma$ which are homogenous polynomials of degree $d$ in the fibre variables). When $E=F=M \times \mathbb{C}$ the trivial line bundle, we denote $\operatorname{Sym}^{d}(E, F)$ simply by $\operatorname{Sym}^{d}(M)$. We have the following facts:
(i): The associated graded module to the filtered $C^{\infty}(M)$-module $\mathcal{D}^{\infty}(E, F)$ is the algebra $\operatorname{Sym}^{\infty}(E, F):=$ $\oplus_{d=0}^{\infty} \operatorname{Sym}^{d}(E, F)$. The natural quotient map of $C^{\infty}(M)$-modules:

$$
\begin{aligned}
\mathcal{D}^{d}(E, F) & \rightarrow \operatorname{Sym}^{d}(E, F) \\
P & \mapsto \sigma_{L}(P)
\end{aligned}
$$

is called the leading symbol map. When $E=F, \operatorname{Sym}^{\infty}(E, E)$ is a commutative algebra, and the map $\sigma_{L}$ is an algebra homomorphism $\mathcal{D}^{\infty}(E, E) \rightarrow \operatorname{Sym}^{\infty}(E, E)$. In a local coordinate system, $\sigma_{L}(P)$ as defined above coincides with the leading symbol defined earlier in Definition 5.5.4.
(ii): If $\xi \in T_{x}^{*}(M)$ is any cotangent vector, and $f$ is any smooth function satisfying $d f(x)=\xi$, then the leading symbol can be computed from the formula:

$$
\sigma_{L}(P)(\xi)=\lim _{t \rightarrow \infty} t^{-d}\left(e^{-i t f} P e^{i t f}\right)(x)
$$

(iii): $P$ is elliptic iff $\sigma_{L}(P)(\omega)$ is a bundle isomorphism at all points of $T^{*} M \backslash 0_{M}$, where $0_{M}$ denotes the zero section of $T^{*} M$.

Proof: First we note that

$$
\mathcal{D}^{d}(E, F)=C^{\infty}(M ; \operatorname{hom}(E, F)) \otimes_{C^{\infty}(M)} \mathcal{D}^{d}(M)
$$

and similarly

$$
\operatorname{Sym}^{d}(E, F)=C^{\infty}(M ; \operatorname{hom}(E, F)) \otimes_{C^{\infty}(M)} \operatorname{Sym}^{d}(M)
$$

hence we need prove all the assertions for the case of $E=F=M \times \mathbb{C}$, the trivial bundle, and then left-tensor everything with $C^{\infty}(M, \operatorname{hom}(E, F))$ to get it for general $E$ and $F$. The second simplification one can make is to reduce it to $M=\mathbb{R}^{n}$. This is done by first covering $M$ with charts $U_{i}$ with each $U_{i}$ diffeomorphic to $\mathbb{R}^{n}$. In fact, for any $U \subset M$ open, we can define the left and right $C^{\infty}(U)$ module $\mathcal{D}^{d}(U)$ by the definition above (applied to $M=U$ ). Indeed, for $V \supset U$ any two open subsets, there are the natural restriction maps $\chi(V) \rightarrow \chi(U)$ which preserves commutators of vector fields, and also the restriction map $C^{\infty}(V) \rightarrow C^{\infty}(U)$. Thus we have a natural restriction map $\mathcal{D}^{1}(V) \rightarrow \mathcal{D}^{1}(U)$ of first order differential operators. Thus restriction maps result:

$$
\mathcal{T}(V):=\oplus_{k}\left(\otimes^{k} \mathcal{D}^{1}(V)\right) \rightarrow \mathcal{T}(U)
$$

which are algebra homomorphisms. Clearly the ideal $\mathcal{I}_{V}$ generated by $P_{1} \otimes P_{2}-P_{2} \otimes P_{1}-\left[P_{1}, P_{2}\right] \otimes 1$ in $\mathcal{T}(V)$ maps to the corresponding ideal $I_{U} \subset \mathcal{T}(U)$, one has a natural restriction algebra homomorphism
$\mathcal{D}^{\infty}(V) \rightarrow \mathcal{D}^{\infty}(U)$ which maps $\mathcal{D}^{d}(V)$ to $\mathcal{D}^{d}(U)$. The fact that $\mathcal{D}^{d}$ is a sheaf of left and right modules over the sheaf $\mathcal{C}^{\infty}$ follows from the facts that (i) $\mathcal{D}^{1}$ is a sheaf, which implies that $\mathcal{T}$ is a sheaf, and (ii) $\mathcal{I}$ is also a sheaf of ideals inside $\mathcal{T}$ (verify!). Similarly, one forms the symbol sheaf $\mathrm{Sym}^{d}$, whose sections over $U \subset M$ are precisely the sections $\sigma$ in $C^{\infty}\left(T^{*}(U), \pi^{*} \mathbb{C}\right)$ satisfying $\sigma(t \xi)=t^{d} \sigma(\xi)$. The symbol map also becomes a sheaf map with all of these definitions.

Because of the sheaf theoretic machinery above, all the assertions of the lemma need to be verified only locally, i.e. on $M=\mathbb{R}^{n}$. In this setting, $\mathcal{D}^{1}$ is the $C^{\infty}\left(\mathbb{R}^{n}\right)$ module of all operators of the kind $\sum_{i} a_{i}(x) \partial_{i}+b(x)$ where $a_{i}, b \in C^{\infty}$. Then clearly $\mathcal{D}^{1} / \mathcal{D}^{0}$ is the space $\chi$ of smooth vector fields on $\mathbb{R}^{n}$. It is trivial to check that for any smooth vector field $X(x)=\sum_{i=1}^{n} a_{i}(x) D_{i, x},\left(\right.$ where $\left.D_{i}=\frac{1}{\sqrt{-1}} \partial_{i}\right)$, the smooth function $\sigma$ on $T^{*}\left(\mathbb{R}^{n}\right)$ defined by:

$$
\sigma\left(\xi_{x}\right)=\sqrt{-1}\left[\xi_{x}(X(x))\right] \quad \text { for } \quad \xi_{x} \in T_{x}^{*}\left(\mathbb{R}^{n}\right)
$$

satisfies $\sigma(t \xi)=t \sigma(\xi)$ by the linearity of the cotangent vector $\xi_{x}$. Since $\sigma\left(d x_{i, x}\right)=a_{i}(x)$ by this definition, it is natural to write

$$
\sigma(\xi)=\sigma\left(\sum_{i=1}^{n} \xi_{i} d x_{i, x}\right)=\sum_{i} a_{i}(x) \xi_{i}
$$

which gives precisely the leading symbol of $P=\sum_{i} a_{i}(x) D_{i, x}$. Conversely, given a $\sigma \in C^{\infty}\left(T^{*}\left(\mathbb{R}^{n}\right)\right)$ satisfying $\sigma(t \xi)=t \sigma(\xi)$, it follows that $\sigma$ is a linear functional on $T_{x}^{*}\left(\mathbb{R}^{n}\right)$, and one gets a $C^{\infty}$ vector field in $\chi$ by setting

$$
X(x)=\sum_{i=1}^{n} \sigma\left(d x_{i, x}\right) D_{i, x}
$$

It is checked immediately that these maps are inverses of each other. More generally, if $P=\sum_{|\alpha| \leq d} a_{\alpha} D_{x}^{\alpha}$ is a differential operator in $\mathcal{D}^{d}$, then $\sigma_{L}(P)$ is $\sum_{|\alpha|=d} a_{\alpha} \xi^{\alpha}$, which being a homogeneous polynomial of degree $d$, satisfies $\sigma_{L}(P)(t \xi)=t^{d} \sigma_{L}(P)$. The space of smooth functions on $T^{*}(M)$ and obeying this scaling property are precisely those functions which are homogeneous polynomials of degree $d$ in the variables $\xi_{1}, . ., \xi_{n}$, and so $\operatorname{Sym}^{d}$ is exactly $\sigma_{L}\left(\mathcal{D}^{d}\right)$. Indeed, this definition of $\sigma_{L}$ agrees with the earlier one in Definition 5.5.4. Now it is trivially checked that $\sigma_{L}(P Q)=\sigma_{L}(P) \sigma_{L}(Q)$. Thus (i) is proved.

To see (ii), note that if $d f(x)=\xi$, then $\partial_{j, x} f=\xi_{j}$, and hence for a $C^{\infty}$ function $g$, we have:

$$
D_{j, x}\left(e^{i t f} g\right)(x)=t e^{i t f} \partial_{j, x} f(x) g(x)+e^{i t f} D_{j, x} g(x)=t e^{i t f} \xi_{j} g(x)+e^{i t f}\left(D_{j, x} g\right)(x)
$$

More generally, using Leibnitz formula for differentiating a product:

$$
\begin{aligned}
D_{1, x}^{\alpha_{1}} \ldots D_{n, x}^{\alpha_{n}}\left(e^{i t f} g\right)(x) & =t^{|\alpha|} e^{i t f} \xi_{1}^{\alpha_{1}} \ldots \xi_{n}^{\alpha_{n}} g(x) \\
& +\quad \text { (terms involving strictly lower powers of } t)
\end{aligned}
$$

from which it follows that

$$
\lim _{t \rightarrow \infty} t^{-d}\left(e^{-i t f} P e^{i t f}\right)(x)=\sigma_{L}(P)
$$

and (ii) is proved.
(iii) is clear because saying that $\sigma_{L}(P)(\xi) \neq 0$ for all $|\xi|$ large enough is equivalent to saying that it is non-zero for all $\xi \neq 0$, by homogeneity of $\sigma_{L}(P)$. The lemma follows.

For $f$ a smooth function on $M$, and $P \in \mathcal{D}^{d}(M, E)$ a differential operator, we denote by $(a d f) P$ the differential operator $f P-P f$. Using the fact that $\left[\mathcal{D}^{0}, \mathcal{D}^{1}\right] \subset \mathcal{D}^{0}$ and induction, it is easy to see that $(a d f) P \in \mathcal{D}^{d-1}(M, E)$, so that $(\operatorname{ad} f)^{d} P$ is a zero-th order operator.

Corollary 9.3.4. Let $P$ and $f$ as in (iii) of the Lemma 9.3.3 above. Then
(i):

$$
\sigma_{L}(P)=\frac{(-i)^{d}}{d!}(a d f)^{d} P
$$

(ii): Let $P^{*}$ be the adjoint of $P$, defined with respect to some Hermitian inner products on $E, F$. Then

$$
\sigma_{L}\left(P^{*}\right)(\xi)=\left(\sigma_{L}(P)\right)^{*}(\bar{\xi})
$$

Proof: Note that

$$
\frac{d}{d t}\left(e^{-i t f} P e^{i t f}\right)=(-i) e^{-i t f}[(a d f) P] e^{i t f}
$$

so that we inductively have:

$$
\left(\frac{d}{d t}\right)^{d}\left(e^{-i t f} P e^{i t f}\right)=(-i)^{d} e^{-i t f}\left[(a d f)^{d} P\right] e^{i t f}=(-i)^{d}(a d f)^{d} P
$$

since $(a d f)^{d} P$ is in $\mathcal{D}^{0}$, and commutes with $e^{i t f}$. Applying L'Hospital's rule to the formula in (iii) of 9.3.3, we have (i)

To see (ii), if $d f(x)=\xi$, we have by (iii) of 9.3.3 above that:

$$
\sigma_{L}\left(P^{*}\right)(\xi)=\lim _{t \rightarrow \infty} t^{-d}\left(e^{-i t f} P^{*} e^{i t f}\right)=\lim _{t \rightarrow \infty} t^{-d}\left(e^{-i t \bar{f}} P e^{i t \bar{f}}\right)^{*}=\left(\sigma_{L}(P)\right)^{*}(\bar{\xi})
$$

The corollary follows.

### 9.4. Elliptic Complexes.

Definition 9.4.1. Let $\left\{E^{i}\right\}_{i=0}^{m}$ be complex vector bundles with Hermitian metrics. Say that a sequence of differential operators:

$$
\ldots \rightarrow C^{\infty}\left(M, E^{i}\right) \xrightarrow{P_{i}} C^{\infty}\left(M, E^{i+1}\right) \rightarrow \ldots
$$

is an elliptic complex if:
(i): $P_{i+1} \circ P_{i}=0$ for all $i$.
(ii): The associated symbol sequence

$$
\ldots \rightarrow \pi^{*} E^{i} \xrightarrow{\sigma_{L}\left(P_{i}\right)(\xi)} \pi^{*} E^{i+1}
$$

is exact for all $\xi \neq 0$ (i.e. all $\xi \in T^{*} M \backslash M$ ).
(iii): The order of each $P_{i}$ is $d>0$. (For most elliptic complexes of concern to us, $d=1$ ).

Clearly, if we only have a two term sequence $C^{\infty}\left(M, E^{0}\right) \xrightarrow{P} C^{\infty}\left(M, E^{1}\right)$, then this two term complex is elliptic iff $P$ is an elliptic operator of order $d>0$.

Before looking at some examples of elliptic complexes, let us note the following:
Lemma 9.4.2. Let $\left\{C^{\infty}\left(M, E^{*}\right), P_{*}\right\}$ be a complex of differential operators (i.e. $P_{i+1} \circ P_{i}=0$ for all $i$ ). Define the Laplacian of this complex by:

$$
\Delta_{P}^{i}=P_{i}^{*} P_{i}+P_{i-1} P_{i-1}^{*}: C^{\infty}\left(M, E^{i}\right) \rightarrow C^{\infty}\left(M, E^{i}\right)
$$

Then the complex above is elliptic iff $\Delta_{P}^{i}$ is an elliptic operator for each $i$.

Proof: Let us denote $\sigma_{L}\left(P_{i}\right)=p_{i}$. Let us assume that the complex is elliptic. Then, from (ii) of the Corollary 9.3.4, and (i) of 9.3.3 that $\sigma_{L}$ is an algebra homomorphism, it follows that:

$$
\sigma_{L}\left(\Delta_{P}^{i}\right)(\xi)=p_{i}^{*}(\xi) p_{i}(\xi)+p_{i-1}(\xi) p_{i-1}^{*}(\xi)
$$

If for some $e \in \pi^{*} E^{i}, \sigma_{L}\left(\Delta_{P}^{i}\right)(\xi) e=0$, and $\xi \neq 0$, it follows that with respect to the Hermitian inner product $\langle-,-\rangle$ on $\pi^{*} E^{i}$, we have:

$$
\left\langle p_{i}(\xi) e, p_{i}(\xi) e\right\rangle+\left\langle p_{i-1}^{*}(\xi) e, p_{i-1}^{*}(\xi) e\right\rangle=0
$$

which implies that $p_{i}(\xi) e=0$ and $p_{i-1}^{*}(\xi) e=0$. Since the complex is elliptic, and $\xi \neq 0$, it follows that $e=$ $p_{i-i}(\xi) v$ for $v \in \pi^{*}\left(E^{i-1}\right)$. Since $p_{i-1}^{*}(\xi) e=0$, it follows that $p_{i-1}^{*}(\xi) p_{i-1}(\xi) v=0$. Thus $\left\langle v, p_{i-1}^{*}(\xi) p_{i-1}(\xi) v\right\rangle=$ 0 , which implies that $p_{i-1}(\xi) v=e=0$. Thus $\sigma_{L}\left(\Delta_{P}^{i}\right)(\xi): \pi^{*} E^{i} \rightarrow \pi^{*} E^{i}$ is a monomorphism, and hence an isomorphism. That is $\Delta_{P}^{i}$ is elliptic.

The converse is similar, and left as an exercise.

Remark 9.4.3. Note that if $\mathcal{P}$ is an elliptic complex, and of finite length (i.e $E^{i}=0$ for $i \gg 0$ ), and $\operatorname{dim} M>0$, then we have

$$
\sum_{i=0}^{\infty} \operatorname{rank} E^{2 i}=\sum_{i=0}^{\infty} \operatorname{rank} E^{2 i+1}
$$

This is because we can choose a $\xi \neq 0$ in $T_{x}^{*} M$, and the fact that the symbol complex:

$$
\ldots \rightarrow E_{x}^{i} \xrightarrow{\sigma_{L}(\xi)} E_{x}^{i+1} \rightarrow \ldots
$$

is exact means that the alternating sum:

$$
\sum_{i=0}^{\infty}(-1)^{i} \operatorname{dim} E_{x}^{i}=\sum_{i=0}^{\infty}(-1)^{i} \operatorname{rank} E^{i}=0
$$

which implies our assertion.

Example 9.4.4 (The de-Rham Complex). Set $E^{i}=\Lambda^{i}\left(T_{\mathbb{C}}^{*} M\right)$, the $i$-th exterior power of the (complexified) cotangent bundle of $M$. Then consider the de-Rham complex:

$$
\ldots \rightarrow C^{\infty}\left(M, E^{i}\right)=: \Lambda^{i}(M, \mathbb{C}) \xrightarrow{d_{i}} \Lambda^{i+1}(M, \mathbb{C}) \rightarrow \ldots
$$

If $\xi_{x}=\sum_{i} \xi_{j} d x_{j, x}$ is a real cotangent vector, in some local coordinate system, then since for $\omega=\sum_{|I|=i} \omega_{I} d x_{I} \in$ $\Lambda^{i}(M, \mathbb{C})$ we have the representation of $d \omega$ in local coordinates:

$$
d \omega=\sum_{j} d x_{j} \partial_{j} \wedge \omega=\sqrt{-1}\left(\sum_{j} d x_{j} D_{x_{j}}\right) \wedge \omega
$$

it follows that $\sigma_{L}(d)(\xi)=\sqrt{-1}\left(\sum_{j} d x_{j} \xi_{j}\right) \wedge(-)=i \xi \wedge(-)$. One already knows that this is a complex of differential operators, i.e. $d_{i+1} \circ d_{i}=0$, so to show that the complex is elliptic, it is enough to show that the operator $e(\xi):=\xi \wedge(-)$ is exact for $\xi \neq 0$. Since $\xi \neq 0$, we may complete it to a basis $\left\{e_{i}\right\}_{i=1}^{p}$ of $\left.\pi^{*}\left(T^{*} M\right)\right)_{\xi}=T_{x}^{*}(M)$ with $e_{p}=\xi$. Then, each $\alpha \in \Lambda^{p}\left(T_{\mathbb{C}, x}^{*}(M)\right)$ may be uniquely written as:

$$
\alpha=\alpha_{1}+\xi \wedge \alpha_{2}
$$

where $\alpha_{1}, \alpha_{2}$ do not involve $\xi=e_{p}$. If $\xi \wedge \alpha=0$, it follows that $\xi \wedge \alpha_{1}=0$, but since $\alpha_{1}$ does not involve $\xi$, this implies $\alpha_{1}=0$. Thus $\alpha=\xi \wedge \alpha_{2}$. This proves that the de-Rham complex is elliptic.

From the lemma 9.4.2 above, it follows that all the Laplacians $\Delta^{i}=d d^{*}+d^{*} d$ of the de-Rham complex are elliptic operators.

Example 9.4.5 (Twisted Dolbeault Complex). Let $M$ be a compact complex manifold of $\operatorname{dim}_{\mathbb{C}} M=n$. Let $E$ be a holomorphic vector bundle on $M$ of $\operatorname{rk}_{\mathbb{C}} E=k$. We have the following well known decomposition (as complex vector bundles) for the complexification of the real tangent bundle $T_{\mathbb{R}} M$ :

$$
T_{\mathbb{C}} M=T_{\mathbb{R}} M \otimes_{\mathbb{R}} \mathbb{C}=T^{1,0} M \oplus T^{0,1} M
$$

where $T^{1,0} M$ is the holomorphic tangent bundle of $M$, and $T^{0,1} M$ is its complex conjugate bundle, and called the anti-holomorphic tangent bundle of $M$. In a local holomorphic coordinate chart $U \subset M$, we may write $v \in T^{1,0} M_{\mid U}$ as:

$$
v=\sum_{j=1}^{n} \alpha_{j} \frac{\partial}{\partial z_{j}}
$$

and correspondingly $w \in T^{0,1} M_{\mid U}$ as:

$$
w=\sum_{j=1}^{n} \beta_{j} \frac{\partial}{\partial \bar{z}_{j}}
$$

The decomposition of $T_{\mathbb{C}} M$ leads to a corresponding decomposition of $T_{\mathbb{C}}^{*} M=\operatorname{hom}_{\mathbb{R}}\left(T_{\mathbb{R}} M, \mathbb{C}\right)$ as:

$$
T_{\mathbb{C}}^{*} M=\left(T^{1,0} M\right)^{*} \oplus\left(T^{0,1} M\right)^{*}
$$

Thus

$$
\Lambda^{i}\left(T_{\mathbb{C}}^{*} M\right)=\oplus_{p+q=i}\left(\Lambda^{p}\left(T^{1,0} M\right)^{*} \otimes \Lambda^{q}\left(T^{0,1} M\right)^{*}\right)=: \oplus_{p+q=i} \Lambda^{p, q}\left(T_{\mathbb{C}}^{*} M\right)
$$

Again, in a local holomorphic chart over $U$, and element $\omega \in \Lambda^{p, q}\left(T^{*} M\right)_{\mid U}$ has the representation:

$$
\omega=\sum_{|I|=p,|J|=q} \alpha_{I J} d z_{I} \wedge d \bar{z}_{J}
$$

where we use the notation:

$$
d z_{I}:=d z_{i_{1}} \wedge d z_{i_{2}} \wedge \ldots \wedge d z_{i_{p}}, \quad d \bar{z}_{J}=d \bar{z}_{j_{1}} \wedge d \bar{z}_{j_{2}} \wedge \ldots \wedge d \bar{z}_{j_{q}}
$$

We can tensor all this with the bundle $E$, and thus we have:

$$
\Lambda^{i}\left(T^{*} M\right) \otimes_{\mathbb{C}} E=\oplus_{p+q=i} \Lambda^{p, q}\left(T_{\mathbb{C}}^{*} M\right) \otimes_{\mathbb{C}} E
$$

Thus, for smooth sections of the above bundle, we have:

$$
\Lambda^{i}(M, E):=C^{\infty}\left(M, \Lambda^{i}\left(T^{*} M\right) \otimes_{\mathbb{C}} E\right)=\oplus_{p+q=i} \Lambda^{p, q}(M, E)
$$

where $\Lambda^{p, q}(M, E):=C^{\infty}\left(M, \Lambda^{p, q}\left(T_{\mathbb{C}}^{*} M\right) \otimes_{\mathbb{C}} E\right)$. Again, for $U$ a coordinate chart, a section $\omega \in \Lambda^{p, q}(U, E)$ has the representation:

$$
\omega=\sum_{|I|=p,|J|=q} \alpha_{I J} d z_{I} \wedge d \bar{z}_{J}
$$

where $\alpha_{I J} \in C^{\infty}(U, E)$ are smooth sections of $E_{\mid U}$.
Now we can define the Dolbeault operator

$$
\bar{\partial}^{E}: \Lambda^{p, q}(M, E) \rightarrow \Lambda^{p, q+1}(M, E)
$$

by defining it on local representations as follows. On a coordinate chart $U$, write:

$$
\omega_{\mid U}=\sum_{|I|=p,|J|=q} \alpha_{I J} d z_{I} \wedge d \bar{z}_{J}
$$

with $\alpha_{I J} \in C^{\infty}(U, E)=\Lambda^{0,0}(U, E)$, and set

$$
\bar{\partial}^{E} \omega_{\mid U}=\sum_{|I|=p,|J|=q} \bar{\partial}^{E} \alpha_{I J} \wedge d z_{I} \wedge d \bar{z}_{J}
$$

where

$$
\bar{\partial}^{E} \alpha_{I J}=\sum_{j=1}^{n} \frac{\partial \alpha_{I J}}{\partial \bar{z}_{j}} d \bar{z}_{j}
$$

The thing to verify is that all this is globally defined, and the reason it is globally defined is that $\bar{\partial}^{E} \alpha$ is globally defined as an element of $\Lambda^{0,1}(U, E)$ for $\alpha \in C^{\infty}(U, E)=\Lambda^{0,0}(U, E)$, and $U \subset M$ any open set. For, over a $W$ satisfying $E_{\mid W}$ is holomorphically trivial, we can write $\alpha$ as $\alpha=\sum_{i=1}^{k} \alpha_{i} e_{i}$ where $\alpha_{i}$ are smooth functions on $W$, and $\left\{e_{i}\right\}_{i=1}^{k}$ is a holomorphic frame for $E_{\mid W}$. Then we set:

$$
\bar{\partial}^{E} \alpha=\sum_{i=1}^{k} \bar{\partial} \alpha_{i} e_{i}
$$

where, on a coordinate chart with coordinates $z_{1}, . ., z_{n}$, we have

$$
\bar{\partial} \alpha_{i}=\sum_{j=1}^{n} \frac{\partial \alpha_{i}}{\partial \bar{z}_{j}} d \bar{z}_{j}
$$

the usual $\bar{\partial}$ operator on smooth complex valued functions. That this $\bar{\partial}$-operator on smooth functions is welldefined follows from the fact that coordinate changes on $M$ are holomorphic.

If we change to another holomorphic frame $\left\{f_{j}\right\}_{j=1}^{k}$ for $E_{\mid V}$, where $V \subset M$ is another open set, we have the transition relation $e_{i}=\sum_{j} g_{j i} f_{j}$, where $g_{i j}$ are holomorphic functions on $V \cap W$, and thus

$$
\alpha=\sum_{i} \alpha_{i} e_{i}=\sum_{j}\left(\sum_{i} g_{j i} \alpha_{i}\right) f_{j}
$$

and since $g_{j i}$ are holomorphic, we have $\bar{\partial}\left(g_{j i} \alpha_{i}\right)=g_{j i} \bar{\partial} \alpha_{i}$, so that

$$
\sum_{i} \bar{\partial} \alpha_{i} e_{i}=\sum_{j}\left(\sum_{i} g_{j i} \bar{\partial} \alpha_{i}\right) f_{j}=\sum_{j} \bar{\partial}\left(\sum_{i} g_{j i} \alpha_{i}\right) f_{j}
$$

which shows that our definition of $\bar{\partial}^{E} \alpha$ makes global sense, independent of local holomorphic frames.
It is now easy to check, using local coordinates, that $\bar{\partial} \circ \bar{\partial}=0$, so that we have the twisted Dolbeault Complex

$$
\ldots \rightarrow \Lambda^{p, q}(M, E)=C^{\infty}\left(M, \Lambda^{p, q} T_{\mathbb{C}}^{*}(M) \otimes_{\mathbb{C}} E\right) \xrightarrow{\bar{\partial}^{E}} \Lambda^{p, q+1}(M, E) \rightarrow \ldots
$$

of differential operators. That is, we are taking the complex vector bundle $E^{q}:=\Lambda^{p, q} T_{\mathbb{C}}^{*} M \otimes E$, with a fixed $p$. We can easily equip these smooth complex vector bundles with some Hermitian metrics, arising from a Hermitian metric on the bundles $T_{\mathbb{C}} M$ and $E$.

To check that this is an elliptic complex, one needs to calculate the symbol of $\bar{\partial}^{E}$. We note that the the complex vector bundle $\left(T^{0,1} M\right)^{*}$ can be identified with the real cotangent bundle $T^{*} M$, by forgetting its complex structure, and with this notation $\bar{\partial}=i \sum_{j} d \bar{z}_{j} \frac{\partial}{\partial \bar{z}_{j}} \wedge(-)$, so that its symbol is given by:

$$
\sigma_{L}(\bar{\partial})(\xi)=\frac{i}{2} \sigma_{L}\left(\sum_{j} d \bar{z}_{j}\left(D_{x_{j}}+i D_{y_{j}}\right) \wedge(-)\right)=\frac{i}{2} \sum_{j} d \bar{z}_{j}\left(\xi_{j}^{1}+i \xi_{j}^{2}\right)=\frac{i}{2} \xi \wedge(-)
$$

where $\xi=\sum_{j} \xi_{j} d \bar{z}_{j}=\sum_{j}\left(\xi_{j}^{1}+i \xi_{j}^{2}\right) d \bar{z}_{j} \in T^{0,1}(M)^{*}$. The reason that $\sigma_{L}(\bar{\partial})(\xi)$ is exact for $\xi \neq 0$ is the same as that for the de Rham complex above, so we omit the argument.

### 9.5. The Hodge Theorem for Elliptic Complexes.

Definition 9.5.1. Let $\mathcal{P}$ denote an elliptic complex:

$$
\ldots . \rightarrow C^{\infty}\left(M, E^{i}\right) \xrightarrow{P_{i}} C^{\infty}\left(M, E^{i+1}\right) \rightarrow \ldots
$$

on a compact Riemannian manifold $M$. We define the $i$-th cohomology of this complex to be the $\mathbb{C}$-vextor space:

$$
H^{i}(M, \mathcal{P}):=\frac{\operatorname{ker} P_{i}}{\operatorname{Im} P_{i-1}}
$$

For example, in the case of the de-Rham complex of Example 9.4.4 above, this gives the de Rham cohomology of $M$ (with complex coefficients). In the case of the twisted Dolbeault complex of Example 9.4.5 above, it gives the $(p, q)$-Dolbeault cohomology with coefficients in $E$, and is denoted by $H^{p, q}(M, E)$, which algebraic geometers write as $H^{q}\left(M, \Omega^{p}(E)\right)$ for reasons we needn't explore here.

Theorem 9.5.2 (Hodge Theorem for Elliptic Complexes). Let $\mathcal{P}$ be an elliptic complex on a compact Riemannian manifold $M$. Let $\Delta_{P}^{i}: C^{\infty}\left(M, E^{i}\right) \rightarrow C^{\infty}\left(M, E^{i}\right)$ be the Laplacian introduced in the Lemma 9.4.2. Then:
(i): $H^{i}(M, \mathcal{P}) \simeq \operatorname{ker} \Delta_{P}^{i}$, and this cohomology is a finite dimensional space.
(ii): (Kodaira-Hodge decomposition) The $L_{2}$-space of sections $L_{2}\left(M, E^{i}\right)=H_{0}\left(M, E^{i}\right)$ admits the $L_{2^{-}}$ orthogonal direct sum decomposition:

$$
L_{2}\left(M, E^{i}\right)=\operatorname{ker} \Delta_{P}^{i} \oplus P_{i}^{*}\left(H_{d}\left(M, E^{i+1}\right)\right) \oplus P_{i-1}\left(H_{d}\left(M, E^{i-1}\right)\right)
$$

where each space on the right is a closed Hilbert subspace.

Proof: Let us first prove (ii), and then (i) will follow as a consequence. Since $\mathcal{P}$ is an elliptic complex, by the Lemma 9.4.2, the operators

$$
\Delta_{P}^{i}: C^{\infty}\left(M, E^{i}\right) \rightarrow C^{\infty}\left(M, E^{i}\right)
$$

are elliptic operators of the same order $2 d \geq 2$, where $d=$ ord $P_{i} \geq 1$. Also by its definition, it is self-adjoint. By (i) of Proposition 8.4.8, the kernel

$$
\mathcal{H}_{P}^{i}:=\operatorname{ker} \Delta_{P}^{i}
$$

is a finite dimensional subspace inside $C^{\infty}\left(M, E^{i}\right)$, and therefore a finite-dimensional closed subspace of $H_{d}\left(M, E^{i}\right)$ for all $d$. Let $\pi: L_{2}\left(M, E^{i}\right)=H_{0}\left(M, E^{i}\right) \rightarrow \mathcal{H}_{P}^{i}$ be the $L_{2}$-orthogonal projection. Then for all $f \in L_{2}\left(M, E^{i}\right)$, we have $f-\pi(f) \in \mathcal{H}_{P}^{i, \perp}$. Then let

$$
G_{P}^{i}: L_{2}\left(M, E^{i}\right) \rightarrow L_{2}\left(M, E^{i}\right)
$$

denote the Green operator from Proposition 8.4.8. By (a) of (iii) in that proposition, we have:

$$
\Delta_{P}^{i} G_{P}^{i}(f-\pi(f))=f-\pi(f)
$$

Again, by (a) of (iii) of the Proposition 8.4.8, we have $G_{P}^{i}(\pi(f))=0$, since $\pi(f) \in \operatorname{ker} \Delta_{P}^{i}$. Thus we have:

$$
f=\pi(f)+\Delta_{P}^{i} G_{P}^{i}(f)=\pi(f)+P_{i}^{*}\left(P_{i} G_{P}^{i} f\right)+P_{i-1}\left(P_{i-1}^{*} G_{P}^{i} f\right) \quad \text { for all } \quad f \in L_{2}\left(M, E^{i}\right)
$$

By the construction of the Green operator in 8.4.8, $G_{P}^{i} f \in H_{2 d}\left(M, E^{i}\right)$, which implies that $P_{i} G_{P}^{i} f \in H_{d}\left(M, E^{i+1}\right)$. Similarly, $P_{i-1}^{*} G_{P}^{i} f \in H_{d}\left(M, E^{i-1}\right)$. The computation above therefore shows that:

$$
L_{2}\left(M, E^{i}\right)=\mathcal{H}_{P}^{i}+P_{i-1}\left(H_{d}\left(M, E^{i-1}\right)+P_{i}^{*}\left(H_{d}\left(M, E^{i+1}\right)\right.\right.
$$

We denote the last two spaces above by $\operatorname{Im} P_{i-1}$ and $\operatorname{Im} P_{i}^{*}$ respectively.
To check that the decomposition is orthogonal, we easily check that $\mathcal{H}_{P}^{i}=\operatorname{ker} P_{i} \cap \operatorname{ker} P_{i-1}^{*}$ from the definition of $\Delta_{P}^{i}$. Hence for $\alpha \in \mathcal{H}_{P}^{i}$, we have:

$$
\left(\alpha, P_{i}^{*} \beta\right)=\left(P_{i} \alpha, \beta\right)=0
$$

for all $\beta \in H_{d}\left(M, E^{i+1}\right)$. Hence $\mathcal{H}_{P}^{i}$ is orthogonal to $\operatorname{Im} P_{i}^{*}$. Similarly, it is orthogonal to $\operatorname{Im} P_{i-1}$. Finally, if we have $\alpha=P_{i-1} \beta$ and $\gamma=P_{i}^{*} \delta$, then:

$$
(\alpha, \gamma)=\left(P_{i-1} \beta, P_{i}^{*} \delta\right)=\left(P_{i} P_{i-1} \beta, \delta\right)=0
$$

since $P_{i} P_{i-1}=0$. This shows that $\operatorname{Im} P_{i}^{*}$ and $\operatorname{Im} P_{i-1}$ are also mutually orthogonal. We need to check that both these images are closed. Note that if $\alpha \in L_{2}\left(M, E^{i}\right)$ and $\alpha \in \mathcal{H}_{P}^{i}+P_{i-1}\left(H_{d}\left(M, E^{i-1}\right)\right)$, then $P_{i} \alpha=0$. Conversely, if $P_{i}: L_{2}\left(M, E^{i}\right) \rightarrow H_{-d}\left(M, E^{i+1}\right)$ annihilates $\alpha$, we write:

$$
\alpha=\alpha_{1}+P_{i-1} \beta+P_{i}^{*} \gamma
$$

by the decomposition above, where $\gamma \in H_{d}\left(M, E^{i+1}\right)$. Now note that $P_{i} \alpha=0$ implies that the element $P_{i} P_{i}^{*} \gamma \in H_{-d}\left(M, E^{i+1}\right)$ is zero. This implies that under the natural pairing $\langle-,-\rangle$ of $H_{d}\left(M, E^{i+1}\right)$ and $H_{-d}\left(M, E^{i+1}\right)$ (see (iii) of Proposition 4.2.2), we have:

$$
\left\langle\gamma, P_{i} P_{i}^{*} \gamma\right\rangle=\left(P_{i}^{*} \gamma, P_{i}^{*} \gamma\right)_{0}=0
$$

which implies that $P_{i}^{*} \gamma=0$, and $\alpha \in \mathcal{H}_{P}^{i}+\operatorname{Im} P_{i-1}$. Thus $\operatorname{Im} P_{i}^{*}$ is precisely the orthogonal complement of the subspace ker $P_{i}$ in $L_{2}\left(M, E^{i}\right)$, and since the orthogonal complement of any subspace is closed, we have $\operatorname{Im} P_{i}^{*}$ is closed. Similarly, $\operatorname{Im} P_{i-1}$ is the orthogonal complement of $\operatorname{ker} P_{i-1}^{*}$ in $L_{2}\left(M, E^{i}\right)$ and also closed. This proves (ii).

In fact, since both $G_{P}^{i}$ and $\Delta_{P}^{i}$ map smooth forms to smooth forms, as do $P_{i-1}$ and $P_{i}^{*}$, and $\mathcal{H}_{P}^{i} \subset$ $C^{\infty}\left(M, E^{i}\right)$, we can restrict the decomposition above to obtain:

$$
C^{\infty}\left(M, E^{i}\right)=\mathcal{H}_{P}^{i} \oplus P_{i-1}\left(C^{\infty}\left(M, E^{i-1}\right)\right) \oplus P_{i}^{*}\left(C^{\infty}\left(M, E^{i+1}\right)\right)
$$

which is $L_{2}$-orthogonal, but of course left hand space and the two right hand spaces are no longer closed in $L_{2}\left(M, E^{i}\right)$. Again it is readily checked that

$$
\operatorname{ker}\left\{P_{i}: C^{\infty}\left(M, E^{i}\right) \rightarrow C^{\infty}\left(M, E^{i}\right)\right\}=\mathcal{H}_{P}^{i} \oplus P_{i-1}\left(C^{\infty}\left(M, E^{i-1}\right)\right)
$$

which implies that

$$
H^{i}(M, \mathcal{P})=\frac{\operatorname{ker}\left\{P_{i}: C^{\infty}\left(M, E^{i}\right) \rightarrow C^{\infty}\left(M, E^{i}\right)\right\}}{P_{i-1}\left(C^{\infty}\left(M, E^{i-1}\right)\right)}=\mathcal{H}_{P}^{i}
$$

and, indeed the natural composite map:

$$
\mathcal{H}_{P}^{i} \rightarrow \operatorname{ker}\left\{P_{i}: C^{\infty}\left(M, E^{i}\right) \rightarrow C^{\infty}\left(M, E^{i}\right)\right\} \rightarrow H^{i}(M, \mathcal{P})
$$

is the required isomorphism. This proves (i), and the theorem follows.

Corollary 9.5.3 (Hodge-deRham Theorem). By the Theorem 9.5.2 above applied to the elliptic de Rham complex of Example 9.4.4 above, we have that the $i$-th de Rham cohomology of a compact manifold $M$ satisfies:

$$
H_{d R}^{i}(M, \mathbb{C}) \simeq \mathcal{H}^{i}
$$

where $\mathcal{H}^{i}=\operatorname{ker}\left\{\Delta^{i}: \Lambda^{i}(M, \mathbb{C}) \rightarrow \Lambda^{i+1}(M, \mathbb{C})\right\}$ is the space of harmonic $i$-forms on $M$. In particular, by the above theorem, this cohomology is finite dimensional. This is, incidentally, provable by using the de Rham theorem which is highly non-trivial, and the fact that a compact smooth manifold is a finite CW-complex, which again uses non-trivial Morse Theory. That is, the finite dimensionality of the de Rham cohomology of a compact smooth manifold, whichever way one chooses to prove it, is a very deep result.

Corollary 9.5.4 (Hodge-Dolbeault Theorem). By the Theorem 9.5.2 applied to the twisted Dolbeault complex of Example 9.4.5, it follows that the twisted Dolbeault cohomology $H^{p, q}(M, E)$ of a compact complex manifold $M$ and holomorphic coefficient bundle $E$ satisfies:

$$
H^{p, q}(M, E) \simeq \mathcal{H}^{p, q}(M, E)
$$

where the space on the right is the kernel of the Hodge-Dolbeault Laplacian $\square:=\bar{\partial}^{E} \bar{\partial}^{E *}+\bar{\partial}^{E *} \bar{\partial}^{E}$ inside $\Lambda^{p, q}(M, E)$. Again the theorem implies that this Dolbeault cohomology is finite dimensional. In the case when $E$ is the trivial line bundle $M \times \mathbb{C}$, and $p=0$, the Dolbeault cohomology $H^{p, q}(M, E)$ is simply denoted $H_{\bar{\partial}}^{0, q}(M)$ or simply $H^{0, q}(M)$. Then, by the above, the alternating sum:

$$
\sum_{q=0}^{n}(-1)^{q} \operatorname{dim} H^{0, q}(M)
$$

is finite. Again, that this is finite for the situation above has to be proved as above, and is a very deep fact.
9.6. Index of an elliptic complex. We observed in (a) of (iii) in Proposition 8.4.8 that for a formally self-adjoint elliptic differential operator $P: C^{\infty}(M, E) \rightarrow C^{\infty}(M, E)$ of order $d>0$, the (finite dimensional) cokernel Coker $P=(\operatorname{Im} P)^{\perp}=\operatorname{ker} P$, so that the index of the Fredholm operator $P: H_{d}(M, E) \rightarrow H_{0}(M, E)$ is ind $P=\operatorname{dim} \operatorname{ker} P-\operatorname{dim}$ Coker $P=0$. Thus we won't get any interesting index by considering the indices of the elliptic (of order 2d) Laplacians $\Delta_{P}^{i}$ of an elliptic complex $\mathcal{P}$. On the other hand, a profound idea due to Dirac (who introduced it to explain electron spin) suggests that we find a "square root" of the Laplacian to get an interesting index.

What one does instead is construct an operator of order $d$ as follows.

Definition 9.6.1 (The Dirac operator of an elliptic complex). Let $M$ be a compact oreinted Riemannian manifold, and let $\mathcal{P}$ be the elliptic complex:

$$
\ldots \rightarrow C^{\infty}\left(M, E^{i}\right) \xrightarrow{P_{i}} C^{\infty}\left(M, E^{i+1}\right) \rightarrow \ldots .
$$

where $P_{i}$ is a differential operator of order $d>0$ for each $i$. Let us assume that this complex is of finite length, i.e. $E^{i}=0$ for $i$ large enough. Define $E^{+}=\oplus_{i=0}^{\infty} E^{2 i}$ and $E^{-}=\oplus_{i=0}^{\infty} E^{2 i+1}$. Note that by the Remark 9.4.3, the smooth complex vector bundles $E^{+}$and $E^{-}$have the same rank. Then we define the following operators of order $d$ :

$$
\begin{aligned}
& D^{+}:=P_{+}+P_{-}^{*}: C^{\infty}\left(M, E^{+}\right) \rightarrow C^{\infty}\left(M, E^{-}\right) \\
& D^{-}:=P_{-}+P_{+}^{*}: C^{\infty}\left(M, E^{-}\right) \rightarrow C^{\infty}\left(M, E^{+}\right)
\end{aligned}
$$

where $P^{+}:=\oplus_{i} P_{2 i}, P_{-}=\oplus_{i} P_{2 i+1}$. These operators are called the Dirac operators of the elliptic complex $\mathcal{P}$.

Proposition 9.6.2. In the setting of the Definition 9.6 .1 above, we have:
(i): $D^{+}$and $D^{-}$are formal adjoints of each other, and are differential operators of order $d$.
(ii): The composite $D^{-} D^{+}=\oplus_{i \geq 0} \Delta_{P}^{2 i}$ and similarly the composite $D^{+} D^{-}=\oplus_{i \geq 0} \Delta_{P}^{2 i+1}$. Thus the two term complex:

$$
D^{+}: C^{\infty}\left(M, E^{+}\right) \rightarrow C^{\infty}\left(M, E^{-}\right)
$$

is an elliptic complex, with associated Laplacian being

$$
\Delta_{P}^{+}:=D^{-} D^{+}=\oplus_{i \geq 0} \Delta_{P}^{2 i}: C^{\infty}\left(M, E^{+}\right) \rightarrow C^{\infty}\left(M, E^{+}\right)
$$

This Laplacian $\Delta_{P}^{+}$is elliptic and (formally) self-adjoint. Similarly one can construct the other elliptic formally self-adjoint Laplacian $\Delta_{P}^{-}:=D^{+}: D^{-}$acting on $C^{\infty}\left(M, E^{-}\right)$.
(iii): The operators

$$
D^{ \pm}: H_{d}\left(M, E^{ \pm}\right) \rightarrow L_{2}\left(M, E^{\mp}\right)
$$

are elliptic, and hence Fredholm. Their Fredholm index is given by:

$$
\operatorname{ind} D^{+}=\sum_{i=0}^{\infty}(-1)^{i}\left(\operatorname{dim} \operatorname{ker} \Delta_{P}^{i}\right)=\sum_{i=0}^{\infty}(-1)^{i} \operatorname{dim} H^{i}(M, \mathcal{P})=-\operatorname{ind} D^{-}
$$

Proof: The assertion (i) is clear from the definitions.
(ii) is also clear from the definitions. The assertion that this two term complex is elliptic follows from the fact that the associated Laplacian is precisely $\Delta_{P}^{+}=\oplus_{i \geq 0} \Delta_{P}^{2 i}$, which is elliptic, and formally self adjoint, by the Lemma 9.4.2.

We have seen that a two term complex is elliptic iff the operator in this complex is elliptic, so $D^{+}$(and hence its formal adjoint $D^{-}$) is an elliptic operator. One easily checks that $D^{+} f=0$ iff $\Delta_{P}^{+} f=D^{-} D^{+} f=0$ for $f \in H_{d}\left(M, E^{+}\right)$, by using the fact that

$$
\left(D^{+} f, g\right)=\left(f, D^{-} g\right), \text { for all } f \in H_{d}\left(M, E^{+}\right), g \in L_{2}\left(M, E^{-}\right)
$$

which follows from the duality of $H_{d}\left(M, E^{+}\right)$and $H_{-d}\left(M, E^{+}\right)$of (iii) in Proposition 4.2.2 and that the above formula holds for $f, g$ smooth (i.e. $D^{+}$and $D^{-}$are formal adjoints of each other). Likewise for the adjoint $D^{-}$, we have $f \in \operatorname{ker} D^{-}$iff $f \in \operatorname{ker} \Delta_{P}^{-}$. Hence the index of $D^{+}$and $D^{-}$satisfy: ind $D^{+}=\operatorname{dim} \operatorname{ker} D^{+}-\operatorname{dim} \operatorname{ker} D^{-}=\sum_{i \geq 0}\left(\operatorname{dim} \operatorname{ker} \Delta_{P}^{2 i}-\operatorname{dim} \operatorname{ker} \Delta_{P}^{2 i+1}\right)=\sum_{i \geq 0}(-1)^{i} \operatorname{dim} \operatorname{ker} \Delta_{P}^{i}=-\operatorname{ind} D^{-}$

The fact that $\operatorname{dim} \operatorname{ker} \Delta_{P}^{i}=\operatorname{dim} H^{i}(M, \mathcal{P})$ follows from (i) of the Hodge Theorem 9.5.2. The proposition follows.

Note that for the De Rham complex of Example 9.4.4, the Dirac operator is $d+d^{*}=d+\delta$, and its index is the Euler characteristic of $M$. For the twisted Dolbeault complex of Example 9.4.5, the associated Dirac operator is $\bar{\partial}^{E}+\bar{\partial}^{E *}$, and its index is the quantity $\sum_{q}(-1)^{q} H^{p, q}(M, E)$.

Remark 9.6.3. Aside from the fact that the Dirac operator construction leads to an interesting index, it also shows that no generality is lost by considering two-term elliptic complexes instead of a general elliptic complex of finite length. We will henceforth restrict ourselves to this setting for analytical considerations, though finite length elliptic complexes will always be in the background because they arise from natural geometric considerations, e.g. the de Rham complex, the twisted Dolbeault complex, and the signature and spin complexes that will arise later.

## 10. Heat kernels

10.1. Heat Operators on Compact Manifolds. We now confine ourselves to the setting of the Proposition 9.6.2. That is, we have two Hermitian smooth vector bundles $E^{ \pm}$on $M$, and elliptic operators of order $d>0$ $D^{+}$and $D^{-}$fulfilling all the conclusions of 9.6.2. In particular, by the Propositions 8.4.8 and 8.4.9, we know that the spectra of $\Delta_{P}^{+}$and $\Delta_{P}^{-}$are discrete, with absolute values of eigenvalues $\left|\lambda_{n}\right| \geq C n^{\delta}$ for some $\delta>0$. Actually, one can say more:

Proposition 10.1.1. In the setting of Proposition 9.6.2, the spectrum of $\Delta_{P}^{+}$and $\Delta_{P}^{-}$satisfies $\lambda_{n} \geq 0$. Thus if we arrange the eigenvalues of $\Delta_{P}^{+}$in non-decreasing order:

$$
\lambda_{0} \leq \lambda_{1} \leq \ldots \leq \lambda_{n} \leq \ldots
$$

we have constants $C, \delta>0$ so that $\lambda_{n} \geq C n^{\delta}$ for all $n$. Likewise for $\Delta_{P}^{-}$.

Proof: Let $e_{n}$ be a basis of $L_{2}$-orthonormal smooth eigensections in $L_{2}\left(M, E^{+}\right)$, for the elliptic self-adjoint operator $\Delta_{P}^{+}$, via (d) of the Proposition 8.4.8. Then

$$
\lambda_{n}=\left(\Delta_{P}^{+} e_{n}, e_{n}\right)=\left(D^{-} D^{+} e_{n}, e_{n}\right)=\left(D^{+} e_{n}, D^{+} e_{n}\right) \geq 0
$$

where the right equality follows from the fact that $D^{+} e_{n}$ is also smooth, and $D^{+}$and $D^{-}$are formal adjoints of each other by (i) of 9.6.2. The last assertion follows from the Proposition 8.4.9.

Proposition 10.1.2. Let $t \in(0, \infty)$. Define the operator $e^{-t \Delta_{P}^{+}}$by defining its action on the eigensections $e_{n}$ of the last proposition by $e^{-t \Delta_{P}^{+}} e_{n}=e^{-t \lambda_{n}} e_{n}$ (i.e. by "functional calculus"). Then this extends to a bounded self-adjoint operator:

$$
e^{-t \Delta_{P}^{+}}: L_{2}\left(M, E^{+}\right) \rightarrow L_{2}\left(M, E^{+}\right)
$$

called the heat operator of $\Delta_{P}^{+}$. For all $t \in(0, \infty)$, this operator is infinitely smoothing, gets defined on $H_{d}\left(M, E^{+}\right)$for all $d$, and when viewed as an operator $H_{d} \rightarrow L_{2}$, is compact, for all $d$. The analogous statement holds for $e^{-t \Delta_{P}^{-}}$.

Proof: Write an element $f \in L_{2}\left(M, E^{+}\right)$as:

$$
f=\sum_{n=0}^{\infty} a_{n} e_{n}
$$

where $\sum_{n}\left|a_{n}\right|^{2}=\|f\|^{2}<\infty$. Since we have $\lambda_{n} \geq C n^{\delta}$ by the last Proposition 10.1.1, we have $e^{-t \lambda_{n}} \leq e^{-t C n^{\delta}}$ for all $n$. Since $t>0$, it follows that there is a constant $A(t)$ such that $e^{-t \lambda_{n}} \leq A(t)$ for all $n$. Thus:

$$
\sum_{n}\left|e^{-t \lambda_{n}} a_{n}\right|^{2} \leq A(t)^{2} \sum_{n}\left|a_{n}\right|^{2}
$$

and the heat operator $e^{-t \Delta_{P}^{+}}$is a bounded operator on $L_{2}\left(M, E^{+}\right)$, with operator norm $\leq A(t)$. It is self-adjoint since $\Delta_{P}^{+}$is formally self adjoint, and smooth functions are dense in $L_{2}\left(M, E^{+}\right)$.

To see that it is infinitely smoothing, note that by the Corollary 6.2.3 (Garding inequality) applied to the elliptic operator $Q:=\left(\Delta_{P}^{+}\right)^{k}$ (which is of order $2 k d$ ), we have that the Sobolev $2 k d$-norm is given by:

$$
\left\|e_{n}\right\|_{2 k d}^{2}=\left\|Q e_{n}\right\|_{0}^{2}+\left\|e_{n}\right\|_{0}^{2}=\left(\lambda_{n}^{k}+1\right)
$$

Again, since $t>0$ and $\lambda_{n} \geq C n^{\delta}$, it easily follows that

$$
\sum_{n=0}^{\infty} e^{-2 t \lambda_{n}}\left(\lambda_{n}^{k}+1\right) \leq B_{k}(t)<\infty
$$

Thus, for the partial sum $g_{N}:=\sum_{n=0}^{N} e^{-t \lambda_{n}} a_{n} e_{n}$, we will have

$$
\begin{aligned}
\left\|g_{N}\right\|_{2 k d} & \leq \sum_{n=0}^{N} e^{-t \lambda_{n}}\left|a_{n}\right|\left\|e_{n}\right\|_{2 k d}=\sum_{n=0}^{N} e^{-t \lambda_{n}}\left(\lambda_{n}^{k}+1\right)^{1 / 2}\left|a_{n}\right| \\
& \leq\left(\sum_{n=0}^{N} e^{-2 t \lambda_{n}}\left(\lambda_{n}^{k}+1\right)\right)^{1 / 2}\left(\sum_{n=0}^{N}\left|a_{n}\right|^{2}\right)^{1 / 2} \leq B_{k}(t)^{1 / 2}\|f\|_{0}
\end{aligned}
$$

From which it follows that $e^{-t \Delta_{P}^{+}} f \in H_{2 k d}\left(M, E^{+}\right)$for all $k$, which implies that it is smooth by the Sobolev Lemma (iv) of Proposition 4.2.2.

Since $e^{-t \Delta_{P}^{+}}$is infinitely smoothing, it is in $\Psi^{d-1}(M)$ for all $d$, and a bounded operator $H_{d}\left(M, E^{+}\right)$into $H_{1}\left(M, E^{+}\right)$for all $d$. By Rellich's Lemma (vi) of 4.2.2, since the inclusion $H_{1} \subset H_{0}=L_{2}\left(M, E^{+}\right)$is compact,

$$
e^{-t \Delta_{P}^{+}}: H_{d}\left(M, E^{+}\right) \rightarrow L_{2}\left(M, E^{+}\right)
$$

is a compact operator for all $d$.

Proposition 10.1.3 (Some facts about the Heat Operator). In the setting of the previous proposition, we have the following:
(i): The for $f \in L_{2}\left(M, E^{+}\right)$, we let $f_{\mathcal{H}^{+}}$denote the orthogonal projection $\pi(f)$ to the finite dimensional $\Delta^{+}$-harmonic space ker $\Delta^{+}=\oplus_{i \geq 0} \operatorname{ker} \Delta_{P}^{2 i}$ by the Kodaira-Hodge decomposition of (ii) in 9.5.2. Then

$$
\lim _{t \rightarrow 0} e^{-t \Delta^{+}} f=f ; \quad \lim _{t \rightarrow \infty} e^{-t \Delta^{+}} f=f_{\mathcal{H}^{+}} \quad \text { for all } f \in L_{2}\left(M, E^{+}\right)
$$

where the convergence, of course, is in the $L_{2}$-norm.
(ii): If $f \in C^{\infty}\left(M, E^{+}\right)$, then $e^{-t \Delta^{+}} f$ converges to $f_{\mathcal{H}^{+}}$as $t \rightarrow \infty$ and to $f$ as $t \rightarrow 0$ in the norm $\|-\|_{k, \infty}$ for all $k$.
(iii): For $t>0$, there is a smooth integral kernel:

$$
k_{t}^{+} \in C^{\infty}\left(M \times M, \operatorname{hom}_{\mathbb{C}}\left(\pi_{2}^{*} E^{+}, \pi_{1}^{*} E^{+}\right)\right)
$$

(where $\pi_{1}, \pi_{2}$ are the first and second projections of $M \times M$ to $M$ ) satisfying:

$$
\left(e^{-t \Delta^{+}} f\right)(x)=\int_{M} k_{t}^{+}(x, y) f(y) d V(y) \text { for all } f \in C^{\infty}\left(M, E^{+}\right)
$$

(iv): For $t \in(0, \infty)$, the sum $\sum_{i=0}^{\infty} e^{-t \lambda_{i}}$, called the trace of the heat operator for obvious reasons, and denoted by $\operatorname{tr} e^{-t \Delta^{+}}$is given by the integral:

$$
\operatorname{tr} e^{-t \Delta^{+}}=\int_{M} \operatorname{tr} k_{t}^{+}(x, x) d V(x)
$$

Analogous facts obtain for $e^{-t \Delta^{-}}$.

Proof: Let $e_{n}$ be $L_{2}$-orthogonal eigensections for $\Delta^{+}$corresponding to the eigenvalues $\lambda_{n}$. Assume that in our non-decreasing arrangement of eigenvalues $\lambda_{1}=\lambda_{2}=\ldots \lambda_{p}=0$, so that $\left\{e_{n}\right\}_{n=1}^{p}$ is an orthonormal basis for $\mathcal{H}^{+}$. Also $\lambda_{p+1}>0$. Now expand $f \in L_{2}\left(M, E^{+}\right)$as $f=f_{\mathcal{H}^{+}}+\sum_{n \geq p+1} a_{n} e_{n}$. Then $e^{-t \Delta^{+}} f=$ $f_{\mathcal{H}^{+}}+\sum_{n \geq p+1} e^{-t \lambda_{n}} a_{n} e_{n}$, and:

$$
\left\|e^{-t \Delta^{+}} f-f_{\mathcal{H}^{+}}\right\|^{2}=\sum_{n \geq p+1} e^{-2 t \lambda_{n}}\left|a_{n}\right|^{2} \leq e^{-2 t \lambda_{p+1}} \sum_{n \geq p+1}\left|a_{n}\right|^{2} \leq e^{-t \lambda_{p+1}}\|f\|^{2}
$$

which clearly shows, since $\lambda_{p+1}>0$, that $\lim _{t \rightarrow \infty} e^{-t \Delta^{+}} f=f_{\mathcal{H}^{+}}$and the second assertion of (i) follows.
For the first assertion, note that:

$$
\left\|e^{-t \Delta^{+}} f-f\right\|^{2}=\sum_{n \geq p+1}\left(e^{-t \lambda_{n}}-1\right)^{2}\left|a_{n}\right|^{2}
$$

Now, given any $\epsilon>0$, choose an $N>p+1$ such that $\sum_{n \geq N+1}\left|a_{n}\right|^{2} \leq \epsilon$. Also since $\lambda_{n} \geq \lambda_{p+1}>0$ for $n \geq p+1$ by the Proposition 10.1.1, we can choose $\eta>0$ so that $\left(e^{-t \lambda_{n}}-1\right)^{2} \leq \epsilon$ for $t \in(0, \eta)$ and all the finitely many $n$ satisfying $p+1 \leq n \leq N$. Then we estimate:

$$
\begin{aligned}
\sum_{n \geq p+1}\left(e^{-t \lambda_{n}}-1\right)^{2}\left|a_{n}\right|^{2} & \leq \sum_{p+1 \leq n \leq N}\left(e^{-t \lambda_{n}}-1\right)^{2}\left|a_{n}\right|^{2}+C \sum_{n \geq N+1}\left|a_{n}\right|^{2} \\
& \leq \epsilon\|f\|^{2}+C \epsilon \text { for } 0<t<\eta
\end{aligned}
$$

which proves that $\lim _{t \rightarrow 0} e^{-t \Delta^{+}} f=f$ in $L_{2}$ and the first assertion of (i) follows.
Now we prove (ii). In view the Sobolev Embedding Theorem (iv) of Proposition 4.2.2, and the Corollary 6.2.3 (Garding-Friedrichs inequality applied to the elliptic operator $\Delta^{+k}$ of order $2 k d$ ), it is enough to show that that for $f \in C^{\infty}\left(M, E^{+}\right)\left(\right.$contained in $H_{2 k d}\left(M, E^{+}\right)$for all $\left.k \geq 1\right)$ :
(a): $\left(\Delta^{+}\right)^{k} e^{-t \Delta^{+}} f$ converges to $\left(\Delta^{+}\right)^{k} f$ in $L_{2}\left(M, E^{+}\right)$as $t \rightarrow 0$ (resp. converges to $\left(\Delta^{+}\right)^{k} f_{\mathcal{H}^{+}}$, which incidentally is zero for $k \geq 1$, as $t \rightarrow \infty)$ and,
(b): $e^{-t \Delta^{+}} f$ converges to $f$ in $L_{2}\left(M, E^{+}\right)$as $t \rightarrow 0$ (resp. converges to $f_{\mathcal{H}^{+}}$in $L_{2}\left(M, E^{+}\right)$as $\left.t \rightarrow \infty\right)$.

The statement (b) follows from (i) above. For the statement (a), note that $\left(\Delta^{+}\right)^{k} e^{-t \Delta^{+}} f=e^{-t \Delta^{+}}\left(\Delta^{+}\right)^{k} f$, and $\Delta^{+k}\left(f_{\mathcal{H}^{+}}\right)=\left(\Delta^{+k} f\right)_{\mathcal{H}^{+}}$, and thus (a) follows by applying (i) to the section $\Delta^{+k} f \in L_{2}\left(M, E^{+}\right)$. This proves (ii).

Now we give the construction for $k_{t}^{+}$. For the smooth eigensection $e_{n}$ of $E^{+}$corresponding to the eigenvalue $\lambda_{n}$, denote by $e_{n}^{*}$ the section of $E^{+*}=\operatorname{hom}\left(E^{+}, \mathbb{C}\right)$ defined by $e_{n}^{*}(x)(w)=\left\langle w, e_{n}(x)\right\rangle_{x}$ for $w \in E_{x}^{+}$(remember Hermitian metrics are linear in the first slot, and conjugate linear in the second!). Then $e_{n}^{*}(y) \otimes e_{n}(x)$ becomes a smooth section of $\operatorname{hom}\left(\pi_{2}^{*} E^{+}, \pi_{1}^{*} E^{+}\right)$, its value at a an element $v \in E_{y}^{+}=\left(\pi_{2}^{*} E^{+}\right)_{(x, y)}$ being the element $e_{n}^{*}(y)(v) e_{n}(x)=\left\langle v, e_{n}(y)\right\rangle e_{n}(x) \in E_{x}^{+}=\left(\pi_{1}^{*} E^{+}\right)_{(x, y)}$.

Define the formal sum:

$$
k_{t}(x, y)=\sum_{i=0}^{\infty} e^{-t \lambda_{n}}\left(e_{n}^{*}(y) \otimes e_{n}(x)\right)
$$

Note that $L_{2}\left(M \times M, \operatorname{hom}\left(\pi_{2}^{*} E^{+}, \pi_{1}^{*} E^{+}\right)\right)$has a canonical $L_{2}$-inner product arising out of the natural tensor product Hermitian metric on the bundle $\operatorname{hom}\left(\pi_{2}^{*} E^{+}, \pi_{1}^{*} E^{+}\right)$. The corresponding global inner product on $M \times M$ (with respect to the volume element of the product Riemannian metric) has the orthonormal basis $\left\{e_{m}^{*}(y) \otimes e_{n}(x)\right\}$. So the series above certainly converges in $L_{2}$-norm, by the estimate $\lambda_{n} \geq C n^{\delta}$ of Proposition 10.1.1. To show that this kernel is a smooth section on $M \times M$, we apply the elliptic operators $\Delta_{y}^{+k} \times \Delta_{x}^{+j}$ for arbitrary $j$ and $k$, and note that the differentiated series will have coefficients $\lambda_{n}^{k+j} e^{-t \lambda_{n}}$, to which again $\lambda_{n} \geq C n^{\delta}$ may be applied, to show that this differentiated series is again $L_{2}$ over $M \times M$. Now appeal to the Sobolev Lemma and Garding-Friedrichs as always.

To see it is the required kernel, we compute its effect on each $e_{m}$ :

$$
\begin{array}{r}
\int_{y \in M} k_{t}(x, y) e_{m}(y) d V(y)=\sum_{n=0}^{\infty} e^{-t \lambda_{n}} \int_{y \in M} e_{n}^{*}(y) \otimes e_{n}(x)\left(e_{m}(y)\right) d V(y) \\
=\sum_{n=0}^{\infty} e^{-t \lambda_{n}} e_{n}(x) \int_{y \in M}\left\langle e_{n}(y),\left(e_{m}(y)\right\rangle d V(y)=\sum_{n=0}^{\infty} e^{-t \lambda_{n}} e_{n}(x)\left(e_{n}, e_{m}\right)=e^{-t \lambda_{m}} e_{m}(x)\right.
\end{array}
$$

since $\left(e_{n}, e_{m}\right)=\delta_{n m}$ by $L_{2}$-orthornormality of $e_{n}$ 's. This shows that the integral operator defined by $k_{t}$ has the same effect on each $e_{m}$ as the heat operator $e^{-t \Delta^{+}}$, and the two operators are therefore the same on $L_{2}\left(M, E^{+}\right)$. This proves (iii).

To see (iv), we first define what we mean by $\operatorname{tr} k_{t}^{+}(x, x) . k_{t}$ is a smooth section of the bundle $\operatorname{hom}_{\mathbb{C}}\left(\pi_{2}^{*} E^{+}, \pi_{1}^{*} E^{+}\right)$. The maps $\pi_{1}$ and $\pi_{2}$ agree on the diagonal, and indeed if one identifies the diagonal $\Delta_{M}$ inside $M \times M$ with $M$ via the map $(x, x) \mapsto x$, the bundles $\pi_{2}^{*} E^{+}$and $\pi_{1}^{*} E^{+}$both get identified with the bundle $E^{+}$. Thus restricting the smooth section $k_{t}^{+}$to the diagonal gives the smooth section, denoted by $k_{t}^{+}(x, x)$, of the bundle $\operatorname{hom}_{\mathbb{C}}\left(E^{+}, E^{+}\right)$. On this bundle there is the natural trace map:

$$
\begin{aligned}
\operatorname{tr}: \operatorname{hom}_{\mathbb{C}}\left(E^{+}, E^{+}\right) & \rightarrow \mathbb{C} \\
T_{x} & \mapsto \sum_{i}\left\langle T_{x}\left(f_{i}\right), f_{i}\right\rangle_{x}
\end{aligned}
$$

where $f_{i}$ is any $\langle-,-\rangle_{x}$ orthonormal basis of $E_{x}^{+}$, (viz. it is the invariant trace of $T_{x}: E_{x}^{+} \rightarrow E_{x}^{+}$).
Now we simply calculate, for $x \in M$, and $f_{i}$ some $\langle-,-\rangle_{x}$-orthonormal basis of $E_{x}^{+}$:

$$
\begin{aligned}
\operatorname{tr} k_{t}^{+}(x, x) & =\sum_{i} \sum_{n=0}^{\infty} e^{-t \lambda_{n}}\left\langle\left(e_{n}^{*}(x) \otimes e_{n}(x)\right)\left(f_{i}\right), f_{i}\right\rangle_{x} \\
& =\sum_{n=0}^{\infty} e^{-t \lambda_{n}} \sum_{i}\left\langle\left\langle f_{i}, e_{n}(x)\right\rangle e_{n}(x), f_{i}\right\rangle_{x} \\
& =\sum_{n=0}^{\infty} e^{-t \lambda_{n}} \sum_{i}\left|\left\langle e_{n}(x), f_{i}\right\rangle_{x}\right|^{2}=\sum_{n=0}^{\infty} e^{-t \lambda_{n}}\left\|e_{n}(x)\right\|_{x}^{2}
\end{aligned}
$$

which implies that:

$$
\int_{M} \operatorname{tr} k_{t}^{+}(x, x) d V(x)=\sum_{n=0}^{\infty} e^{-t \lambda_{n}} \int_{M}\left\|e_{n}(x)\right\|_{x}^{2} d V(x)=\sum_{n=0}^{\infty} e^{-t \lambda_{n}}\left(e_{n}, e_{n}\right)=\sum_{n=0}^{\infty} e^{-t \lambda_{n}}
$$

since $e_{n}$ 's form an orthonormal basis with respect to the global inner product $(-,-)$. This proves (iv), and the proposition follows.

Remark 10.1.4. The proof of assertion (ii) of the foregoing proposition reflects a classical fact about the heat operator $e^{-t \Delta^{+}}$, which is that it starts with an arbitrarily irregular $f$ (a distribution, i.e. in some $H_{s}\left(M, E^{+}\right)$) at $t=0$, and makes it smooth at any positive time $t>0$. Indeed as $t \rightarrow \infty$, it converts the irregular $f$ into its smooth harmonic part $f_{\mathcal{H}^{+}}$. Thus it time-evolves the irregular initial data $f$ into a smooth section for any $t>0$, and into its smooth harmonic part as $t \rightarrow \infty$.

Remark 10.1.5. We explicitly constructed the kernel for the heat operator $e^{-t \Delta^{+}}$. However, it is a fact that an operator (defined on distributional sections $\mathcal{D}^{\prime}(M, E)$ ) on a compact Riemannian manifold is infinitely smoothing iff it is given by an integral operator with a smooth integral kernel. For convenience's sake, let us consider an operator $K$ which maps $H_{s}(M, \mathbb{C}) \rightarrow L_{2}(M, \mathbb{C})$ as a bounded operator, for all $s$, and whose image is contained in $C^{\infty}(M, \mathbb{C})$. We know that on $\mathbb{R}^{n}$, the Dirac distribution $\delta_{x}$ is a compactly supported distribution lying in $H_{-k}\left(\mathbb{R}^{n}\right)$ for all $k>n / 2$ (see the Corollary 3.2.2). If $M$ is of dimension $n$, since the support of $\delta_{x}$ is $x$, it becomes an element of $H_{-k}(M)$ for all $n>k / 2$ (by using a partition of unity definition of $H_{s}(M)$ ). Thus $K\left(\delta_{x}\right)$ is a smooth function. Define:

$$
k(y, x):=K\left(\delta_{x}\right)(y)
$$

One now has to verify that this is the required integral kernel. For the converse, one has to verify that integral operators with smooth integral kernels on a compact manifold are infinitely smoothing, by differentiating under the integral sign using compactness of $M$, or using the Sobolev Embedding Theorem coupled with clever uses of integral inequalities.

As we have remarked earlier, integral operators with smooth integral kernels do not give rise to infinitely smoothing operators on non-compact manifolds. For example, the Fourier transform on $\mathbb{R}$ is an integral operator with smooth kernel $e^{-i \xi \cdot x}$, but converts a smooth function like $\left(1+x^{2}\right)^{-1}$ into a non-smooth function.

Proposition 10.1.6 (Facts about the heat kernel).
(i): The section $k_{t}^{+}(x, y)$ defined in (iii) of the Proposition 10.1.3 satisfies the pointwise adjointness formula:

$$
\left\langle k_{t}^{+}(x, y) v, w\right\rangle_{x}=\left\langle v, k_{t}^{+}(y, x) w\right\rangle_{y} \text { for } v \in\left(\pi_{2}^{*} E^{+}\right)_{(x, y)}=E_{y}^{+}, w \in\left(\pi_{1}^{*} E^{+}\right)_{(x, y)}=E_{x}^{+}
$$

(ii): $k_{t}^{+}(x, y)$ satisfies the heat equations

$$
\left(\frac{\partial}{\partial t}+\Delta_{x}^{+}\right) k_{t}^{+}(x, y)=0=\left(\frac{\partial}{\partial t}+\left(\Delta^{+}\right)_{y}^{\vee}\right) k_{t}^{+}(x, y) \text { for } t \in(0, \infty),(x, y) \in M \times M
$$

where

$$
\left(\Delta^{+}\right)^{\vee}: C^{\infty}\left(M, E^{+*}\right) \rightarrow C^{\infty}\left(M, E^{+*}\right)
$$

is the pointwise dual of $\Delta^{+}$with respect to the Hermitian metric $\langle-,-\rangle$on $E^{+}$.
(iii): If $f \in L_{2}\left(M, E^{+}\right)$is a square integrable section, we have seen in the Proposition 10.1.2 that $e^{-t \Delta^{+}} f$ is smooth in $x$. It is also smooth in $t$ for $t \in(0, \infty)$, and if we define $F(x, t):=e^{-t \Delta^{+}} f$, the $F$ satisfies:

$$
\begin{aligned}
\left(\frac{\partial}{\partial t}+\Delta^{+}\right) F(x, t) & =0 \text { for } t \in(0, \infty), x \in M \\
F(x, 0) & :=\lim _{t \rightarrow 0} F(x, t)=f
\end{aligned}
$$

There are completely analogous statements for $k_{t}^{-}$and $\Delta^{-}$.

Proof: To see (i), we note that for $v \in E_{y}^{+}, w \in E_{x}^{+}$:

$$
\left\langle\left(e_{n}^{*}(y) \otimes e_{n}(x)\right)(v), w\right\rangle_{x}=\left\langle\left\langle v, e_{n}(y)\right\rangle_{y} e_{n}(x), w\right\rangle_{x}=\left\langle v, e_{n}(y)\right\rangle_{y}\left\langle e_{n}(x), w\right\rangle_{x}
$$

and interchanging the roles of $x$ and $y, v$ and $w$, we have:

$$
\left\langle\left(e_{n}^{*}(x) \otimes e_{n}(y)(w), v\right\rangle_{y}=\left\langle w, e_{n}(x)\right\rangle_{x}\left\langle e_{n}(y), v\right\rangle_{y}\right.
$$

Since the right hand sides of the two equations above are complex, conjugates of each other, we have:

$$
\left\langle\left(e_{n}^{*}(y) \otimes e_{n}(x)\right)(v), w\right\rangle_{x}=\left\langle v, e_{n}^{*}(x) \otimes e_{n}(y)(w)\right\rangle_{y}
$$

from which (i) follows by multiplying by $e^{-t \lambda_{n}}$ and summing over $n$.
To see (ii), note that since the series for $k_{t}(x, y)$ is absolutely and uniformly convergent in both variables, as are the differentiated series with respect to $\partial_{t}=\frac{\partial}{\partial t}$ and $\Delta_{x}^{+}$and $\Delta_{y}^{+}$(from the eigenvalue estimate $\lambda_{n} \geq C n^{\delta}$ ), we can apply these operators term by term. Hence:

$$
\begin{aligned}
\partial_{t}\left(\sum_{n} e^{-t \lambda_{n}} e_{n}^{*}(y) \otimes e_{n}(x)\right. & =\sum_{n}\left(-\lambda_{n} e^{-t \lambda_{n}} e_{n}^{*}(y) \otimes e_{n}(x)\right. \\
-\sum_{n} e^{-t \lambda_{n}} e_{n}^{*}(y) \otimes \Delta_{x}^{+} e_{n}(x) & =-\Delta_{x}^{+}\left(k_{t}(x, y)\right)
\end{aligned}
$$

Also, from the equation $\Delta_{y}^{+} e_{n}(y)=\lambda_{n} e_{n}(y)$, one finds that for the pointwise adjoint operator $\Delta_{y}^{+\vee}$ defined by the adjoint formula:

$$
\left\langle\Delta^{+\vee} \psi, f\right\rangle=\left\langle\psi, \Delta^{+} f\right\rangle, \quad f \in C^{\infty}\left(M, E^{+}\right), \quad \psi \in C^{\infty}\left(M, E^{+*}\right)
$$

one easily finds that $\left(\Delta^{+}\right)^{\vee} e_{n}^{*}=\lambda_{n} e_{n}^{*}$, and the second formula of (ii) follows as well.
To see (iii) note that if $f \in L_{2}\left(M, E^{+}\right)$, we may write $f=\sum_{n} a_{n} e_{n}$, with $\sum_{n}\left|a_{n}\right|^{2}=\|f\|^{2}<\infty$. Furthermore, $F(x, t)=\sum_{n} e^{-t \lambda_{n}} a_{n} e_{n}$ is a series which lies in $H_{s}\left(M, E^{+}\right)$for all $s$, and converges in $\|-\|_{s}$ for each $s$ (meaning the Sobolev $s$-norm of the tails $\sum_{n \geq N} e^{-t \lambda_{n}} a_{n} e_{n}$ converges to 0 for all $s$, by the facts that $\lambda_{n} \geq C n^{\delta}$, and $\left.\left\|e_{n}\right\|_{2 k d}^{2}=\lambda_{n}^{k}+1\right)$. Hence the series on the right converges in $\|-\|_{\infty, k}$ for all $k$, by the Sobolev Embedding Theorem (iv) of 4.2.2. Hence if one applies $\partial_{t}$, or $\Delta^{+}$term-by-term to this series, the resulting series converge to $\partial_{t}(F(x, t))$ and $\Delta^{+} F(x, t)$ respectively. However, upon term by term differentiation we have:

$$
\partial_{t}\left(e^{-t \Delta^{+}} f\right)=-\sum_{n} \lambda_{n} e^{-t \lambda_{n}} a_{n} e_{n}=-\sum_{n} e^{-t \lambda_{n}} a_{n} \Delta^{+} e_{n}=\Delta^{+}\left(e^{-t \Delta^{+}} f\right)
$$

since $\Delta^{+}$and $e^{-t \Delta^{+}}$commute. This proves that $\partial_{t} F(x, t)+\Delta^{+} F(x, t)=0$. The fact that $\lim _{t \rightarrow 0} F(x, t)=f$ follows from (i) of Proposition 10.1.3. The proposition follows.

An analogue of (iii) can be proved for $f \in H_{s}\left(M, E^{+}\right)$and any $s$, (i.e. for all distributional sections $f \in H_{-\infty}\left(M, E^{+}\right)=\mathcal{D}^{\prime}\left(M, E^{+}\right)$, but we omit the proof. It is completely analogous, because $f$ can still be expanded in a Fourier series $\sum_{n} a_{n} e_{n}$. One needs to note that $e_{n}$ need no longer be orthonormal in $\|-\|_{2 k d}$, but we still have $\left(e_{n}, e_{m}\right)_{2 k d}=\left(\lambda_{n}^{k}+1\right) \delta_{n m}$ for all $k \neq 0$ (by proving the analogue of Corollary 6.2.3 for $k \leq 0$, which in turn stems from the duality of $H_{2 k d}$ and $H_{-2 k d}$ from (iii) of Proposition 4.2.2).
10.2. An integral formula for the index of $D^{+}$. The following proposition is the key to the entire heatequation approach for the index theorem.

Theorem 10.2.1 (McKean-Singer). Let $M$ be a compact Riemannian manifold, and $\mathcal{P}$ an elliptic complex on $M$. Let:

$$
D^{ \pm}: C^{\infty}\left(M, E^{ \pm}\right) \rightarrow C^{\infty}\left(M, E^{\mp}\right)
$$

be the corresponding Dirac operators, as in Definition 9.6.1, and let

$$
k_{t}^{ \pm}(x, y) \in C^{\infty}\left(M \times M, \operatorname{hom}_{\mathbb{C}}\left(\pi_{2}^{*} E^{ \pm}, \pi_{1}^{*} E^{ \pm}\right)\right)
$$

denote the heat kernels of the heat evolution operators $e^{-t \Delta \pm}$ respectively, as in (iii) of the Proposition 10.1.3. Then:

$$
\operatorname{ind} D^{+}=\int_{M}\left(\operatorname{tr} k_{t}^{+}(x, x)-\operatorname{tr} k_{t}^{-}(x, x)\right) d V(x)=-\operatorname{ind} D^{-}
$$

In particular, the quantity on the right is an integer independent of $t$.

Proof: Let $\lambda_{n} \geq 0$ and $\mu_{m} \geq 0$ be the eigenvalues of the two Laplacians $\Delta^{+}=D^{-} D^{+}$and $\Delta^{-}=D^{+} D^{-}$ respectively. Let the eigensections of $\Delta^{+}$be denoted $e_{n}$, which are orthonormal in $L_{2}\left(M, E^{+}\right)$with respect to its $L_{2}$ - inner product, which we denote by $(-,-)_{+}\left(\right.$with $(-,-)_{-}$denoting the $L_{2}$-inner product on $\left.L_{2}\left(M, E^{-}\right)\right)$.

We now note that if $e_{n}$ is an eigensection of $\Delta^{+}$with eigenvalue $\lambda_{n}$, we have:

$$
\Delta^{-} D^{+} e_{n}=\left(D^{+} D^{-}\right) D^{+} e_{n}=D^{+}\left(D^{-} D^{+}\right) e_{n}=D^{+}\left(\Delta^{+} e_{n}\right)=\lambda_{n} D^{+} e_{n}
$$

so that $D^{+} e_{n}$, if non-zero, is an eigensection of $\Delta^{-}$corresponding to the same eigenvalue $\lambda_{n}$.
Furthermore, for all $n, m$, by the fact that $\lambda_{n}$ and $\lambda_{m}$ are both real, and the adjointness of $D^{+}$and $D^{-}$, we have:

$$
\lambda_{n}\left(e_{n}, e_{m}\right)_{+}=\left(D^{-} D^{+} e_{n}, e_{m}\right)_{+}=\left(D^{+} e_{n}, D^{+} e_{m}\right)_{-}=\left(e_{n}, D^{-} D^{+} e_{m}\right)_{+}=\lambda_{m}\left(e_{n}, e_{m}\right)_{+}
$$

which implies in view of the foregoing that:
(i): If $\lambda_{n} \neq 0$ (i.e. $\lambda_{n}>0$ ), the section $D^{+} e_{n}$ is a non-zero eigensection of $\Delta^{-}$corresponding to the same eigenvalue $\lambda_{n}$ as $e_{n}$.
(ii): For $n \neq m$ we have $D^{+} e_{m}$ orthogonal to $D^{+} e_{m}$.

Similar facts obtain for the eigensections $f_{n} \in C^{\infty}\left(M, E^{-}\right)$of the other Laplacian $\Delta^{-}$. From this it follows that for $\lambda_{n} \neq 0$ and $\mu_{m} \neq 0, D^{+}$maps the finite-dimensional $\lambda_{n}$-eigenspace of $\Delta^{+}$isomorphically into a subspace of the $\lambda_{n}$-eigenspace of $\Delta^{-}$, and $D^{-}$similarly maps the finite-dimensional $\mu_{m}$-eigenspace of $\Delta^{-}$ isomorphically into a subspace of the $\mu_{m}$-eigenspace of $\Delta^{+}$. It follows that the non-zero eigenvalues $\lambda_{n}>0$ of $\Delta^{+}$are in bijective correspondence with the non-zero eigenvalues $\mu_{m}>0$ of $\Delta^{-}$, and also occur with exactly the same multiplicity. Thus:

$$
\operatorname{tr} e^{-t \Delta^{+}}-\operatorname{tr} e^{-t \Delta^{-}}=\sum_{n} e^{-t \lambda_{n}}-\sum_{m} e^{-t \mu_{m}}=\sum_{\lambda_{n}=0} 1-\sum_{\mu_{m}=0} 1=\operatorname{dim} \operatorname{ker} \Delta^{+}-\operatorname{dim} \operatorname{ker} \Delta^{-}
$$

But the left hand side of this equation is precisely:

$$
\int_{M}\left(\operatorname{tr} k_{t}^{+}(x, x)-\operatorname{tr} k_{t}^{-}(x, x)\right) d V(x)
$$

by (iv) of Proposition 10.1.3, and the right hand side of the equation is ind $D^{+}=-$ind $D^{-}$by the proof of (iii) in Proposition 9.6.2.

The last assertion is clear in view of the fact that ind $D^{+}$is independent of $t$. The theorem follows.
Now, in the sequel, the main aim is to identify the integrand

$$
\operatorname{str} k_{t}(x, x):=\operatorname{tr} k_{t}^{+}(x, x)-\operatorname{tr} k_{t}^{-}(x, x)
$$

called the supertrace of the heat evolution operator. This is impossible in full generality. However, one can do what is called an asymptotic expansion in powers of $t^{1 / d}$ (where $d$ is the order of the differential operator $D^{+}$) for small times $t$, and using the fact that the left hand side is independent of $t$, compute just the coefficient of $t^{0}$ (the constant term) in this symptotic expansion. That such an asymptotic expansion exists in general is proved in Gilkey. However, since we shall be interested only in four specific elliptic complexes (the de-Rham, TwistedDolbeault, Signature and Spin complexes), for each of which the corresponding Laplacian $\Delta^{+}=D^{-} D^{+}$is a generalised Laplacian, i.e. a second order operator whose leading symbol is the same as that of the classical Laplace-Beltrami operator, we will concentrate only on such elliptic complexes.

## 11. Fundamental solutions

To motivate whatever follows, we need to construct the heat kernel for the Laplacian $\Delta=-\sum_{j} \partial_{j}^{2}$ on $\mathbb{R}^{n}$. Our assertion of the existence of a heat kernel in Proposition 10.1.3 doesn't quite apply, since $\mathbb{R}^{n}$ is non-compact, and does not have a discrete spectrum. But fortunately, one can explicitly write down the heat kernel (or Gauss kernel as it is sometimes known) in the case of $\mathbb{R}^{n}$.

### 11.1. The Euclidean heat kernel.

Proposition 11.1.1 (Euclidean heat kernel). For $x, y \in \mathbb{R}^{n}$, define the function:

$$
k_{t}(x, y)=(4 \pi t)^{-n / 2} e^{-|x-y|^{2} / 4 t}
$$

(i): $k_{t}(x, y)$ is symmetric in $x$ and $y$, and is a fundamental solution to the heat equation, viz.,

$$
\left(\partial_{t}+\Delta_{x}\right) k_{t}(x, y)=0=\left(\partial_{t}+\Delta_{y}\right) k_{t}(x, y)
$$

(ii): For $f \in L_{2}\left(\mathbb{R}^{n}\right)$ the function:

$$
F(x, t)=e^{-t \Delta} f:=\int_{\mathbb{R}^{n}} k_{t}(x, y) f(y) d y
$$

is a smooth function of both $t$ and $x$, and satisfies:

$$
\left(\partial_{t}+\Delta\right) F(x, t)=0
$$

(iii): Let $y \in \mathbb{R}^{n}$, and let $\delta_{y}$ denote the Dirac distribution at $y$. Then there is a smooth function $w(-, t) \in$ $\mathcal{S}\left(\mathbb{R}^{n}\right)$ such that:

$$
\text { (a) }\left(\partial_{t}+\Delta\right) w(x, t)=0 \text { for all } x \in \mathbb{R}^{n}, t>0
$$

and

$$
\text { (b) } \quad \lim _{t \rightarrow 0}(f, w(-, t))=f(y) \text { for all } f \in \mathcal{S}\left(\mathbb{R}^{n}\right)
$$

This $w(-, t)$ is called a fundamental solution of the heat equation on $\mathbb{R}^{n}$ with pole at $y$, and is uniquely determined by the conditions (a) and (b).
(Caution: in this proposition, $d y$ denotes Euclidean volume element $d y=d y_{1} \ldots d y_{n}$, and is related to the earlier volume element $d V(y)$ of $\S 1$ by $d V(y)=(2 \pi)^{-n / 2} d y$.)

Proof: Direct differentiation yields that:

$$
\begin{aligned}
\partial_{t}\left(t^{-n / 2} e^{-|x-y|^{2} / 4 t}\right) & =\left(t^{-n / 2} \frac{|x-y|^{2}}{4 t^{2}}+(-n / 2) t^{-n / 2-1}\right) e^{-|x-y|^{2} / 4 t} \\
& =\left(\frac{-n}{2}+\frac{|x-y|^{2}}{4 t}\right) t^{-n / 2-1} e^{-|x-y|^{2} / 4 t} \\
\Delta_{x}\left(t^{-n / 2} e^{-|x-y|^{2} / 4 t}\right) & =-t^{-n / 2} \sum_{i} \partial_{x, i}\left(-\frac{\left(x_{i}-y_{i}\right)}{2 t} e^{-|x-y|^{2} / 4 t}\right)=t^{-n / 2} \sum_{i}\left(\frac{1}{2 t}-\frac{\left(x_{i}-y_{i}\right)^{2}}{4 t^{2}}\right) e^{-|x-y|^{2} / 4 t} \\
& =t^{-n / 2-1}\left(\frac{n}{2}-\frac{|x-y|^{2}}{4 t}\right) e^{-|x-y|^{2} / 4 t}
\end{aligned}
$$

from which the statement (i) follows. For the second, note that if we denote the function:

$$
\rho_{t}(x)=(2 t)^{-n / 2}\left(e^{-|x|^{2} / 4 t}\right)
$$

then $F(x, t)=\rho_{t} * f$, where the convolution is the same as the one introduced in $\S 1$ (i.e. in the space variables, with respect to the volume $\left.d V(y)=(2 \pi)^{-n / 2} d y_{1} \ldots d y_{n}\right)$. Since $\rho_{t}$ is in the Schwartz class (since $t>0$ ), and $f \in L_{2}\left(\mathbb{R}^{n}\right)=H_{0}\left(\mathbb{R}^{n}\right)$ implies $f$ is a tempered distribution, it follows that the convolution $F(x, t)$ is smooth in the space variable $x$ by the Lemma 1.4.7. and also that $\Delta_{x} F(x, t)=\left(\Delta \rho_{t}\right) * f$.

In the time variable, one uses that $\partial_{t} \rho_{t}(x)=\left(-n / 2+|x|^{2} / 4 t\right) t^{-1} \rho_{t}(x)$ and the Dominated convergence theorem to take $\partial_{t}$ under the integral sign and get

$$
\partial_{t}(F(x, t))=\left(\partial_{t} \rho_{t}\right) * f
$$

Hence we have:

$$
\left(\partial_{t}+\Delta_{x}\right)(F(x, t))=\left(\partial_{t}+\Delta\right)\left(\rho_{t}\right) * f=\int_{\mathbb{R}^{n}}\left(\partial_{t}+\Delta_{y}\right) k_{t}(x, y) f(y) d y=0
$$

by applying (i). This proves (ii).

To see (iii), note that the Dirac distribution $\delta_{y}$ is a tempered distribution (as we remarked in Example 1.3.5, it is a compactly supported distribution), and so taking the convolution with the Schwartz class function $\rho_{t}$ for $t>0$

$$
w(x, t):=\rho_{t} * \delta_{y}
$$

gives a smooth function (see Proposition 1.4.7). To see that it is in the Schwartz class, note that its Fourier transform is $\widehat{\rho}_{t}(\xi) e^{-i \xi \cdot y}$, which is in $\mathcal{S}\left(\mathbb{R}^{n}\right)$ since $\widehat{\rho_{t}}$ is in $\mathcal{S}\left(\mathbb{R}^{n}\right)$. If one writes down the formula for the convolution of distributions, we find:

$$
w(x, t)=\delta_{y}\left(\rho_{t}^{x}\right)=\rho_{t}^{x}(y)=(2 t)^{-n / 2} e^{-|x-y|^{2} / 4 t}=(2 \pi)^{n / 2} k_{t}(x, y)
$$

which is clearly a smooth function of $x \in \mathbb{R}^{n}$ for $t>0$. That it satisfies the heat equation is an immediate consequence of (i) that $\left(\partial_{t}+\Delta_{x}\right) k_{t}(x, y)=0$ for $t>0$. This shows (a) of (iii).

To see (b), note that for $f \in \mathcal{S}\left(\mathbb{R}^{n}\right)$, we have

$$
(f, w(-, t))=\int_{\mathbb{R}^{n}} f(x) w(x, t) d V(x)=\int_{\mathbb{R}^{n}} \rho_{t}(x-y) f(x) d V(x)=\int_{\mathbb{R}^{n}} \rho_{t}(y-x) f(x) d V(x)=\left(\rho_{t} * f\right)(y)
$$

and the proposition follows by noting that for $\phi(x)=\rho_{\frac{1}{2}}(x)$ :

$$
\int_{\mathbb{R}^{n}} \phi(x) d V(x)=\int_{\mathbb{R}^{n}} \rho_{\frac{1}{2}}(x) d V(x)=\int_{\mathbb{R}^{n}} e^{-|x|^{2} / 2} d V(x)=1
$$

and that by setting $\epsilon^{2}=2 t$, we have

$$
\left.\phi_{\epsilon}(x)=\epsilon^{-n} \phi(x / \epsilon)=(2 t)^{-n / 2} \rho_{\frac{1}{2}}(x / \epsilon)=(2 t)^{-n / 2} e^{-|x|^{2} / 2 \epsilon^{2}}=(2 t)^{-n / 2}\right) e^{-|x|^{2} / 4 t}=\rho_{t}(x)
$$

But by the Lemma 1.2 .3 , we have $\phi_{\epsilon}$ are approximate identities, and $\phi_{\epsilon} * f=\rho_{t} * f \rightarrow f$ uniformly on $\mathbb{R}^{n}$, for all $f \in \mathcal{S}\left(\mathbb{R}^{n}\right)$. Thus $(f, w(-, t)) \rightarrow\left(\rho_{t} * f\right)(y)$ has limit $f(y)$ as $t \rightarrow 0$. This proves (b). To see that (a) and (b) uniquely determine $w(-, t)$, let us assume $u(-, t) \in \mathcal{S}\left(\mathbb{R}^{n}\right)$ also satisfies (a) and (b). Denoting $w_{t}:=w(-, t), u_{t}:=u(-, t)$ for notational convenience, note that the time derivative of the $L_{2}$-norm of $w_{t}-u_{t}$ is given by (because both $w_{t}$ and $u_{t}$ are rapidly decreasing, Stokes Formula is applicable):

$$
\partial_{t}\left(w_{t}-u_{t}, w_{t}-u_{t}\right)=-2\left(\Delta\left(w_{t}-u_{t}\right), w_{t}-u_{t}\right)=-2\left(d\left(w_{t}-u_{t}\right), d\left(w_{t}-u_{t}\right)\right) \leq 0
$$

and so $\left\|w_{t}-u_{t}\right\|^{2}$ is a non-increasing function of $t$. Also, by (b), for every $f$ we have:

$$
\lim _{t \rightarrow 0}\left(f, w_{t}-u_{t}\right)=f(y)-f(y)=0
$$

which means $\lim _{t \rightarrow 0}\left\|w_{t}-u_{t}\right\| \rightarrow 0$, and by the fact that $\left\|w_{t}-u_{t}\right\|$ is non-increasing in $t$, it follows that $w_{t} \equiv u_{t}$ for all $t>0$. This proves the proposition.
11.2. Fundamental solutions of the Heat equation for the Dirac Laplacians. Now let $M$ be a compact Riemannian manifold, and let $D^{ \pm}$be the Dirac operators introduced in Definition 9.6.1.

Definition 11.2.1 (Fundamental solutions). Let $x \in M$ and let $v \in E_{x}^{+}$. We say that a smooth section $w(-, t) \in C^{\infty}\left(M, E^{+}\right)$is a fundamental solution with pole $(x, v)$ if:
(i): The section $w(-, t)$ satisfies the heat equation for $\Delta^{+}$, viz.

$$
\left(\partial_{t}+\Delta^{+}\right) w(x, t)=0 \text { for all } x \in M, \quad t>0
$$

(ii): $\lim _{t \rightarrow 0}(s, w(-, t))=\langle s(x), v\rangle_{x}$ for all $s \in C^{\infty}\left(M, E^{+}\right)$

The second condition means that $w_{t}$ approaches the "Dirac distributional-section" (at the point $x$ ) which is given by $\delta_{x} v$ as $t \rightarrow 0$.

One can obviously make a similar definition for $E^{-}$and $\Delta^{-}$.

Proposition 11.2.2 (Existence and uniqueness of fundamental solutions). Let $M, E^{ \pm}$be as above. Then given $v \in E_{x}^{+}$, there exists a fundamental solution with pole $(x, v)$ to the heat equation for $\Delta^{+}$, and this solution is unique. Likewise for $E^{-}$and $\Delta^{-}$.

Proof: We merely apply (iii) of the Proposition 10.1.3 to $f=\delta_{x} v$, this "Dirac distributional section" $\delta_{x} v$. We also note that this $f \in H_{-k}\left(M, E^{+}\right)$for all $k>n / 2$, and by the Remark 10.1.5, we will have that

$$
w(z, t)=\int_{M} k_{t}^{+}(z, y) \delta_{x}(y) v d V(y)=k_{t}^{+}(z, x) v
$$

is a smooth section of $E^{+}$for $t>0$. Indeed, the right hand side is clearly smooth in $z$ since $k_{t}^{+}$is smooth in $z$ ( $x$ is held fixed here) for $t>0$. Furthermore, for $t>0$, we have

$$
\left(\partial_{t}+\Delta^{+}\right)(w(z, t))=\left(\partial_{t}+\Delta_{z}^{+}\right)\left(k_{t}^{+}(z, x) v\right)=0
$$

by using (ii) of Proposition 10.1.6.
For the convergence as $t \rightarrow 0$, we have:

$$
\begin{aligned}
(s, w(-, t)) & =\int_{M}\langle s(z), w(z, t)\rangle_{z} d V(z)=\int_{M}\left\langle s(z), k_{t}^{+}(z, x) v\right\rangle_{z} d V(z) \\
& =\int_{M}\left\langle k_{t}^{+}(x, z) s(z), v\right\rangle_{x} d V(z)=\left\langle\int_{M} k_{t}^{+}(x, z) s(z) d V(z), v\right\rangle_{x} \\
& =\left\langle\left(e^{-t \Delta^{+}} s\right)(x), v\right\rangle_{x}
\end{aligned}
$$

where we have used the adjointness-symmetry property (i) of Proposition 10.1.6 to arrive at the second line. Now, by (ii) of Proposition 10.1.3, we have by the smoothness of $s$ that $\lim _{t \rightarrow \infty} e^{-t \Delta^{+}} s \rightarrow s$ in the $\|-\|_{\infty, 0}$ (i.e the convergence is uniform over $M$ ), which means that the limit at $x$ satisfies:

$$
\lim _{t \rightarrow 0}\left(e^{-t \Delta^{+}} s\right)(x)=s(x)
$$

and hence $\lim _{t \rightarrow 0}(s, w(-, t))=\langle s(x), v\rangle_{x}$, and our assertion follows. Likewise for $E^{-}$and $\Delta^{-}$.
To see uniqueness, just verbatim repeat the argument for uniqueness given in (iii) of the Proposition 11.1.1, only noting that for $w_{t}-u_{t}$, we have:

$$
-\left(\Delta^{+}\left(w_{t}-u_{t}\right), w_{t}-u_{t}\right)=-\left(D^{+}\left(w_{t}-u_{t}\right), D^{+}\left(w_{t}-u_{t}\right)\right) \leq 0
$$

by the formal-adjointness of $D^{+}$and $D^{-}$proved in (i) of Proposition 9.6.2.

Exercise 11.2.3. For $M=S^{1}$, and $E^{+}=\Lambda^{0} T^{*}(M), E^{-}=\Lambda^{1} T^{*}(M), D^{+}=d, D^{-}=\delta$ (i.e. the two term deRham elliptic complex for $S^{1}$, whose associated Dirac complex is itself), explicitly write down the heat kernels $k_{t}^{+}$and $k_{t}^{-}$and carry out the verifications of all the preceding propositions in this subsection and the previous one by hand.

## 12. Asymptotic expansions of the heat kernel

This approach is due to Minakshisundaram and Pleijel. First, assuming that one has an asymptotic expansion, one computes the coefficients in this expansion by substituting in the heat equation and equating coefficients term-by-term. Then one appeals to elliptic estimates to prove that the formal procedure above makes sense.

### 12.1. Asymptotic expansions.

Definition 12.1.1. Let $f$ be any function on $(0, \infty)$. A formal series $\sum_{k=0}^{\infty} a_{k} t^{n_{k}}$ (where $n_{k} \in \mathbb{Z}$ ) is said to be an asymptotic expansion for $f$ near 0 if:
(i): $n_{k}<n_{k+1}$ for all $k$ (so that $n_{k} \rightarrow \infty$ as $k \rightarrow \infty$ ), and,
(ii): For each $l \geq 0$, there exists a $C_{l} \geq 0$ such that

$$
\left|f(t)-\sum_{k=0}^{l} a_{k} t^{n_{k}}\right| \leq C_{l} t^{n_{l+1}}
$$

This will be denoted by $f(t) \sim \sum_{k=0}^{\infty} a_{k} t^{n_{k}}$ (Compare with asymptotic expansions of symbols introduced in Definition 5.3.2).

For example, for the function $k_{t}(x, y)=(4 \pi t)^{-n / 2} e^{-|x-y|^{2} / 4 t}$ introduced above, regarded as a function of $t$, we would have that:

$$
(4 \pi)^{-n / 2} \sum_{k=0}^{\infty}\left(\frac{|x-y|^{2 k}}{4^{k} k!}\right) t^{-n / 2-k}
$$

is an asymptotic expansion near 0 . Note that the expansion starts with $t^{-n / 2}$.
For a motivation, knowing the heat kernel on $\mathbb{R}$ for the Laplacian $\Delta=-\partial_{x}^{2}$, let us try to find an asymptotic expansion for the heat kernel of the operator $L=\Delta+b(x) \partial_{x}+c(x)$ where $b$ and $c$ are smooth functions. It is in fact enough to find the fundamental solution: $u(x, t)$ satisfying $\left(\partial_{t}+L\right) u(x, t)=0$, and $\lim _{t \rightarrow 0} u(x, t)=\delta_{x}$. Then one gets the heat kernel by $k_{t}(x, y)=u(x-y, t)$ (verify!). To this end we have:

Proposition 12.1.2. Let $L=\Delta+b(x) \partial_{x}+c(x)$ as above, where $b$ and $c$ are smooth functions on $\mathbb{R}$. Then there is an asymptotic fundamental solution to the corresponding heat equation $\left(\partial_{t}+L\right) u(x, t)=0$. That is, there is a formal series:

$$
(4 \pi t)^{-1 / 2} e^{-x^{2} / 4 t}\left(u_{0}(x)+t u_{1}(x)+\ldots+t^{k} u_{k}(x)+\ldots\right)
$$

where $u_{j}(x)$ are smooth functions of $x$, with $u_{0}(0)=1$, such that for the partial sum

$$
S_{k}(x, t):=(4 \pi t)^{-1 / 2} e^{-x^{2} / 4 t}\left(\sum_{j=0}^{k} t^{j} u_{j}(x)\right)
$$

we have:

$$
\left(\partial_{t}+L\right) S_{k}(x, t)=(4 \pi t)^{-1 / 2} e^{-x^{2} / 4 t} t^{k} r_{k}(x)
$$

where $r_{k}(x)$ is a smooth function of $x$. Furthermore, $u_{j}(0)$ are algebraic expressions (i.e. polynomials) in the jets (derivatives of all orders) of $b$ and $c$ at 0 .

Proof: The idea is to determine the $u_{j}(x)$ by a recursive formula. So in the PDE $\left(\partial_{t}+L\right) u(x, t)=0$, let us substitute the series

$$
(4 \pi t)^{-1 / 2} e^{-x^{2} / 4 t}\left[u_{0}(x)+t u_{1}(x)+t^{2} u_{2}(x)+\ldots\right]
$$

for $u(x, t)$. The coefficient of $t^{k}$ in the expression within square-brackets is $u_{k}$.
Note that the formal series on differentiating with respect to $x$ is

$$
\partial_{x} u(x, t)=(4 \pi t)^{-1 / 2} e^{-x^{2} / 4 t}\left[-\frac{x}{2 t}\left(u_{0}+t u_{1}+\ldots\right)+\left(u_{0}^{\prime}+t u_{1}^{\prime}+\ldots\right)\right]
$$

where $u_{i}^{\prime}$ denotes $\partial_{x} u_{i}$. Note that the coefficient of $t^{k}$ in the expression within square-brackets is:

$$
u_{k}^{\prime}-\frac{x}{2} u_{k+1}
$$

Differentiating again with respect to $x$, we have:

$$
\partial_{x}^{2} u(x, t)=(4 \pi t)^{-1 / 2} e^{-x^{2} / 4 t}\left[\left(\frac{x^{2}}{4 t^{2}}-\frac{1}{2 t}\right)\left(u_{0}+t u_{1}+\ldots\right)-\frac{x}{t}\left(u_{0}^{\prime}+t u_{1}^{\prime}+\ldots\right)+\left(u_{0}^{\prime \prime}+t u_{1}^{\prime \prime}+\ldots\right)\right]
$$

The coefficient of $t^{k}$ in the expression within square-brackets is:

$$
\frac{x^{2}}{4} u_{k+2}-\frac{1}{2} u_{k+1}-x u_{k+1}^{\prime}+u_{k}^{\prime \prime}
$$

Taking the $t$-derivative, we have:

$$
\partial_{t} u(x, t)=(4 \pi t)^{-1 / 2} e^{-x^{2} / 4 t}\left[\left(u_{1}+2 t u_{2}+3 t^{2} u_{3}+\ldots\right)+\left(\frac{x^{2}}{4 t^{2}}-\frac{1}{2 t}\right)\left(u_{0}+t u_{1}+\ldots\right)\right]
$$

The coefficient of $t^{k}$ in the expression within square brackets is:

$$
(k+1) u_{k+1}+\frac{x^{2}}{4} u_{k+2}-\frac{1}{2} u_{k+1}=\left(k+\frac{1}{2}\right) u_{k+1}+\frac{x^{2}}{4} u_{k+2}
$$

Now substitute this into the heat equation for $L$ to get:

$$
\left(\partial_{t}+L\right) u(x, t)=\left(\partial_{t}-\partial_{x}^{2}+b(x) \partial_{x}+c(x)\right) u(x, t)=(4 \pi t)^{-1 / 2} e^{-x^{2} / 4 t}\left[\sum_{k=-2}^{\infty} \alpha_{k} t^{k}\right]
$$

where:

$$
\begin{aligned}
\alpha_{k} & =\left(k+\frac{1}{2}\right) u_{k+1}+\frac{x^{2}}{4} u_{k+2}-\frac{x^{2}}{4} u_{k+2}+\frac{1}{2} u_{k+1}+x u_{k+1}^{\prime}-u_{k}^{\prime \prime}+b(x) u_{k}^{\prime}-b(x) \frac{x}{2} u_{k+1}+c(x) u_{k} \\
& =x u_{k+1}^{\prime}+\left(k+1-\frac{x b(x)}{2}\right) u_{k+1}+L u_{k}
\end{aligned}
$$

Setting $\alpha_{k}=0$ gives a recursive differential equation for $u_{k+1}$ in terms of $u_{k}$. That is, the equation:

$$
\begin{equation*}
x u_{k+1}^{\prime}+\left(k+1-\frac{x b(x)}{2}\right) u_{k+1}+L u_{k}=0 \tag{25}
\end{equation*}
$$

Since $u_{-1}=0$ by definition, we have on substituting $k=-1$ in the equation (25) above the following differential equation for $u_{0}$ :

$$
u_{0}^{\prime}-\frac{b(x)}{2} u_{0}=0
$$

which implies that $u_{0}=A e^{-\frac{1}{2} \int_{0}^{x} b(y) d y}$ for some constant $A$, and setting the requirement that $u_{0}(0)=1$ implies that

$$
u_{0}=e^{-\frac{1}{2} \int_{0}^{x} b(y) d y}
$$

More generally, consider the integrating factor:

$$
R_{k}(x)=x^{k+1} e^{-\frac{1}{2} \int_{0}^{x} b(y) d y}
$$

we get $\log R_{k}(x)=(k+1) \log x-\frac{1}{2} \int_{0}^{x} b(y) d y$ so that:

$$
\frac{1}{R_{k}(x)} \frac{d}{d x}\left(R_{k}(x) u_{k+1}\right)=\left(\frac{k+1}{x}-\frac{b(x)}{2}\right) u_{k+1}+u_{k+1}^{\prime}=\frac{1}{x}\left[\left((k+1)-\frac{x b(x)}{2}\right) u_{k+1}+x u_{k+1}^{\prime}\right]=-\frac{1}{x} L u_{k}
$$

by (25), so that

$$
u_{k+1}(x)=\frac{-1}{R_{k}(x)}\left(\int_{0}^{x} \frac{R_{k}(y)}{y} L u_{k}(y) d y\right)
$$

gives the explicit recursive formula for $u_{k+1}$ in terms of $u_{k}$.
Now if we take the partial sum:

$$
S_{k}(x)=(4 \pi t)^{-\frac{1}{2}} e^{-x^{2} / 4 t}\left(u_{0}+t u_{1}+t^{2} u_{2}+\ldots+t^{k} u_{k}\right)
$$

with $u_{k}$ defined as above, then write:

$$
\left(\partial_{t}+L\right) S_{k}=(4 \pi t)^{-\frac{1}{2}} e^{-x^{2} / 4 t}\left[\beta_{-2} t^{-2}+\ldots . \beta_{k} t^{k}\right]
$$

since $\alpha_{j}$ contains no contribution from the term $t^{j+2} u_{j+2}$, we have that the coefficient $\beta_{j}=\alpha_{j}$ for all $j \leq k-1$. Also $\beta_{k}=L u_{k}$, since the rest of the expression for $\alpha_{k}$ involves $u_{k+1}$.

So we finally have:

$$
\left(\partial_{t}+L\right) S_{k}=(4 \pi t)^{-1 / 2} t^{k} e^{-x^{2} / 4 t}\left(L u_{k}\right)
$$

which implies the differential equation asserted for $S_{k}$.
We need to show that the $u_{k}$ 's defined above are smooth. We do this by induction. The function $u_{0}=$ $\exp \left(-\frac{1}{2} \int b(y) d y\right)$ is clearly smooth by definition. Also the integrating factor $R_{k}$ is given by:

$$
R_{k}(x)=x^{k+1} u_{0}(x)
$$

from the above proof. Hence if we inductively assume that $u_{k}$ is smooth, we will have:

$$
-\frac{R_{k}(y)}{y} L u_{k}(y)=y^{k} \gamma_{k}(y)
$$

where $\gamma_{k}(y)=-u_{0}(y) L u_{k}(y)$ is smooth in $y$. Hence the integral:

$$
-\int_{0}^{x} \frac{R_{k}(y)}{y} L u_{k}(y) d y=\int_{0}^{x} y^{k} \gamma_{k}(y) d y=x^{k+1} \rho_{k}(x)
$$

where $\rho_{k}(x)$ is a smooth function in $x$ (using integration by parts, for example). Thus the formula for $u_{k+1}(x)$ in the proof above reads

$$
u_{k+1}(x)=\frac{1}{R_{k}(x)}\left(-\int_{0}^{x} \frac{R_{k}(y)}{y} L u_{k}(y) d y\right)=\frac{1}{R_{k}(x)}\left(x^{k+1} \rho_{k}(x)\right)=\frac{\rho_{k}(x)}{u_{0}(x)}
$$

which is clearly smooth in $x$ since $u_{0}$ is a nowhere vanishing smooth function. Note that adding a constant of integration to the indefinite integral $\int_{0}^{x} \frac{R_{k}(y)}{y} L u_{k}(y) d y$ will destroy this property, because we need this integral to yield the factor $x^{k+1}$. Hence, by induction, all the $u_{k}$ are smooth.

The final assertion is that $u_{k}(0)$ are polynomial expressions in the various jets (higher derivatives) of $b$ and $c$ at zero. Indeed, we claim that $u_{k}^{(r)}(0)$ are all polynomials in the various jets of $b$ and $c$ at 0 . We do this by double induction on $k$ and $r$. For $k=0$, by definition $u_{0}(0)=1$, and $u_{0}^{\prime}(x)=\frac{b(x)}{2} u_{0}(x)$ implies by Leibnitz rule that:

$$
u_{0}^{(r+1)}(0)=\frac{1}{2} \sum_{0 \leq j \leq r} \frac{r!}{(r-j)!j!} b^{(r-j)}(0) u_{0}^{(j)}(0)
$$

so that induction on $r$ shows that our claim is true for $k=0$. Assume inductively that it is true for $u_{k}$, i.e. $u_{k}^{(r)}(0)$ is a polynomial in the various jets of $b$ and $c$ at 0 for all $r$. Since $L=-\partial_{x}^{2}+b(x) \partial_{x}+c(x)$, it follows by the induction hypothesis that $\left(L u_{k}\right)^{(r)}(0)$ is also a polynomial in the various jets of $b$ and $c$ at 0 for all $r$. From the equation (25) it follows that:

$$
u_{k+1}(0)=\frac{1}{k+1}\left(L u_{k}\right)(0)
$$

so that the claim is true for $u_{k+1}(0)$. Differentiating the equation $(25)(r+1)$ times with respect to $x$ yields:

$$
x u_{k+1}^{(r+2)}+(r+1) u_{k+1}^{(r+1)}+\left(k+1-\frac{x b(x)}{2}\right) u_{k+1}^{(r+1)}+\sum_{0 \leq j \leq r} A_{j}(x) u_{k+1}^{(j)}+\left(L u_{k}\right)^{(r+1)}=0
$$

where $A_{j}(x)$ is a polynomial in $x, b(x), \ldots, b^{(j)}(x)$. Setting $x=0$ in this equation shows that:

$$
(r+k+2) u_{k+1}^{(r+1)}(0)=-\sum_{0 \leq j \leq r} A_{j}(0) u_{k+1}^{(j)}(0)-\left(L u_{k}\right)^{(r+1)}(0)
$$

Since $A_{j}(0)$ is a polynomial in the various jets of $b$ at 0 , and by the last para $\left(L u_{k}\right)^{(r+1)}(0)$ is a polynomial in the jets of $b$ and $c$ at 0 , the last equation above implies by induction on $r$ that $u_{k+1}^{(r+1)}$ is a polynomial in the jets of $b$ and $c$ at 0 . The proposition follows.

Let us prove another technical lemma which will be used later on.
Lemma 12.1.3. Let $g=\sum_{i, j} g_{i j} d x_{i} \otimes d x_{j}$ be a Riemannian metric on a suitably small ball $U$ around the origin in $\mathbb{R}^{n}$, and let $g^{i j}$ be the corresponding metric on 1- forms, i.e. $g^{i j}=g_{i j}^{-1}$ is the matrix inverse of $g$. This metric defines a Riemannian distance on this ball, which we denote by $\delta$. Define a smooth function on $U$ :

$$
f(x, t)=(4 \pi t)^{-n / 2} \exp \left(\frac{-\delta(0, x)^{2}}{4 t}\right)
$$

Then

$$
\partial_{t} f-\sum_{i, j} g^{i j} \partial_{i} \partial_{j} f=\left(\frac{1}{t} a_{1}+a_{2}\right) f
$$

where $\partial_{i}=\frac{\partial}{\partial x_{i}}$, and $a_{1}, a_{2}$ are smooth functions of $x$, with $a_{1}(0)=0$.

Proof: Let us denote the scalar Laplacian (on functions) on $U$ by $\Delta$. We claim that

$$
\Delta=-\sum_{i, j} g^{i j} \partial_{i} \partial_{j}+L
$$

where $L$ is a 1st-order operator. This is because we saw in the Example 9.4.4 that $\sigma_{L}(d)=i \xi \wedge(-)$, and also by (ii) of the Corollary 9.3.4 that $\sigma_{L}\left(d^{*}\right)=(-i \xi) \angle(-)$, the adjoint of $\sigma_{L}(d)$. Thus $\sigma_{L}\left(d^{*} d\right)=|\xi|^{2}=\sum_{i, j} g^{i j} \xi^{i} \xi^{j}$. Thus

$$
\Delta=\sum_{i, j} g^{i j} D_{x, i} D_{x, j}+L=-\sum_{i j} g^{i j} \partial_{i} \partial_{j}+L
$$

where $L=\sum_{i} \alpha_{i}(x) \partial_{i}+\beta(x)$ is a first-order operator.
Now note that for the first-order operator $L=\sum_{i} \alpha_{i}(x) \partial_{i}+\beta(x)$ as above, we have

$$
L f=(4 \pi t)^{-n / 2} \sum_{i} \alpha_{i}(x)\left[-\frac{1}{4 t} \partial_{i}\left(\delta(0, x)^{2}\right) \exp \left(\frac{-\delta(0, x)^{2}}{4 t}\right)\right]+\beta(x) f=\left(\frac{1}{t} c_{1}(x)+c_{2}(x)\right) f
$$

where $c_{i}$ are smooth, and also

$$
c_{1}(0)=\frac{-1}{4} \sum_{i} \alpha_{i}(0) \partial_{i}\left(\delta(0, x)^{2}\right)(0)=\frac{-1}{4} \sum_{i} \alpha_{i}(0)\left(\sum_{j} g_{i j} x_{j}\right)(0)=0
$$

since

$$
\delta(0, x)=\|x\|+o\left(\|x\|^{2}\right)=\left(\sum_{i, j} g_{i j}(0) x_{i} x_{j}\right)^{1 / 2}+o\left(\|x\|^{2}\right)
$$

where $\|x\|$ denotes the norm of $x$ in the tangent space $T_{0}\left(\mathbb{R}^{n}\right)$ with respect to $g_{i j}(0)$.
Thus it is enough to prove that:

$$
\partial_{t} f+\Delta f=\left(\frac{1}{t} a_{1}+a_{2}\right) f
$$

where $a_{i}$ are smooth, and $a_{1}(0)=0$. Now it is convenient to use geodesic polar coordinates on $U$, i.e. polar coordinates on $T_{0}\left(\mathbb{R}^{n}\right)=\mathbb{R}^{n}$ transferred to $U$ by the exponential map. We may shrink $U$ to guarantee that the exponential map is a diffeomorphism of a neighbourhood of 0 in $T_{0}\left(\mathbb{R}^{n}\right)$ onto $U$. In these polar coordinates $\delta(0, x)=r$, and

$$
f(x, t)=(4 \pi t)^{-n / 2} \exp \left(\frac{-r^{2}}{4 t}\right)
$$

which is a radial function. It is also known that if $x=\exp _{0}\left(x_{1}, . ., x_{n}\right)$ is a vector in $T_{0}\left(\mathbb{R}^{n}\right)$, then $r^{2}=$ $\delta(0, x)^{2}=\|x\|^{2}=\sum_{i, j} g_{i j}(0) x_{i} x_{j}$. Thus for the function $f(r, t)$, which does not depend on any of the other polar coordinates $v_{2}, . ., v_{n}$ on the unit sphere, we have:

$$
\partial_{i} f=\partial_{r} f \frac{\partial r}{\partial x_{i}}=\frac{1}{r} \sum_{k} g_{i k}(0) x_{k} \partial_{r} f
$$

Differentiating again, multiplying with $-g^{i j}$ and summing over $i, j$ yields:

$$
\Delta f=-\partial_{r}^{2} f-\frac{(n-1)}{r} \partial_{r} f
$$

Now:

$$
\partial_{r} f=\frac{-r}{2 t} f
$$

and so

$$
\partial_{r}^{2} f=\frac{-1}{2 t} f+\frac{r^{2}}{4 t^{2}} f
$$

Thus

$$
\Delta f=\frac{n}{2 t} f-\frac{r^{2}}{4 t^{2}} f
$$

Finally,

$$
\partial_{t} f=-\frac{n}{2 t} f+\frac{r^{2}}{4 t^{2}} f
$$

Hence $\partial_{t} f+\Delta f=0$, which is of the required form.
12.2. Generalised Laplacians. We will now look at elliptic operators of a special kind, because these will be of primary interest in whatever follows.

In this section, $E$ is to be thought of as either $E^{+}$or $E^{-}$, the complex Hermitian vector bundles arising in the Dirac complex. Also, the operator $\Delta^{E}$ which will be cropping up in this section will be the operators $\Delta^{+}$ or $\Delta^{-}$in our future considerations.

Definition 12.2.1. Let $E$ be a complex vector bundle on a compact Riemannian manifold $M$, with Hermitian metric $\langle-,-\rangle$. Let $P: C^{\infty}(M, E) \rightarrow C^{\infty}(M, E)$ be a differential operator of order 2 . We say that $P$ is a generalised Laplacian if:
(i): $P$ is a formally self-adjoint, viz., $(P f, g)=(f, P g)$ for all $f, g \in C^{\infty}(M, E)$, where $\left(f_{1}, f_{2}\right)$ is the usual global Hermitian inner product on $C^{\infty}(M, E)$ defined by:

$$
\left(f_{1}, f_{2}\right)=\int_{M}\left\langle f_{1}(x), f_{2}(x)\right\rangle_{x} d V(x) \quad f_{i} \in C^{\infty}(M, E)
$$

(ii): The leading symbol of $P$ satisfies:

$$
\sigma_{L}(P)(\xi)=|\xi|^{2} I_{E_{x}}, \quad \xi \in T^{*} M_{x}
$$

In future we will suppress $I_{E_{x}}$ from the notation, with the understanding that the scalar $|\xi|^{2}$ means that scalar times the identity endomorphism of $\left(\pi^{*} E\right)_{\xi}=E_{x}$.

Remark 12.2.2. Using (i) of the Corollary 9.3.4, we have for a second operator that:

$$
\sigma_{L}(P)(\xi)=\frac{-1}{2}(a d f)^{2} P=\frac{-1}{2}[f,[f, P]]
$$

for $f$ such that $d f(x)=\xi$. Thus $P$ is a generalised Laplacian iff $P$ is formally self-adjoint of order 2 and:

$$
[f,[f, P]]=-2|\xi|^{2}=-2|d f|^{2}
$$

for each $f \in C^{\infty}(M)$.

One can easily construct a generalised Laplacian $P$ as above by using a connection $\nabla^{E}$ on the bundle $E$ and the Levi-Civita connection $\nabla$ on the Riemannian manifold $M$, as we see below.

First we recall that there is a trace map defined by:

$$
\begin{aligned}
\operatorname{tr}: C^{\infty}\left(M, T^{*} M \otimes T^{*} M\right) & \rightarrow C^{\infty}(M) \\
f & \mapsto f(g)
\end{aligned}
$$

where $g \in C^{\infty}(M, T M \otimes T M)$ is the Riemannian metric (on the cotangent bundle) given by $g=\sum_{i, j} g^{i j} \partial_{i} \otimes \partial_{j}$ in local coordinates. By tensoring with $I_{E}$, we get a map as below, which we also denote by tr,

$$
\operatorname{tr}: C^{\infty}\left(M, T^{*} M \otimes T^{*} M \otimes E\right) \rightarrow C^{\infty}(M, E)
$$

If $s \in C^{\infty}\left(M, T^{*} M \otimes T^{*} M \otimes E\right)$, then in a local coordinate system $x_{i}$ at a point, we may write

$$
s=\sum_{i, j} s\left(\partial_{i}, \partial_{j}\right) d x_{i} \otimes d x_{j}
$$

where $s\left(\partial_{i}, \partial_{j}\right)$ is a local smooth section of $E$. Then

$$
\begin{equation*}
\operatorname{tr} s=\sum_{i, j} s\left(\partial_{i}, \partial_{j}\right) d x_{i} \otimes d x_{j}\left(\sum_{k, l} g^{k, l} \partial_{k} \otimes \partial_{l}\right)=\sum_{i, j} s\left(\partial_{i}, \partial_{j}\right) g^{i j} \tag{26}
\end{equation*}
$$

Now let

$$
\nabla^{E}: C^{\infty}(M, E) \rightarrow C^{\infty}\left(M, T^{*} M \otimes E\right)
$$

be a connection on $E$. By taking the natural "tensor product connection" $I_{T^{*} M} \otimes \nabla^{E}+\nabla \otimes I_{E}$ (where $\nabla$ is the Levi-Civita connection on $T^{*} M$, we also get a connection:

$$
\nabla^{T^{*} M \otimes E}: C^{\infty}\left(M, T^{*} M \otimes E\right) \rightarrow C^{\infty}\left(M, T^{*} M \otimes T^{*} M \otimes E\right)
$$

Definition 12.2.3 (The operator $\Delta^{E}$ ). Define the second order differential operator:

$$
\Delta^{E}=-\operatorname{tr}\left(\nabla^{T^{*} M \otimes E} \circ \nabla^{E}\right): C^{\infty}(M, E) \rightarrow C^{\infty}(M, E)
$$

Lemma 12.2.4. The operator $\Delta^{E}$ defined above is a generalised Laplacian.

Proof: First, if $s \in C^{\infty}(M, E)$, we have, in local coordinates $x_{i}$ at a point:

$$
\nabla^{E} s=\sum_{i} d x_{i} \otimes \nabla_{i}^{E} s
$$

where we set $\nabla_{i}^{E} s:=\nabla_{\partial_{i}}^{E} s$ to simplify notation.
Then we compute $\nabla^{T^{*} M \otimes E}$ of both sides:

$$
\begin{aligned}
\nabla^{T^{*} M \otimes E} \nabla^{E} s & =\sum_{i} \nabla^{T^{*} M \otimes E}\left(d x_{i} \otimes \nabla_{i}^{E} s\right)=\sum_{i} \nabla\left(d x_{i}\right) \otimes \nabla_{i}^{E} s+\sum_{i} d x_{i} \otimes \nabla^{E} \nabla_{i}^{E} s \\
& =\sum_{i, j}\left(d x_{j} \otimes \nabla_{j}\left(d x_{i}\right) \otimes \nabla_{i}^{E} s+d x_{i} \otimes d x_{j} \otimes \nabla_{j}^{E} \nabla_{i}^{E} s\right)
\end{aligned}
$$

where $\nabla$ denotes the Levi-Civita connection. Now note that for a tangent vector $Y$ :

$$
\nabla_{j}\left(d x_{i}\right)(Y)=\partial_{j}\left(d x_{i}(Y)\right)-d x_{i}\left(\nabla_{j} Y\right)=\partial_{j} Y_{i}-d x_{i}\left(\nabla_{j} Y\right)
$$

Thus, for tangent vectors $X=\sum_{i} X_{i} \partial_{i}$ and $Y=\sum_{j} Y_{j} \partial_{j}$, we have:

$$
\begin{equation*}
\left(\nabla^{T^{*} M \otimes E} \nabla^{E} s\right)(X, Y)=\sum_{i, j} X_{j}\left(\partial_{j} Y_{i}-d x_{i}\left(\nabla_{j} Y\right)\right) \nabla_{i}^{E} s+\sum_{i, j} X_{i} Y_{j} \nabla_{j}^{E} \nabla_{i}^{E} s \tag{27}
\end{equation*}
$$

Now again note that the second term of the equation (27) above is:

$$
-\sum_{i, j} X_{j} d x_{i}\left(\nabla_{j} Y\right) \nabla_{i}^{E} s=-\sum_{j} X_{j} \sum_{i}\left(\nabla_{j} Y\right)_{i} \nabla_{i}^{E} s=-\sum_{j} X_{j} \nabla_{\nabla_{j} Y}^{E} s=-\nabla_{\nabla_{X} Y}^{E} s
$$

Also:

$$
\nabla_{X}^{E} \nabla_{Y}^{E} s=\sum_{j} X_{j} \nabla_{j}^{E}\left(\sum_{i} Y_{i} \nabla_{i}^{E} s\right)=\sum_{j} X_{j} Y_{i} \nabla_{j}^{E} \nabla_{i}^{E} s+X_{j}\left(\partial_{j} Y_{i}\right) \nabla_{i}^{E} s
$$

which are precisely the first and third terms of (27). In conclusion:

$$
\left(\nabla^{T^{*} M \otimes E} \nabla^{E} s\right)(X, Y)=\nabla_{X}^{E} \nabla_{Y}^{E} s-\nabla_{\nabla_{X} Y}^{E} s
$$

Thus, by applying the equation (26) above, we find that:

$$
\Delta^{E} s=-\sum_{i, j} g^{i j}\left(\nabla_{i}^{E} \nabla_{j}^{E}-\sum_{k} \Gamma_{i j}^{k} \nabla_{k}^{E}\right) s
$$

where the Christoffel symbols $\Gamma_{i j}^{k}$ are defined by

$$
\nabla_{i} \partial_{j}=\sum_{k} \Gamma_{i j}^{k} \partial_{k}
$$

Since the leading symbols of $\nabla_{i}$ is just $\partial_{i}$, it follows that $\Delta^{E}$ has the same leading symbol as the operator given in local coordinates by:

$$
-\sum_{i j} g^{i j} \partial_{i} \partial_{j}=\sum_{i, j} g^{i j} D_{x, i} D_{x, j}
$$

But his leading symbol precisely $\sum_{i, j} g^{i j} \xi^{i} \xi^{j}=|\xi|^{2}$, the symbol of the Laplacian. So $\Delta^{E}$ is a generalised Laplacian. We remark here that the leading symbol depends only on the Riemannian metric $g$ on $M$, and does not depend on the connection $\nabla^{E}$ on $E$.

Remark 12.2.5. We have already remarked in the proof of the Lemma 12.1.3 that for the usual LaplaceBeltrami operator on functions, (i.e. the Laplacian on $C^{\infty}(M)$ of the deRham complex) that $\sigma_{L}(\Delta)=$ $-\sum_{i, j} g^{i j}(x) \partial_{i} \partial_{j}$. Thus, from the above proposition it follows that no matter what connection one puts on $E$, we have:

$$
\Delta^{E}=\Delta+L
$$

where $L$ is a first order differential operator. $L$, of course, will depend on $E$. We will study it in greater detail later, and see the connection with the Bochner and Weitzenbock formulas.
12.3. Fundamental solutions of the Heat Equation for generalised Laplacians. By the Proposition 11.2.2, we have the existence and uniqueness of a fundamental solution $w(x, t)$ to the Dirac Laplacians $\Delta^{ \pm}$. For the elliptic complexes we consider in the sequel, all of these Dirac Laplacians $\Delta^{ \pm}$will be generalised Laplacians. (Indeed, they will all arise as $\Delta^{E}$ as in Definition 12.2.3, and Lemma 12.2 .4 will imply that they are generalised Laplacians). The proof of the existence and uniqueness of the fundamental solution $w(x, t)$ used the eigensections and eigenvalues of $\Delta^{ \pm}$, which gives little information about the behaviour of the fundamental solution, because one cannot explicitly compute these eigenvalues and eigensections.

The objective of this section is to gain more information by actual construction of the fundamental solution of Proposition 11.2.2, by starting out with a Gaussian type fundamental solution as in $\mathbb{R}^{n}$, and applying an iterated approximation process using asymptotic solutions for generalised Laplacians. Because this iterative procedure is explicit, it will in theory "solve" the problem of computing the fundamental solution.

Since by definition a generalised Laplacian $\Delta^{E}$ has the same leading symbol as the Laplace-Beltrami operator $\Delta$, it follows that

$$
P=\Delta+L
$$

where

$$
L=\sum_{i=1}^{n} b_{i}(x) \partial_{i}+c(x)
$$

is a first order operator. We have a handle on the fundamental solution for $\Delta$ by Lemma 12.1.3, we can try to mimic the argument of Proposition 12.1.2 to obtain a fundamental solution for the heat equation:

$$
\left(\partial_{t}+P\right) u=0
$$

by asymptotic methods.

Proposition 12.3.1. Let $P: C^{\infty}(M, E) \rightarrow C^{\infty}(M, E)$ be a generalised Laplacian on the compact Riemannian manifold $M$. Let 0 be some preassigned point on $M$. Then, on a suitably small neighbourhood $U$ of 0 , there is an asymptotic solution:

$$
u(x, t) \sim(4 \pi t)^{-n / 2} \exp \left(-\frac{\delta(0, x)^{2}}{4 t}\right)\left(u_{0}(x)+t u_{1}(x)+\ldots t^{k} u_{k}(x)+. .\right)
$$

where $\delta(0, x)$ is the Riemannian distance between 0 and $x$ in the metric $g$, and $u_{k}(x)$ are smooth sections of $E$ on $U$. That is to say,

$$
\left(\partial_{t}+P\right) S_{k}(x, t)=(4 \pi t)^{-n / 2} \exp \left(-\frac{\delta(0, x)^{2}}{4 t}\right) t^{k} r_{k}(x) \quad \text { for } x \in U, \quad t \in(0, \infty)
$$

where $S_{k}(x, t)$ is the partial sum $\sum_{j=0}^{k} t^{j} u_{j}(x)$, and $r_{k}(x)$ is a smooth function on $U . u_{0}(0)$ can be given any preassigned non-zero (vector) value $v \in E_{0}$, and for all $k$, each component of $u_{k}(0)$ is a polynomial in the $v_{i}(0)$ and the various jets of $g^{i j}, b_{i}$ and $c$ at $0\left(b_{i}(x)\right.$ and $c(x)$ being the coefficients occurring in the first order operator $L:=P-\Delta$ as in the last paragraph).

Proof: First of all note that we may use a coordinate chart $U$ around 0 on which $E$ is trivial, and which is diffeomorphic to $\mathbb{R}^{n}$. So, we take $M=\mathbb{R}^{n}$, and $E$ the trivial bundle. By coordinatewise application, we can also assume that $E$ is the trivial line bundle. Since $P$ is a generalised Laplacian, we can take:

$$
P=\Delta+L=-\sum_{i, j} g^{i j}(x) \partial_{i} \partial_{j}+L, \text { where } L=\sum_{i} b_{i}(x) \partial_{i}+c(x)
$$

and $b_{i}(x)$ and $c(x)$ are smooth. 0 is the origin in $\mathbb{R}^{n}$. We will use the geodesic normal coordinates $\left(x_{1}, . ., x_{n}\right)$ introduced in the proof of Lemma 12.1.3.

Now if $f, v \in C^{\infty}(U)$ are two smooth functions, we have by Leibnitz's formula:

$$
\begin{align*}
P(f v) & =-\sum_{i, j} g^{i j}(x) \partial_{i} \partial_{j}(f v)+\sum_{i} b_{i}(x) \partial_{i}(f v)+c(x)(f v) \\
& =-\sum_{i, j} g^{i j}(x)\left(v \partial_{i} \partial_{j} f+2\left(\partial_{i} v\right)\left(\partial_{j} f\right)+f \partial_{i} \partial_{j} v\right)+\sum_{i} b_{i}(x)\left(f \partial_{i} v+v \partial_{i} f\right)+c(x) f v \\
& =f P v-v \sum_{i, j} g^{i j}(x) \partial_{i} \partial_{j} f-2 \sum_{i, j} g^{i j}(x)\left(\partial_{i} v\right)\left(\partial_{j} f\right)+v \sum_{i} b_{i}(x) \partial_{i} f \tag{28}
\end{align*}
$$

Thus we have,

$$
\begin{equation*}
\frac{1}{f}\left(\partial_{t}(f v)+P(f v)\right)=\left(\partial_{t} v+P v\right)+\frac{v}{f}\left(\partial_{t} f-\sum_{i, j} g^{i j}(x) \partial_{i} \partial_{j} f\right)-\frac{2}{f} \sum_{i, j} g^{i j}(x)\left(\partial_{i} v\right)\left(\partial_{j} f\right)+\frac{v}{f} \sum_{i} b_{i}(x) \partial_{i} f \tag{29}
\end{equation*}
$$

Now set $f=f(x, t)=(4 \pi t)^{-n / 2} \exp \left(\frac{-\delta(0, x)^{2}}{4 t}\right)$ in the above formula. By the Lemma 12.1.3, we have (upon shrinking $U$ if necessary) that:

$$
\partial_{t} f-\sum_{i, j} g^{i j} \partial_{i} \partial_{j} f=\left(\frac{1}{t} a_{1}+a_{2}\right) f
$$

where $a_{i}$ are smooth, and $a_{1}(0)=0$. Also, since we are using geodesic normal coordinates, we have $r^{2}=$ $\|x\|^{2}=\sum_{i, j} g_{i, j}(0) x_{i} x_{j}=\sum_{i} x_{i}^{2}$. Now $f$ being a radial function (i.e. only a function of $r$ ), we have:

$$
\frac{1}{f} \partial_{i} f=\frac{1}{f} \partial_{r} f \partial_{i} r=-\frac{2 r}{4 t} \frac{x_{i}}{r}=-\frac{x_{i}}{2 t}
$$

Substituting these two facts into the equation (29), we have:

$$
\frac{1}{f}\left(\partial_{t}(f v)+P(f v)\right)=\left(\partial_{t} v+P v\right)+v\left(\frac{1}{t} a_{1}+a_{2}\right)-2 g^{i j} \partial_{i} v\left(\frac{-x_{j}}{2 t}\right)+v \sum_{j} b_{j}(x)\left(-\frac{x_{j}}{2 t}\right)
$$

which implies that:

$$
\begin{equation*}
\left.\left(\partial_{t}(f v)+P(f v)\right)=f\left[\partial_{t} v+P v\right)+u\left(\frac{1}{t} a_{1}+a_{2}\right)+\sum_{j} \frac{x_{j}}{t}\left(\sum_{i} g^{i j} \partial_{i} v-\frac{1}{2} v b_{j}\right)\right] \tag{30}
\end{equation*}
$$

Since we are using geodesic normal coordinates, the radial vector field $\partial_{r}$ has length one, and is in the same direction as $x=\sum_{j} x_{j} e_{j}$, where $e_{j}(x)=\sum g^{i j}(x) \partial_{i}$ is an orthonormal frame at $x \in U$. Thus

$$
\partial_{r}=\frac{1}{\|x\|} \sum_{j} x_{j} e_{j}=\frac{1}{r} \sum_{i, j} g^{i j}(x) x_{j} \partial_{i}
$$

Substituting this into the equation (30) above, we get:

$$
\begin{equation*}
\left(\partial_{t}(f v)+P(f v)\right)=f\left[\left(\partial_{t} v+P v\right)+u\left(\frac{1}{t} a_{1}+a_{2}\right)-\left(\sum_{j} \frac{x_{j} b_{j}}{2 t} v+\frac{r}{t} \partial_{r} v\right)\right] \tag{31}
\end{equation*}
$$

Now let $v=\sum_{k=0}^{\infty} t^{k} u_{k}(x)$, so that

$$
u(x, t)=f v=(4 \pi t)^{-n / 2} \exp \left(\frac{-\delta(0, x)^{2}}{4 t}\right)\left(\sum_{k} t^{k} u_{k}(x)\right)
$$

Since we want to satisfy $\left(\partial_{t}+P\right) u=\left(\partial_{t}(f v)+P(f v)\right)=0$, we compute the coefficient of $t^{k}$ in the box-brackets on the right hand side of equation (31) and set it equal to zero. The coefficient of $t^{k}$ on the right hand side of equation (31) is

$$
\begin{aligned}
(k+1) u_{k+1} & +P u_{k}+a_{1} u_{k+1}+a_{2} u_{k}-\frac{1}{2} \sum_{j} b_{j} x_{j} u_{k+1}+r \partial_{r} u_{k+1} \\
& =r \partial_{r} u_{k+1}+\left(k+1+a_{1}-\frac{1}{2} \sum_{j} x_{j} b_{j}\right) u_{k+1}+a_{2} u_{k}+P u_{k}
\end{aligned}
$$

which leads to the recursive differential equation:

$$
\begin{equation*}
r \partial_{r} u_{k+1}+\left(k+1+a_{1}-\frac{1}{2} \sum_{j} x_{j} b_{j}\right) u_{k+1}+\left(P+a_{2}\right) u_{k}=0 \tag{32}
\end{equation*}
$$

From the Lemma 12.1.3, we have $a_{1}(0)=0$, so we may write $a_{1}(x)=\sum_{j} x_{j} A_{j}(x)$ for some smooth functions $A_{j}$ by the first order Taylor expansion. We write $x=r y$ where $\|y\|=1$, i.e. $y$ is on the unit sphere $S^{n-1}$. Then let $B(r, y):=-2 \sum_{j}\left(y_{j} A_{j}(r, y)+y_{j} b_{j}(r, y)\right)$. If we also write $\Lambda:=P+a_{2}$, our equation above becomes:

$$
r \partial_{r} u_{k+1}+\left(k+1-\frac{1}{2} r B(r, y)\right) u_{k+1}+\Lambda u_{k}=0
$$

This equation is an ODE in the variable $r$, with $y \in S^{n-1}$ being treated as a smooth parameter, and identical to the earlier single-variable equation (25), with $r$ playing the role of $x, B(r, y)$ playing the role of the earlier $b(x)$, and $\Lambda$ playing the role of the earlier $L$. Thus it is solved along any ray $y=y_{0} \in S^{n-1}$ by exactly the same procedure as in the Proposition 12.1.2. The resulting $u_{j}$ are smooth in $x$, because of the inductive formula

$$
u_{k+1}(r, y)=-\frac{1}{R_{k}(x)} \int_{0}^{r} \frac{R_{k}(x) \Lambda u_{k}(x)}{\|x\|} d(\|x\|) ; \quad R_{k}(x)=\|x\|^{k+1} \exp \left(-\int_{0}^{r} B(x) d(\|x\|)\right)
$$

using the same argument as in Proposition 12.1.2, after noting the fact that $B$ and the coefficients of the differential operator $\Lambda$ are smooth in $x\left(\right.$ since $b_{j}(x), c(x), a_{1}(x), a_{2}(x)$ are all smooth in $\left.x\right)$.

To see the last assertion about $u_{k}(0)$, we have as before that $u_{k}(0)$ will be algebraic in the various jets of $B$ and the coefficients of $\Lambda$ at 0 . That is, they will be algebraic in the jets of $g^{i j}, b_{j}, c, a_{1}$ and $a_{2}$ at 0 . We just need to show that the jets of $a_{1}$ and $a_{2}$ at 0 are algebraic in the jets of $g^{i j}$ at 0 . If we go back to the proof of Lemma 12.1.3, we find that $a_{1}$ and $a_{2}$ defined there are precisely $c_{1}$ and $c_{2}$, where $c_{1}$ and $c_{2}$ are defined by:

$$
c_{1}(x)=-\frac{1}{4} \sum_{i} \alpha_{i}(x) \partial_{i} r^{2}=-\frac{1}{2} \sum_{i} x_{i} \alpha_{i}(x), \quad c_{2}(x)=\beta(x)
$$

where

$$
\Delta=-\sum_{i, j} g^{i j} \partial_{i} \partial_{j}+\sum_{i} \alpha_{i}(x) \partial_{i}+\beta(x)
$$

Now from the calculation of the Laplacian for the metric $g=g_{i j}$, one knows that $\alpha_{i}(x)$ is an algebraic expression in the first derivatives of $g$, and $\beta(x)$ is an algebraic expression in the second derivatives of $g$. Hence the jets of $c_{1}$ and $c_{2}$ at 0 are algebraic expressions in the jets of $g$ at 0 , from the equations for $c_{1}$ and $c_{2}$ above. The proposition follows.

Proposition 12.3.2 (Duhamel's Principle). Let $M$ be a compact Riemannian manifold, and let

$$
\Delta^{+}: C^{\infty}\left(M, E^{+}\right) \rightarrow C^{\infty}\left(M, E^{+}\right)
$$

be the Dirac Laplacian corresponding to an elliptic complex on $M$. Let us assume that $\Delta^{+}$is of order 2 . Let $\sigma_{t}$ be a smoothly varying section in $C^{\infty}\left(M, E^{+}\right)$, (i.e. $\sigma_{(-)} \in C^{\infty}\left((0, \infty) \times M, p^{*} E^{+}\right)$where $p:(0, \infty) \times M \rightarrow M$ is the second projection). Then there exists a unique smooth solution $\rho_{t}$ which is also smooth in $t$, satisfying:
(i): $\rho_{0}=0$, and
(ii): $\rho_{t}$ satisfies the inhomogenous time-dependent heat equation:

$$
\left(\partial_{t}+\Delta^{+}\right) \rho_{t}=\sigma_{t}
$$

for all $t \in(0, \infty)$.
Likewise for $E^{-}$and $\Delta^{-}$.

Proof: If $\sigma_{t}=\sigma$ were independent of $t$, our $\rho_{t}$ would be $e^{-t \Delta^{+}} \sigma$. In general, we add up the contributions $e^{-(t-s) \Delta^{+}} \sigma_{s}$. That is, define:

$$
\rho_{t}=\int_{0}^{t} e^{-(t-s) \Delta^{+}} \sigma_{s} d s
$$

Note that the integral makes sense, since the integrand is smooth in $s$, and on differentiating both sides with respect to $t$ (and using the dominated convergence theorem), we have:

$$
\begin{aligned}
\partial_{t} \rho_{t} & =e^{-(t-t) \Delta^{+}} \sigma_{t}+\int_{0}^{t} \partial_{t}\left(e^{-(t-s) \Delta^{+}} \sigma_{s}\right) d s \\
& =\sigma_{t}-\int_{0}^{t} \Delta^{+}\left(e^{-(t-s) \Delta^{+}} \sigma_{s}\right) d s \\
& =\sigma_{t}-\Delta^{+} \rho_{t}
\end{aligned}
$$

The uniqueness follows from the fact that for another solution $u_{t}$ satisfying both (i) and (ii), we have:

$$
\partial_{t}\left(\rho_{t}-u_{t}\right)=-\Delta^{+}\left(\rho_{t}-u_{t}\right)
$$

so that

$$
\partial_{t}\left(\rho_{t}-u_{t}, \rho_{t}-u_{t}\right)=-2\left(\Delta^{+}\left(\rho_{t}-u_{t}\right), \rho_{t}-u_{t}\right)=-2\left(D^{+}\left(\rho_{t}-u_{t}\right), D^{+}\left(\rho_{t}-u_{t}\right)\right) \leq 0
$$

which shows that the $L_{2}$-norm $\left\|\rho_{t}-u_{t}\right\|^{2}$ is a non-increasing function of $t$. But since it is zero at $t=0$ by (ii), it follows that it is identically zero.

Corollary 12.3.3. For the $\rho_{t}$ found above, we have the Sobolev norm estimates:

$$
\left\|\rho_{t}\right\|_{2 k} \leq t \sup _{0 \leq s \leq t}\left\|\sigma_{s}\right\|_{2 k}
$$

for all $k=0,1,2 \ldots$, .
Proof: We first note that for any $f \in C^{\infty}\left(M, E^{+}\right)$and for all $\mu \geq 0$ we have $e^{-\mu \lambda_{i}} \leq 1$ for all the eigenvalues $\lambda_{i} \geq 0$ of $\Delta^{+}$, and consequently the inequality of $L_{2}$-norms:

$$
\left\|e^{-\mu \Delta^{+}} f\right\| \leq\|f\| \quad \text { for all } \mu \geq 0
$$

Now, by the Corollary 6.2.3 (Garding's Inequality) it follows that:

$$
\begin{aligned}
\left\|e^{-\mu \Delta^{+}} f\right\|_{2 k}^{2} & =\left\|\Delta^{+k} e^{-\mu \Delta^{+}} f\right\|^{2}+\left\|e^{-\mu \Delta^{+}} f\right\|^{2}=\left\|e^{-\mu \Delta^{+}}\left(\Delta^{+k} f\right)\right\|^{2}+\left\|e^{-\mu \Delta^{+}} f\right\|^{2} \\
& \leq\left\|\Delta^{+, k} f\right\|^{2}+\|f\|^{2}=\|f\|_{2 k}^{2} \quad \text { for all } \mu \geq 0
\end{aligned}
$$

Hence

$$
\left\|\rho_{t}\right\|_{2 k} \leq \int_{0}^{t}\left\|e^{-(t-s) \Delta^{+}} \sigma_{s}\right\|_{2 k} d s \leq \int_{0}^{t}\left\|\sigma_{s}\right\|_{2 k} d s \leq t \sup _{0 \leq s \leq t}\left\|\sigma_{s}\right\|_{2 k}
$$

and the corollary follows.
Now we can prove that "asymptotic fundamental solutions" converge to real fundamental solutions. More precisely, we have:

Proposition 12.3.4. In the setting of the previous proposition, let $w_{t}$ be the unique fundamental solution to the heat equation for the Dirac operator $\Delta^{+}$, with pole at $(x, v)$ (whose existence was proved in the Proposition 11.2.2). Let $u_{t}$ be a smooth section, varying smoothly in $t$ (see last proposition for definition) which satisfies:
(i): For all $s \in C^{\infty}\left(M, E^{+}\right)$, we have

$$
\lim _{t \rightarrow 0}\left(s, u_{t}\right)=\langle s(x), v\rangle_{x}
$$

(That is $u_{t}$ converges to the Dirac distributional section $\delta_{x} v$ as $t \rightarrow 0$ ), and
(ii):

$$
\left(\partial_{t}+\Delta^{+}\right) u_{t}=t^{N} r_{t}(x)
$$

where $r_{t}$ is a smooth section of $E^{+}$, smoothly varying for $t \in(0, \infty)$ and continuous in $t \in[0, \infty)$ and uniformly bounded in the Sobolev $2 k$-norm $\|-\|_{2 k}$ for $t \in[0, T]$ and some $T>0$ and some $k \geq 0$. (This means $\left\|r_{t}\right\|_{2 k} \leq C$ for all $t \in[0, T]$, where $C$ is a positive constant.)

Then we have:

$$
\left\|w_{t}-u_{t}\right\|_{\infty, l} \leq C_{l} t^{N+1} \quad \text { for all } \quad l<2 k-n / 2 \text { and all } t \in(0, T]
$$

where $C_{l}>0$ is some constant.

Proof: By the Duhamel Principle Proposition 12.3.2, there exists a smoothly varying smooth section $\rho_{t}$ of $E^{+}$ satisfying:

$$
\left(\partial_{t}+\Delta^{+}\right) \rho_{t}=t^{N} r_{t}
$$

and also satisfying $\rho_{0}=0$. Then the smoothly varying section $w_{t}:=u_{t}-\rho_{t}$ satisfies:

$$
\left(\partial_{t}+\Delta^{+}\right) w_{t}=\left(\partial_{t}+\Delta^{+}\right) u_{t}-\left(\partial_{t}+\Delta^{+}\right) \rho_{t}=t^{N} r_{t}-t^{N} r_{t}=0
$$

Also, for any smooth section $s \in C^{\infty}\left(M, E^{+}\right)$, we have:

$$
\lim _{t \rightarrow 0}\left(s, w_{t}\right)=\lim _{t \rightarrow 0}\left(s, u_{t}\right)=\langle s(x), v\rangle_{x}
$$

since $\rho_{0}=0$.

Thus $w_{t}$ is the unique fundamental solution of the heat equation with pole at $(x, v)$. It follows that $u_{t}=\rho_{t}+w_{t}$ and by the Corollary 12.3.3

$$
\left\|w_{t}-u_{t}\right\|_{2 k}=\left\|\rho_{t}\right\|_{2 k} \leq t \sup _{0 \leq s \leq t}\left\|s^{N} r_{s}\right\|_{2 k}=t^{N+1} \sup _{0 \leq s \leq T}\left\|r_{s}\right\|_{2 k} \quad \text { for } t \in(0, T]
$$

By the hypothesis on the Sobolev $2 k$-norm $\left\|r_{s}\right\|_{2 k}$ for $s \in[0, T]$ it follows that for $t \in[0, T]$ we have:

$$
\sup _{0 \leq s \leq T}\left\|r_{s}\right\|_{2 k} \leq C \text { for all } s \in[0, T]
$$

Thus it follows that:

$$
\left\|w_{t}-u_{t}\right\|_{2 k} \leq C t^{N+1} \quad \text { for all } t \in(0, T]
$$

Now one uses Sobolev's Embedding Theorem (iv) of Proposition 4.2 .2 which asserts that

$$
\|-\|_{\infty, l} \leq C\|-\|_{2 k} \text { for all } l<2 k-n / 2
$$

to get the assertion for $\|-\|_{\infty, l}$ with $l<2 k-n / 2$.

Theorem 12.3.5 (Asymptotic fundamental solution for the heat equation of a generalised Dirac Laplacian). Let

$$
\Delta^{+}=D^{-} D^{+}: C^{\infty}\left(M, E^{+}\right) \rightarrow C^{\infty}\left(M, E^{+}\right)
$$

be the Dirac Laplacian of the Dirac complex defined by an elliptic complex $\mathcal{P}$ on the compact Riemannian manifold $M$ of dimension $n$. Assume that $\Delta^{+}$is a generalised Laplacian in the sense of Definition 12.2.1. Let $v \in E_{a}^{+}$be some vector, and let $w_{t}$ be the fundamental solution to the heat equation for $\Delta^{+}$with pole at $(a, v)$, which exists and is unique by the Proposition 11.2.2. Then there exists an asymptotic fundamental solution $u(x, t)=u_{t}(x)$ with pole at $(a, v)$ which is given by a formal series:

$$
u(x, t)=(4 \pi t)^{-n / 2} \exp \left(-\frac{\delta(x, a)^{2}}{4 t}\right)\left(u_{0}(x)+t u_{1}(x)+t^{2} u_{2}(x)+\ldots+t^{k} u_{k}(x)+\ldots\right) \quad x \in M, \quad t \in(0, \infty)
$$

where $\delta(x, a)$ denotes the Riemannian distance between $x$ and $a$, and $u_{i}(x)$ are smooth functions of $x$. The value $u_{0}(a)=v$, and in a suitable local coordinate neighbourhood of $a$ the and local framing of $E^{+}$, for every $k$, each component of the vector $u_{k}(a)$ is a polynomial in the $p$-jets at the point $a$ of $g^{i j}$ and the coefficients $b_{i}, c$ occurring in the first-order operator:

$$
\Delta^{+}+\sum_{i j} g^{i j} \partial_{i} \partial_{j}=\sum_{l} b_{l}(y) \partial_{l}+c(y)
$$

This asymptotic solution satisfies:
(i): For each smooth section $s \in C^{\infty}\left(M, E^{+}\right)$,

$$
\lim _{t \rightarrow 0}\left(s, u_{t}\right)=\langle s(a), v\rangle_{a}
$$

(ii): Given any positive integer $N>0$, for the partial sum

$$
S_{m}(x, t):=(4 \pi t)^{-n / 2} \exp \left(-\frac{\delta(x, a)^{2}}{4 t}\right)\left(\sum_{k=0}^{m} t^{k} u_{k}(x)\right)
$$

we have:

$$
\left(\partial_{t}+\Delta^{+}\right) S_{m}(x, t)=t^{N} r_{m, t}(x) \quad \text { for all } m \geq N+n / 2
$$

where $r_{m, t}(x)=r_{m}(x, t)$ is a smoothly varying section in $C^{\infty}\left(M, E^{+}\right)$and continuous for $t \in[0, \infty)$. Indeed, $r_{m}(x, 0) \equiv 0$. If we fix some $T>0$, then its Sobolev $2 k$-norm on $M$ satisfies:

$$
\left\|r_{m, t}\right\|_{2 k} \leq C_{k, m} \quad \text { for all } 2 k \leq m-N-n / 2 \quad \text { and all } t \in[0, T]
$$

Finally,
(iii): For the $T>0$ in (ii) above, we have the norm estimate:

$$
\left\|w_{t}-S_{m}(-, t)\right\|_{l, \infty} \leq C_{l} t^{N+1}
$$

for each $0 \leq l \leq m-N-n$ and all $t \in(0, T]$.
Likewise for $\Delta^{-}$and $E^{-}$.

Proof: Let $U$ be a neighbourhood of $a \in M$ such that $U$ is diffeomorphic to a neighbourhood of 0 in $\mathbb{R}^{n}=T_{a}(M)$ via the exponential map $\exp _{a}: T_{a}(M) \rightarrow M$ of the Riemannian manifold $M$. Since $\Delta^{+}$is a generalised Laplacian by hypothesis, we may further guarantee that $U$ is small enough for the Proposition 12.3.1 to apply to $P=\Delta^{+}$.

By restricting to a ball around $a$ contained in $U$, we may assume without loss of generality that $U=B(a, 3 \epsilon)$. Then, by that Proposition we have a formal series:

$$
\widetilde{u}(x, t) \sim(4 \pi t)^{-n / 2} \exp \left(-\delta(a, x)^{2} / 4 t\right)\left(\widetilde{u}_{0}+t \widetilde{u}_{1}+\ldots+t^{k} \widetilde{u}_{k}+\ldots\right)
$$

where $\widetilde{u}_{i}(x)$ are smooth functions defined on $U$. Furthermore, $\widetilde{u}_{0}(a)=v$, and in a suitable framing of $E^{+}$on $U$, each component of each $\widetilde{u}_{k}(a)$ is a polynomial in the $p$-jets of $g^{i j}, b_{i}, c$ at $a$. Also on $U$ we have, by the proof of Propositions 12.1.2 and 12.3.1 that:

$$
\left(\partial_{t}+\Delta^{+}\right) \widetilde{S}_{m}(x, t)=(4 \pi t)^{-n / 2} t^{m} \exp \left(-\delta(x, a)^{2} / 4 t\right) \Lambda \widetilde{u}_{m}(x)
$$

where $\Lambda=\Delta^{+}+a_{2}$ is also a generalised Laplacian defined on $U$. If $m \geq N+n / 2$, the function

$$
\begin{equation*}
\widetilde{r}_{m}(x, t):=t^{m-N-n / 2} \exp \left(-\delta(x, a)^{2} / 4 t\right) \Lambda \widetilde{u}_{m}(x) \tag{33}
\end{equation*}
$$

is a smoothly varying section of $E_{\mid U}^{+}$, continuous and uniformly bounded in the norm $\|-\|_{x}$ of $E_{x}^{+}$, for all $x \in U$ and all $t \in[0, T]$. That is,

$$
\sup _{x \in U, t \in[0, T]}\left\|\widetilde{r}_{m}(x, t)\right\|_{x}<\infty
$$

Note that the equation (33) above implies that $\widetilde{r}_{m}(x, 0) \equiv 0$ for $m>N+n / 2$. Since

$$
\partial_{i}\left(t^{p} \exp \left(-\delta(x, a)^{2} / 4 t\right)\right)=\left(p-\frac{x_{i}}{2}\right) t^{p-1} \exp \left(-\delta(x, a)^{2} / 4 t\right)
$$

on $U$, we see that for $m \geq N+n / 2+2 k$, the $L_{2}$-norm:

$$
\left\|\partial^{\alpha} \widetilde{r}_{m}\right\|_{0, U}^{2}:=\int_{x \in U}\left\|\partial^{\alpha} \widetilde{r}_{m}(x, t)\right\|_{x}^{2} d V(x)
$$

will be finite and uniformly bounded for all $|\alpha| \leq 2 k \leq m-N-n / 2$. Thus the Sobolev $2 k$-norm of $\widetilde{r}_{m}(-, t)$ on $U$ satisfies:

$$
\left\|\widetilde{r}_{m}(-, t)\right\|_{2 k, U} \leq C_{k, m} \text { for all } 2 k \leq m-N-n / 2, \text { and all } t \in[0, T]
$$

Thus we have:

$$
\begin{equation*}
\left(\partial_{t}+\Delta^{+}\right) \widetilde{S}_{m}(x, t)=t^{N} \widetilde{r}_{m}(x, t) \quad \text { for all } \quad x \in U, \quad \text { and } m>N+n / 2 \tag{34}
\end{equation*}
$$

with $\widetilde{r}_{m}(x, t)$ a smoothly varying section of $E_{\mid U}^{+}$for $t \in(0, \infty)$, continuous in $t \in[0, \infty)$, and uniformly bounded in Sobolev norm $\|-\|_{2 k, U}$ by a positive constant $C_{k, m}$ for all $t \in[0, \infty)$ and all $2 k \leq m-N-n / 2$.

The first step is to globalise $\widetilde{u}(x, t)$ for all $x \in M$. We do this via a cut-off function. Let

$$
\psi: \mathbb{R} \rightarrow[0, \infty)
$$

be a smooth function such that $\psi(s) \equiv 1$ for $|s| \leq \epsilon$ and $\psi(s) \equiv 0$ for $|s| \geq 2 \epsilon$.

To simplify notation, denote $r:=r(x):=\delta(x, a)$. Then define:

$$
u(x, t)=\psi(r(x)) \widetilde{u}(x, t)
$$

so that

$$
u(x, t) \sim(4 \pi t)^{-n / 2} \exp \left(-\frac{\delta(x, a)^{2}}{4 t}\right)\left(u_{0}(x)+t u_{1}(x)+\ldots .+t^{k} u_{k}(x)+\ldots .\right)
$$

where $u_{k}(x):=\psi(r) \widetilde{u}_{k}(x)$. Since $\psi(r(x))$ is identically 1 on $B(0, \epsilon)$, the function $\psi(r(x))$ is a smooth function of $x$, and hence the $u_{k}$ 's defined above are smooth functions of $x$ on all of $M$. Furthermore:

$$
\begin{aligned}
u_{k}(x) & =\widetilde{u}_{k}(x) \quad \text { for } \delta(x, a) \leq \epsilon \\
& =0 \quad \text { for } \quad \delta(x, a) \geq 2 \epsilon
\end{aligned}
$$

Hence the statement about $u_{k}(a)$ follows from the corresponding statements about $\widetilde{u}_{k}(a)$.

For notational convenience, denote:

$$
f(x, t):=(4 \pi t)^{-n / 2} \exp \left(-\frac{\delta(x, a)^{2}}{4 t}\right)
$$

Since $\psi$ is supported in $U$, we have for a smooth section $s \in C^{\infty}\left(M, E^{+}\right)$:

$$
\begin{aligned}
\lim _{t \rightarrow 0}\left(s, u_{t}\right) & =\lim _{t \rightarrow 0} \int_{M}\langle s(x), u(x, t)\rangle_{x} d V(x)=\lim _{t \rightarrow 0} \int_{U}\langle s(x), \psi(r(x)) \widetilde{u}(x, t)\rangle_{x} d V(x) \\
& =\lim _{t \rightarrow 0} \int_{U}\langle\psi(r(x)) s(x), \widetilde{u}(x, t)\rangle_{x} d V(x) \\
& =\lim _{t \rightarrow 0} \int_{U}\left\langle\psi(r(x)) s(x), f(x, t) \widetilde{u}_{0}(x)\right\rangle_{x} d V(x)=\langle\psi(a) s(a), v\rangle_{a} \\
& =\langle s(a), v\rangle_{a}
\end{aligned}
$$

because $\psi(a)=1$ and $\widetilde{u}_{0}(a)=v$ by the Proposition 12.3.1, and $f(x, t)$ is an approximate identity at $x=a$ for compactly supported smooth sections in $U$, and $\psi(r(x)) s(x)$ is a smooth section compactly supported in $U$. This proves (i) of the Theorem.

Now we prove (ii). We have by definition that $S_{m}(x, t)=\psi(r(x)) \widetilde{S}_{m}(x, t)$. Hence

$$
\begin{equation*}
\partial_{t} S_{m}(x, t)=\psi(r) \partial_{t} \widetilde{S}_{m}(x, t) \text { for all } x \in M \tag{35}
\end{equation*}
$$

where the right hand side is interpreted to be identically zero for $x \notin U$ (i.e. $\delta(x, a) \geq 3 \epsilon)$.
On the other hand,

$$
\begin{equation*}
\Delta^{+}\left(\psi(r) \widetilde{S}_{m}(x, t)\right)=\psi(r) \Delta^{+} \widetilde{S}_{m}(x, t)+\mu(r) L \widetilde{S}_{m}(x, t) \tag{36}
\end{equation*}
$$

where

$$
\mu(r):=a(r) \psi^{\prime}(r)+b(r) \psi^{\prime \prime}(r)
$$

and $L\left(=\sum_{i} \alpha_{i}(x) \partial_{i}+\beta(x)\right.$ in $U$, and $\equiv 0$ oustside some $\left.V \supset B(0,2 \epsilon)\right)$ is some first order linear differential operator in the space variables on $M$. We already understand the first term, from the foregoing discussion, and we need to estimate the second term. Since

$$
\widetilde{S}_{m}(x, t)=f(x, t) \sum_{k=0}^{m} t^{k} \widetilde{u}_{k}(x)
$$

we compute for $x \in U$ :

$$
\begin{aligned}
L \widetilde{S}_{m}(x, t) & =\left[\sum_{i=1}^{n} \alpha_{i}(x) \partial_{i}(x)+\beta(x)\right] f \sum_{k=0}^{m} t^{k} \widetilde{u}_{k}(x) \\
& =(L f) \sum_{k=0}^{m} t^{k} \widetilde{u}_{k}(x)+f \sum_{k=0}^{m} t^{k}(L-\beta(x)) \widetilde{u}_{k}(x) \\
& =f\left(\frac{1}{t} c_{1}(x)+c_{2}(x)\right) \sum_{k=0}^{m} t^{k} \widetilde{u}_{k}(x)+f \sum_{k=0}^{m} t^{k} w_{k}(x) \\
& =t^{-1} f(x, t) P_{m}(x, t) \quad \text { for } x \in U
\end{aligned}
$$

where $P_{m}(x, t)$ is a polynomial of degree $m$ in $t$ whose coefficients are smooth sections of $E_{\mid U}^{+}$. Note that we have used the first paragraph of the Lemma 12.1.3 to substitute $L f=\left(\frac{1}{t} c_{1}+c_{2}\right) f$.

Since $\psi^{\prime}(r)$ and $\psi^{\prime \prime}(r)$ identically vanish for $0 \leq r \leq \epsilon$ and $r \geq 2 \epsilon$, it follows that $\mu(r) \equiv 0$ for $0 \leq r \leq \epsilon$ and $r \geq 2 \epsilon$.

Consider the section:

$$
h_{m}(x, t):=t^{-N-1} \mu(r) f(x, t) P_{m}(x, t)
$$

Since $f(x, t) \leq(4 \pi t)^{-n / 2} e^{-\epsilon^{2} / 4 t}$ for $r \geq \epsilon \quad$ and $t \in[0, \infty)$, and $\mu(r)$ vanishes identically for $r \leq \epsilon$ and $r \geq 2 \epsilon$, it follows that the section above is a smooth section of $E^{+}$with compact support in the annulus $\epsilon \leq r \leq 2 \epsilon$, for every $N \geq 0$ and all $t \in[0, \infty)$. At $t=0$, it is the identically zero function. Hence we may write:

$$
\begin{equation*}
\mu(r) L \widetilde{S}_{m}(x, t)=t^{N}\left(t^{-N-1} \mu(r) f(x, t) P_{m}(x, t)\right)=t^{N} h_{m}(x, t) \quad \text { for all } \quad x \in M, \text { and all } m \geq 0 \tag{37}
\end{equation*}
$$

where $h_{m}(x, t)=t^{-N-1} \mu(r) f(x, t) P_{m}(x, t), P_{m}$ being a polynomial of degree $m$ in $t$ whose coefficients are smooth sections in the variable $x \in M, h_{m}(x, t) \equiv 0$ for $r(x)=\delta(x, a) \leq \epsilon$ and $r(x) \geq 2 \epsilon$ and all $t \in[0, \infty)$, with $h_{m}(x, 0) \equiv 0$ on $M$.

To get a hold on the Sobolev $2 k$-norm of $h_{m}(x, t)$, note that $P_{m}$ is a polynomial of degree $m$ in $t$, whose coefficients are smooth sections. Also each spatial derivative of $f(x, t)$ will yield $\left(t^{-1} a_{1}+a_{2}\right) f$, and any spatial derivative of $\mu(r)$ will again yield a smooth function compactly supported in the annulus $\epsilon \leq r \leq 2 \epsilon$. Hence, for the $L_{2}$-norm:

$$
\int_{M}\left\|\partial_{x}^{\alpha} h_{m}(x, t)\right\|_{x}^{2} d V(x) \leq C t^{-N-1-|\alpha|} \int_{\epsilon \leq r \leq 2 \epsilon} f(x, t)^{2} d V(x) \leq C t^{-N-1-|\alpha|} e^{-\epsilon^{2} / 2 t} \quad \text { for } t \in[0, T]
$$

Thus we have:

$$
\begin{equation*}
\sup _{t \in[0, T]}\left\|h_{m}(-, t)\right\|_{2 k}<\infty \quad \text { for all } k \tag{38}
\end{equation*}
$$

Now we can combine all the equations (34), (35), (36) and (37) to compute:

$$
\begin{align*}
\left(\partial_{t}+\Delta^{+}\right) S_{m}(x, t) & =\left(\partial_{t}+\Delta^{+}\right) \psi(r) \widetilde{S}_{m}(x, t)=\psi(r)\left(\partial_{t}+\Delta^{+}\right) \widetilde{S}_{m}(x, t)+\mu(r) L \widetilde{S}_{m}(x, t) \\
& =\psi(r)\left(t^{N} \widetilde{r}_{m}(x, t)\right)+t^{N} h_{m}(x, t) \\
& =t^{N} r_{m}(x, t) \tag{39}
\end{align*}
$$

for all $x \in M$, all $m \geq N+n / 2$ and all $t \in(0, \infty)$, and $r_{m}(x, t):=\psi(r) \widetilde{r}_{m}(x, t)+h_{m}(x, t)$. Also,
(a): From the equations (34) and (37) it follows that $r_{m}(x, 0)=\psi(r(x)) \widetilde{r}_{m}(x, 0)+h_{m}(x, 0) \equiv 0$.
(b): From the statement following equation (34), the fact that

$$
\left\|\psi(r) \widetilde{r}_{m}(-, t)\right\|_{2 k} \leq C\left\|\widetilde{r}_{m}(-, t)\right\|_{2 k, U}
$$

and from the inequality (38), it follows that:

$$
\sup _{t \in[0, T]}\left\|r_{m}(-, t)\right\|_{2 k} \leq C_{k} \quad \text { for all } \quad 2 k \leq m-N-n / 2
$$

This establishes (ii) of the Theorem. The final assertion (iii) now follows from the Corollary 12.3.3.

Example 12.3.6 (The Circle). For the circle, one can explicitly write down the heat kernel, and the fundamental solution by tinkering with the fundamental solution for $\mathbb{R}$.

Let $S^{1}(R)$ denote the circle of radius $R$ around the origin in $\mathbb{R}^{2}$, and let $\theta \in(-\pi, \pi)$ denote the usual angle coordinate in the open set $S^{1}(R) \backslash\{-R\}$. The Riemannian metric is $R^{2} d \theta^{2}$, and the corresponding Riemannian volume of $S^{1}(R)$ is $2 \pi R$. We consider the Dirac complex of the de-Rham complex of the circle, viz. with $E^{+}=\Lambda^{0}\left(T_{\mathbb{C}}^{*} S^{1}(R)\right), E^{-}=\Lambda^{1}\left(T_{\mathbb{C}}^{*} S^{1}(R)\right)$ and $D^{+}=d, D^{-}=\delta$, and $\Delta^{+}=\delta d=-R^{-2} \partial_{\theta}^{2}$ the scalar Laplacian on functions. Since $g^{i j}=g^{11}=R^{-2}$, the scalar Laplacian on functions is $\Delta^{+}=-\sum_{i j} g^{i j} \partial_{i} \partial_{j}=-\frac{1}{R^{2}} \partial_{\theta}^{2}$.

Denote

$$
f(\theta, t):=(4 \pi t)^{-1 / 2} \exp \left(-R^{2} \theta^{2} / 4 t\right)
$$

For $x \in S^{1}(R)$ and $t>0$, define:

$$
u(x, t)=\sum_{n \in \mathbb{Z}} f(\theta+2 n \pi, t)
$$

where $x=R e^{2 \pi i \theta}$. Note that by definition above, which "logarithm" of $x$ we take is immaterial for the definition of $u$. We first need to check that the series above converges for each $x \in S^{1}(R)$, and each $t>0$. But this is clear, since for $t>0$, the factor of $\exp \left(-n^{2} R^{2} \pi^{2} / t\right)$ will occur in the $n$-th term, and the series will converge very rapidly and indeed uniformly and absolutely. Likewise with the $t$-derivative and all $\theta$-derivatives of the series. So it is permissible to differentiate term-by-term and integrate term by term etc.

Since $\left(\partial_{t}-R^{-2} \partial_{\theta}^{2}\right) f(\theta, t)=0$, it follows that $u(x, t)$ satisfies the heat equation. Note that

$$
\lim _{t \rightarrow 0} t^{-\frac{1}{2}} \exp \left(-R^{2} n^{2} \pi^{2} / t\right)=0 \text { for all } n \neq 0, t>0
$$

Also, for any smooth function $s \in C^{\infty}\left(S^{1}(R)\right)$, we can lift $s$ to a smooth function $\widetilde{s}$ which is compactly supported in say $U=(-2 \pi,+2 \pi)$, with $s\left(R e^{i \theta}\right)=\widetilde{s}(\theta)$ for $\theta \in(-\pi, \pi)$. Then it is easy to check that:

$$
\lim _{t \rightarrow 0} \int_{\pi}^{\infty} \widetilde{s}(\theta) f(\theta, t) R d \theta=0, \quad \lim _{t \rightarrow 0} \int_{-\infty}^{-\pi} \widetilde{s}(\theta) f(\theta, t) R d \theta=0
$$

because $f(\theta, t) \leq(4 \pi t)^{-1 / 2} \exp \left(-R^{2} \pi^{2} / 4 t\right) \leq C \exp (-\alpha / t)$ for $\theta \geq \pi$ and $\theta \leq-\pi$, where $\alpha$ and $C$ are some positive constants.

From the two observations above, we have:

$$
\begin{aligned}
\lim _{t \rightarrow 0}\left(s, u_{t}\right) & =\lim _{t \rightarrow 0} \int_{-\pi}^{\pi} s\left(R e^{i \theta}\right)(4 \pi t)^{-1 / 2} \exp \left(-R^{2} \theta^{2} / 4 t\right) R d \theta \\
& =\lim _{t \rightarrow 0} \int_{-\pi}^{\pi} \widetilde{s}(\theta)(4 \pi t)^{-1 / 2} \exp \left(-R^{2} \theta^{2} / 4 t\right) R d \theta \\
& =\lim _{t \rightarrow 0} \int_{-\infty}^{\infty} \widetilde{s}(x / R)(4 \pi t)^{-1 / 2} \exp \left(-x^{2} / 4 t\right) d \theta=\widetilde{s}(0)=s\left(R e^{i 0}\right)
\end{aligned}
$$

Hence, by the uniqueness statement of the Proposition 11.2.2, applied to that $u(x, t)$ is the fundamental solution to the heat equation with pole at $\left(R e^{i 0}, 1\right)$. By suitably translating the space variable of $u(x, t)$, one can write down the fundamental solution with pole at any other point $(x, 1)$.

Now we can determine all the coefficients $u_{i}(a)$ in the asymptotic expansion of the fundamental solution, where $a=R e^{i 0}=R \in S^{1}(R)$. Since

$$
u(x, t)=\sum_{n \in \mathbb{Z}} f(\theta+2 n \pi, t)
$$

and for $n \neq 0$, the term $f(\theta+2 n \pi, t)$ contains the factor $(4 \pi t)^{-1 / 2} \exp \left(-R^{2} n^{2} / 4 t\right) \leq C e^{-\alpha / t}$ for some $C, \alpha>0$, we find that:

$$
\lim _{t \rightarrow 0} t^{-k} f(\theta+2 n \pi, t)=0 \text { for all } k \geq 0 \text { and } n \neq 0
$$

In other words,

$$
u(x, t) \sim(4 \pi t)^{-1 / 2} \exp \left(-R^{2} \theta^{2} / 4 t\right) \quad \text { as } t \rightarrow 0
$$

Now $R \theta=\delta(a, x)$, the Riemannian distance between $a=R e^{i 0}$ and $x=R e^{i \theta}$ in $S^{1}(R)$. So we find, on comparing the expression for $u(x, t)$ in the Proposition 12.3.5 that

$$
u_{0}(a)=1, \quad u_{i}(a)=0 \quad \text { for all } i \geq 1
$$

This fact has a lot of interesting consequences. Note that the eigenvalues of $\Delta^{+}$are precisely $\lambda_{n}=n^{2} / R^{2}$, and the corresponding (normalised) eigenfunctions are $e_{n}(\theta)=(2 \pi R)^{-1 / 2} e^{i n \theta}$, where $n \in \mathbb{Z}$. From the the construction of the fundamental solution of $\Delta^{+}$(from the heat kernel in (iii) of Proposition 10.1.3 and the fundamental solution in Proposition 11.2.1, we have:

$$
u(x, t)=k_{t}^{+}(x, a)=\sum_{n \in \mathbb{Z}} e^{-t \lambda_{n}} e_{n}^{*}(a) \otimes e_{n}(x)=(2 \pi R)^{-1} \sum_{n \in \mathbb{Z}} e^{-t n^{2} / R^{2}} e^{i n \theta} \quad \text { where } x=R e^{i \theta}, \quad a=R e^{i 0}
$$

Since our asymptotic expansion for $u(x, t)$ just consists of the first term and no others, it follows that the partial sum:

$$
S_{m}(x, t)=(4 \pi t)^{-1 / 2} \exp \left(-R^{2} \theta^{2} / 4 t\right) \text { for all } m \geq 0
$$

Then (iii) of the Theorem 12.3.5 (for $l=0$, say) now tells us that

$$
\left\|(2 \pi R)^{-1} \sum_{n \in \mathbb{Z}} e^{-t n^{2} / R^{2}} e^{i n \theta}-(4 \pi t)^{-1 / 2} \exp \left(-R^{2} \theta^{2} / 4 t\right)\right\|_{0, \infty} \leq C t^{N+1} \quad \text { for } t \in(0, T] m \geq N+1
$$

which implies that

$$
(2 \pi R)^{-1} \sum_{n \in Z} e^{-t n^{2} / R^{2}} e^{i n \theta} \sim(4 \pi t)^{-1 / 2} \exp \left(-R^{2} \theta^{2} / 4 t\right) \text { as } t \rightarrow 0 \text { for each } \theta \in(-\pi, \pi)
$$

Setting $\theta=0$ in the above formula, one obtains Jacobi's asymptotic formula

$$
\begin{equation*}
\sum_{n \in Z} e^{-t n^{2} / R^{2}} \sim(2 \pi R)(4 \pi t)^{-1 / 2}=R \sqrt{\pi / t} \quad \text { as } t \rightarrow 0 \tag{40}
\end{equation*}
$$

So here is a beautiful college-level mathematical formula that uses the asymptotic expansion of the heat kernel on a compact Riemannian manifold for its proof!

Also note that the left hand side of (40) is precisely the trace of the heat operator $e^{-t \Delta^{+}}$, so the Jacobi formula above says that:

$$
\lim _{t \rightarrow 0}(4 \pi t)^{1 / 2}\left(\operatorname{tr} e^{-t \Delta^{+}}\right)=2 \pi R=\operatorname{Vol}\left(S^{1}(R)\right)
$$

Thus the $t \rightarrow 0$ asymptotics of the trace of the heat operator encodes the Riemannian volume of $S^{1}(R)$. Indeed, this is a general fact, as we see below.

Proposition 12.3.7 (You can hear the volume of a manifold). For the scalar Laplacian $\Delta: C^{\infty}(M, \mathbb{C}) \rightarrow$ $C^{\infty}(M, \mathbb{C})$, we have:

$$
\lim _{t \rightarrow 0}(4 \pi t)^{n / 2}\left(\operatorname{tr} e^{-t \Delta}\right)=\operatorname{Vol}(M)
$$

Proof: We first remark that for the scalar Laplacian $\Delta$ on any compact Riemannian manifold, the eigenvalues $\lambda_{n} \geq 0$, because

$$
\lambda_{n}=\left(\Delta e_{n}, e_{n}\right)=\left(\delta d e_{n}, e_{n}\right)=\left(d e_{n}, d e_{n}\right)
$$

where $\left\{e_{n}\right\}$ is an orthonormal basis of smooth eigenfunctions, with $e_{n}$ belonging to the eigenvalue $\lambda_{n}$. Because the operator $\Delta$ is elliptic and formally self-adjoint, the Proposition 8.4.9 shows that $\lambda_{n} \geq C n^{\delta}$, and the existence of the heat kernel:

$$
k_{t}(x, y) \in C^{\infty}(M \times M, \mathbb{C})
$$

defined by $k_{t}(x, y)=\sum_{n} e^{-t \lambda_{n}} e_{n}^{*}(y) e_{n}(x)$ goes through exactly as in (iii) of Proposition 10.1.3. The fundamental solution $w(x, t)$ of the heat equation with the pole $(a, 1)$, with $a \in M$ is as before given by

$$
w^{a}(x, t)=k_{t}(x, a)
$$

Then, since the asymptotic expansion and Duhamel Principle carry over to generalised Laplacians on any bundle $E$ (in this case the trivial bundle $M \times \mathbb{C}$ ), we have the conclusions of Theorem 12.3 .5 in this setting as well, though it was stated for Dirac Laplacians.

We also have:

$$
\operatorname{tr}\left(e^{-t \Delta}\right)=\sum_{n} e^{-t \lambda_{n}}=\int_{M} e^{-t \lambda_{n}} \int_{M} e_{n}^{*}(a) e_{n}(a) d V(a)=\int_{M} k_{t}(a, a) d V(a)
$$

On the other hand, we have by (iii) of Theorem 12.3.5 that:

$$
\left\|k_{t}(-, a)-S_{m}^{a}(-, t)\right\|_{0, \infty}=\left\|w_{t}^{a}(x)-S_{m}^{a}(x, t)\right\|_{0, \infty} \leq C t^{N+1} \text { for } m \geq N+n, \quad t \in(0, T]
$$

where $S_{m}^{a}(x, t)$ is the partial sum of the asymptotic solution $u^{a}(x, t)$ with pole $(a, 1)$. On setting $x=a$, this implies that:

$$
\begin{equation*}
\left|k_{t}(a, a)-(4 \pi t)^{-n / 2} \sum_{k=0}^{m} u_{k}(a) t^{k}\right| \leq C t^{N+1} \quad \text { for all } m \geq N+n, \quad t \in(0, T] \tag{41}
\end{equation*}
$$

Note that $s(a)=\lim _{t \rightarrow 0}\left(s, u^{a}(-, t)\right)$ for all $s \in C^{\infty}(M)$. Letting $f^{a}(x, t)=(4 \pi t)^{-n / 2} \exp \left(-\delta(x, a)^{2} / 4 t\right)$ we also have $s(a)=\lim _{t \rightarrow 0}\left(s, f^{a}(-, t)\right)$ for all $s \in C^{\infty}(M)$. Since $u^{a}(x, t)=f^{a}(x, t)\left(u_{0}(x)+O(t)\right)$, we have

$$
s(a)=\lim _{t \rightarrow 0}\left(s, u^{a}(x, t)\right)=\lim _{t \rightarrow 0}\left(s, f^{a}(-, t) u_{0}(-)\right)=\lim _{t \rightarrow 0}\left(\bar{u}_{0} s, f^{a}(-, t)\right)=\bar{u}_{0}(a) s(a)
$$

which implies that $u_{0}(a)=1$. (In fact, we remarked in the proof of Theorem 12.3.5 that $u_{0}(a)=\widetilde{u}_{0}(a)=v$ from the Proposition 12.3.1, if $u^{a}$ is the asymptotic fundamental solution with pole $\left.(a, v)\right)$. Thus, from the equation (41) above, it follows that:

$$
\left|(4 \pi t)^{n / 2} k_{t}(a, a)-\sum_{k=0}^{m} t^{k} u_{k}(a)\right| \leq C t^{N+1+n / 2} \quad \text { for all } \quad m \geq N+n, t \in(0, T]
$$

which implies that

$$
\lim _{t \rightarrow 0}(4 \pi t)^{n / 2} \operatorname{tr} e^{-t \Delta}=\lim _{t \rightarrow 0}(4 \pi t)^{n / 2} \int_{M} k_{t}(a, a) d V(a)=\lim _{t \rightarrow 0} \int_{M} u_{0}(a) d V(a)=\operatorname{Vol}(M)
$$

and the proposition follows.

## 13. Clifford Algebras and Spin Structures

### 13.1. Clifford Algebras.

Definition 13.1.1. Let $V$ be an inner product space, with a symmetric bilinear form $\langle-,-\rangle$. The Clifford algebra on $V$, denoted $C l(V)$ is an associative unital $\mathbb{R}$-algebra together with an $\mathbb{R}$-linear map:

$$
\phi: V \rightarrow C l(V)
$$

satisfying:
(i): $\phi(v)^{2}=-\langle v, v\rangle .1$ for all $v \in V$.
(ii): If $\psi: V \rightarrow A$ is any $\mathbb{R}$-linear map into an associative unital algebra $A$ satisfying $\psi(v)^{2}=-\langle v, v\rangle 1_{A}$ for all $v \in V$, then there exists a unique $\mathbb{R}$-algebra homomorphism $\widetilde{\psi}$ which makes the diagram:

$$
\begin{array}{lll}
V & \xrightarrow{\phi} & C l(V) \\
\psi & \searrow & \downarrow \widetilde{\psi} \\
& & A
\end{array}
$$

commute.

By the usual abstract nonsense, this universal property makes it unique upto $\mathbb{R}$-algebra isomorphism. To construct it, let $\mathcal{T}(V):=\oplus_{i=0}^{\infty}\left(\otimes^{i} V\right)$ be the full real tensor algebra on $V$. Let $1 \in \otimes^{0} V=\mathbb{R}$ be its identity element. Let $\mathcal{I}$ be the two-sided ideal generated by the set

$$
S:=\left\{v \otimes v+\langle v, v\rangle 1: v \in V=\otimes^{1} V \subset \mathcal{T}(V)\right\}
$$

Define $C l(V)=\mathcal{T}(V) / \mathcal{I}$, and let the map $\phi$ be the composite:

$$
V=\otimes^{1} V \hookrightarrow \mathcal{T}(V) \rightarrow \mathcal{T}(V) / \mathcal{I}
$$

Clearly, by definition, $\phi(v)^{2}=(v \otimes v)(\bmod \mathcal{I})=-\langle v, v\rangle 1$, where $1 \in \mathbb{R}$ is the image of $1 \in \mathcal{T}(V)$. It is trivially checked that the universal property of the definition above holds for $\phi: V \rightarrow C l(V)$. We will denote the product of $a, b \in C l(V)$ as $a . b$ or even $a b$ if no confusion is likely.

Proposition 13.1.2. We have the following facts about the Clifford algebra:
(i): The $\operatorname{map} \phi: V \rightarrow C l(V)$ is injective. Hence we may regard $V$ as a subspace of $C l(V)$.
(ii): With the identification of (i) above,

$$
v \cdot w+w \cdot v=-2\langle v, w\rangle 1 \text { for all } v, w \in V \subset C l(V)
$$

(iii): If $\left\{e_{i}\right\}_{i=1}^{n}$ is any $\mathbb{R}$-basis of $V$, then the products

$$
e_{I}:=e_{i_{1}} \cdot e_{i_{2}} \ldots . e_{i_{k}}
$$

where $I=\left(i_{1}<i_{2}<. .<i_{k}\right)$ is a multiindex with $0 \leq k \leq n$ (and $e_{I}:=1$ for the empty multiindex I with $k=0$ ), constitute an $\mathbb{R}$-basis for $C l(V)$. In particular, $\operatorname{dim} C l(V)=2^{n}$.
(iv): There is a natural $\mathbb{Z}_{2}$-grading on $C l(V)$ defined by setting $C l^{0}(V)$ to be the image of the subspace $\oplus_{k=0}^{\infty}\left(\otimes^{2 k}(V)\right) \subset \mathcal{T}(V)$ and $C l^{1}(V)$ to be the image of $\oplus_{k=0}^{\infty}\left(\otimes^{2 k+1}(V)\right) \subset \mathcal{T}(V)$. With this grading $C l(V)$ is a so-called superalgebra, i.e. satisfies:

$$
C l^{i}(V) \cdot C l^{j}(V) \subset C l^{k}(V) \text { where } k=i+j(\bmod 2)
$$

(v): There is a canonical vector space isomorphism (not an algebra homomorphism):

$$
C l(V) \rightarrow \Lambda^{*} V
$$

which takes $v . w$ to $v \wedge w$ for all $v, w \in V$.
(vi): For the identically zero inner product $\langle-,-\rangle \equiv 0$, the Clifford algebra $C l(V)$ is the exterior algebra $\Lambda^{*} V$.

Proof: To see (i), define the degree deg $x$ of an element $x \in \mathcal{T}(V)$ by expanding into homogeneous components

$$
x=\oplus_{i} x_{i}, \quad x_{i} \in \otimes^{i} V
$$

to be the largest $i$ such that $x_{i} \neq 0$. Clearly, $\operatorname{deg}(x \otimes y)=\operatorname{deg} x+\operatorname{deg} y$, and hence the degree of every element in the ideal $\mathcal{I}$ is at least 2. Thus $V \cap \mathcal{I}=\{0\}$ in $\mathcal{T}(V)$, and the map $\phi: V \rightarrow C l(V)$ is injective. This proves (i). We may therefore write $v$ instead of $\phi(v)$ for $v \in V$.

To see (ii), note that for $v, w \in V \subset C l(V)$, we have by the definition of $C l(V)$ :

$$
-(\langle v, v\rangle+\langle w, w\rangle+2\langle v, w\rangle) 1=-\langle v+w, v+w\rangle 1=(v+w)^{2}=v \cdot v+w \cdot w+v \cdot w+w \cdot v
$$

from which it follows that $v \cdot w+w \cdot v=-2\langle v, w\rangle 1$.
To see (iii), we use (ii) to see that $e_{I}$ of the form stated are a spanning set for $C l(V)$, since any word $e_{j_{1}} \cdot e_{j_{2}} \ldots . . e_{j_{k}}$ of any length may be reduced, by using the commutation relations:

$$
e_{i} \cdot e_{j}+e_{j} \cdot e_{i}=-2\left\langle e_{i}, e_{j}\right\rangle 1
$$

to a word of length at most $n$. Their linear independence is left as an exercise. (iv), (v) and (vi) are also straightforward, and their proof is omitted.

Notation: From now on, when we write $C l(V)$, it will be understood that $V$ is an inner product space with a positive definite inner product $\langle-,-\rangle$. Hence, we may always choose an orthonormal basis $\left\{e_{i}\right\}_{i=1}^{n}$ of $V$, and the commutation relations for the basis elements will read:

$$
e_{i} \cdot e_{j}+e_{j} . e_{i}=-2 \delta_{i j} \quad 1 \leq i, j \leq n
$$

Example 13.1.3. If we take $V=\mathbb{R}$, with its usual euclidean inner product $\langle x, y\rangle=x y$, then $C l(\mathbb{R})=\mathbb{C}$, for it is generated as an $\mathbb{R}$-algebra by $e_{1}$ satisfying $e_{1}^{2}=-1$.

If we take $V=\mathbb{R}^{2}$ with its usual euclidean inner product, then $C l(V)$ is generated as an $\mathbb{R}$-algebra by $\left\{e_{1}, e_{2}\right\}$, satisfying:

$$
e_{1}^{2}=e_{2}^{2}=-1, \quad e_{1} e_{2}=-e_{2} e_{1}
$$

Setting $e_{1}=i, e_{2}=j, e_{1} e_{2}=k$, we find that

$$
C l(V)=\mathbb{R} .1 \oplus \mathbb{R} . i \oplus \mathbb{R} . j \oplus \mathbb{R} . k
$$

subject to the relations $i^{2}=j^{2}=k^{2}=-1$, and $i j=k, j k=i, k i=j$. Thus $C l\left(\mathbb{R}^{2}\right)=\mathbb{H}$, the (noncommutative) algebra of quaternions.

Exercise 13.1.4 (Some Clifford Algebras).
(i): Show that for $V=\mathbb{R}^{3}$, with its usual euclidean inner product, we have $C l(V) \simeq \mathbb{H} \oplus \mathbb{H} \eta$ where $\eta:=e_{1} e_{2} e_{3}$. The first summand $\mathbb{H}$ is the span of $1, i:=e_{1} e_{2}, j:=e_{2} e_{3}, k:=e_{3} e_{1}$, and the second summand is the span of $\eta, i \eta, j \eta, k \eta$. Multiplication is given by:

$$
(a+b \eta)(c+d \eta)=(a c+b d)+(a d+b c) \eta \quad a, b, c, d \in \mathbb{H}
$$

(ii): Prove that $C l\left(\mathbb{R}^{2}\right) \otimes_{\mathbb{R}} \mathbb{C}=\mathbb{H} \otimes_{\mathbb{R}} \mathbb{C}=\mathbb{C}(2)$, the algebra of $2 \times 2$ complex matrices. Explicitly, the isomorphism is given by:

$$
i \otimes 1 \mapsto\left(\begin{array}{cc}
\sqrt{-1} & 0 \\
0 & -\sqrt{-1}
\end{array}\right) ; \quad j \otimes 1 \mapsto\left(\begin{array}{cc}
0 & \sqrt{-1} \\
\sqrt{-1} & 0
\end{array}\right) ; \quad k \otimes 1 \mapsto\left(\begin{array}{cc}
0 & -1 \\
1 & 0
\end{array}\right)
$$

where the matrices on the right are the Pauli spin matrices.

Remark 13.1.5. It is possible to write down a complete list of all the real Clifford algebras $C l\left(\mathbb{R}^{n}\right)$, because of the remarkable periodicity theorem which states that:

$$
C l\left(\mathbb{R}^{n+8}\right)=C l\left(\mathbb{R}^{n}\right) \otimes_{\mathbb{R}} \mathbb{R}(16)
$$

where $\mathbb{R}(n)$ denotes the matrix algebra of $n \times n$ real matrices. This reduces us to finding out $C l\left(\mathbb{R}^{n}\right)$ for $n=1, . ., 8$, whose list is as below:

$$
\begin{array}{ccccccccc}
n: & 1 & 2 & 3 & 4 & 5 & 6 & 7 & 8 \\
C l\left(\mathbb{R}^{n}\right): & \mathbb{C} & \mathbb{H} & \mathbb{H} \oplus \mathbb{H} \eta & \mathbb{H}(2) & \mathbb{C}(4) & \mathbb{R}(8) & \mathbb{R}(8) \oplus \mathbb{R}(8) & \mathbb{R}(16)
\end{array}
$$

For a proof of this fact, see the paper "Clifford Modules" by Atiyah- Bott-Shapiro.

We need a little more machinery associated with a Clifford algebra. The first is the involution $*$ defined as follows:

Definition 13.1.6 (The involution $*)$. Let $V$ be a real positive definite inner product space, and $C l(V)$ its Clifford algebra. There is an involution $*$ on the full tensor algebra $\mathcal{T}(V)$ whose effect on decomposable tensors is:

$$
\left(a_{1} \otimes a_{2} \otimes \ldots \otimes a_{k}\right)^{*}=a_{k} \otimes a_{k-1} \otimes \ldots \otimes a_{2} \otimes a_{1}
$$

This involution clearly preserves the set $S=\{v \otimes v+\langle v, v\rangle 1: v \in V\}$ defined in the beginning of this section, and since $(\alpha \otimes \beta)^{*}=\beta^{*} \otimes \alpha^{*}$, we see that $*$ preserves the two-sided ideal $\mathcal{I}$ generated by $S$. Hence it descends to an involution of $C l(V)=\mathcal{T}(V) / \mathcal{I}$. If we let $\left\{e_{i}\right\}_{i=1}^{n}$ be an orthonormal basis for $V$ with respect to $\langle-,-\rangle$, then for the basis of $C l(V)$ introduced in (iii) of Proposition 13.1.2, we have:

$$
\left(e_{i_{1}} \cdot e_{i_{2}} \ldots . . e_{i_{k}}\right)^{*}=e_{i_{k}} \cdot e_{i_{k-1}} \ldots \ldots . e_{i_{2}} \cdot e_{i_{1}}
$$

Clearly, * is the unique involution of $C l(V)$ satisfying:

$$
v^{*}=v \text { for } v \in V \subset C l(V) \text { and }(a . b)^{*}=b^{*} . a^{*} \text { for all } a, b \in C l(V)
$$

As an exercise, the reader may explicitly compute the involution $*$ on the Clifford algebras $C l(\mathbb{R}), C l\left(\mathbb{R}^{2}\right)$ and $C l\left(\mathbb{R}^{3}\right)$ that were determined above.

Definition 13.1.7 (Supercommutators). For a superalgebra $A=A^{0} \oplus A^{1}$ (such as the Clifford algebra), define the supercommutator of two homogeneous elements $x, y \in A$ by:

$$
[x, y]_{s}:=x y-(-1)^{(\operatorname{deg} x)(\operatorname{deg} y)} y x
$$

Extend to arbitrary elements of $A$ by linearity in each slot. For example, if $A=\Lambda^{*} V=\Lambda^{e v} \oplus \Lambda^{o d d}$, then the supercommutator of any two elements is 0 .

For a superalgebra $A$ as above, define the supercentre of $A$ by:

$$
Z_{s}(A):=\left\{x:[x, y]_{s} \equiv 0 \text { for all } y \in A\right\}
$$

Lemma 13.1.8. Let $V$ be a positive definite inner product space. Then the supercentre of the Clifford algebra $\mathrm{Cl}(\mathrm{V})$ consists of the scalars $\mathbb{R} .1$.

Proof: It is clear that $\mathbb{R} .1 \subset Z_{s}(C l(V))$, since the supercommutator of any scalar with any element is just the usual commutator, and the scalars commute with everything in $C l(V)$. On the other hand, we claim that if $[x, v]_{s}=0$ for all $v \in V$, then $x$ is a scalar. For, write $x=x_{0}+x_{1}$, with $x_{i} \in C l^{i}(V)$ in terms of its homogeneous components. Then $[x, v]_{s}=\left[x_{0}, v\right]_{s}+\left[x_{1}, v\right]_{s}$, and since $v$ has homogeneous degree 1 , we have $\left[x_{0}, v\right]_{s}$ is homogeneous of degree 1 , and $\left[x_{1}, v\right]_{s}$ is homogeneous of degree 0 . Thus both $\left[x_{0}, v\right]_{s}$ and $\left[x_{1}, v\right]$ are individually 0 , for all $v \in V$. So it is enough to prove that if $x \in Z_{s}(C l(V))$ is homogeneous and $[x, v]_{s}=0$ for all $v \in V$, then $x=\lambda .1$.

Let $\left\{e_{i}\right\}$ be an orthonormal basis of $V$. Write the homogeneous element $x$ as $x=a+e_{1} . b$, where $a$ and $b$ are independent of $e_{1}$ (by using the basis $e_{I}$ of $C l(V)$ constructed in (iii) of 13.1.2). Then $\operatorname{deg} a=\operatorname{deg} x=\operatorname{deg} b+1$. Hence

$$
\begin{aligned}
{\left[x . e_{1}\right]_{s} } & =\left[a, e_{1}\right]_{s}+\left[e_{1} b, e_{1}\right]_{s}=a e_{1}-(-1)^{\operatorname{deg} a} e_{1} a+e_{1} b e_{1}-(-1)^{\operatorname{deg} a} b e_{1}^{2} \\
& =a e_{1}+(-1)^{\operatorname{deg} a}(-1)^{\operatorname{deg} a} a e_{1}+(-1)^{\operatorname{deg} b} b e_{1}^{2}+(-1)^{\operatorname{deg} a+1} b e_{1}^{2} \\
& =(-1)^{\operatorname{deg} b} 2 b e_{1}^{2}=(-1)^{\operatorname{deg} a} 2 b
\end{aligned}
$$

So that $\left[x, e_{1}\right]_{s}=0$ implies that $b=0$. Thus $x=a+e_{1} b=a$ is independent of $e_{1}$. By the same reasoning, it is independent of $e_{i}$ for all $i$, and hence a scalar. This proves the lemma.

Remark 13.1.9. Note that the usual centre of $C l(V)$ is usually much larger than the scalars. For example, in $C l(\mathbb{R})=\mathbb{C}$, the centre is all of $C l(\mathbb{R})$.
13.2. The Groups Pin and Spin. Let $V=\mathbb{R}^{n}$ with its usual positive definite euclidean inner product. Recall the involution $*$ introduced in Definition 13.1.6.

Definition 13.2.1 $(\operatorname{Pin}(n)$ and $\operatorname{Spin}(n))$. Define the group

$$
\operatorname{Pin}(n):=\left\{x \in C l(V): x \text { is homogeneous, } x x^{*}=x^{*} x=(-1)^{\operatorname{deg} x}, x V x^{*} \subset V\right\}
$$

Further define

$$
\operatorname{Spin}(n)=\operatorname{Pin}(n) \cap C l^{0}(V)
$$

Note that by definition, $x^{*}=x^{-1}$ for all $x \in \operatorname{Spin}(n)$. Also, since the group $C l^{\times}(V)$ of invertible elements in $C l(V)$ is an open subset of the euclidean space $C l(V)$, and the operation of Clifford multiplication is algebraic (by using a basis) and hence smooth, it follows that the closed conditions defining $\operatorname{Pin}(n)$ and $\operatorname{Spin}(n)$ make them closed subgroups of $C l^{\times}(V)$. Hence both are Lie groups by Cartan's theorem.

Note that by definition, there is an action of $\operatorname{Pin}(n)$ on $V=\mathbb{R}^{n}$ given by:

$$
\rho: \operatorname{Pin}(n) \rightarrow G L(n, \mathbb{R})
$$

where $\rho(x) v=x v x^{*}$. We have the following proposition.

Proposition 13.2.2 (Basic facts on $\operatorname{Pin}(n), \operatorname{Spin}(n)$ and $\rho)$.
(i): $\rho(\operatorname{Pin}(n)) \subset O(n, \mathbb{R})$. The sequence:

$$
1 \rightarrow \mathbb{Z}_{2} \rightarrow \operatorname{Pin}(n) \xrightarrow{\rho} O(n, \mathbb{R}) \rightarrow 1
$$

is exact. $\left(\right.$ Here $\left.\mathbb{Z}_{2}=\{+1,-1\} \subset \operatorname{Spin}(n) \subset \operatorname{Pin}(n)\right)$
(ii): Any element $x \in \operatorname{Pin}(n)$ may be expressed as a Clifford product:

$$
x=v_{1} v_{2} \ldots v_{k}
$$

where $v_{i}$ are some unit vectors in $V$.
(iii): $\rho(\operatorname{Spin}(n)) \subset S O(n)$ and the sequence

$$
1 \rightarrow \mathbb{Z}_{2} \rightarrow \operatorname{Spin}(n) \xrightarrow{\rho} S O(n) \rightarrow 1
$$

is exact. An element $x \in \operatorname{Spin}(n)$ iff it is a Clifford product $x=v_{1} \ldots v_{k}$ with $v_{i}$ unit vectors in $V$, and $k$ is even.
(iv): $\operatorname{Spin}(n)$ is connected.
(v): The Lie algebra map $\dot{\rho}$ maps the element $\frac{1}{4} \sum_{i \neq j} a_{i j} e_{i} e_{j} \in C l(V)$ to skew-symmetric matrix $\left[a_{i j}\right]$ in the lie algebra $\operatorname{Lie}(\operatorname{Spin}(n))=\mathfrak{s o}(n)$ thus identifying the above Lie algebra with a subspace of $C l(V)$.

Proof: Note that for $v \in V \subset C l(V)$, we have $\|v\|^{2} .1=-v^{2}$. Further, for $x \in \operatorname{Pin}(n)$, we have $\rho(x) v \in V$ as well, so that

$$
\begin{aligned}
\|\rho(x) v\|^{2} \cdot 1 & =-(\rho(x) v)^{2}=-\left(x v x^{*} x v x^{*}\right)=-(-1)^{\operatorname{deg} x} x v^{2} x^{*} \\
& =(-1)^{\operatorname{deg} x} x\|v\|^{2} \cdot 1 \cdot x^{*}=(-1)^{2 \operatorname{deg} x}\|v\|^{2} \cdot 1=\|v\|^{2} \cdot 1
\end{aligned}
$$

which proves the first assertion of (i).
If $\rho(x)=I d_{V}$ for $x \in \operatorname{Pin}(n)$, then $x v x^{*}=v$ for all $v \in V$. That is, $x v=(-1)^{\operatorname{deg} x} v x$. That is, the supercommutator $[x, v]_{s}=0$ for all $v \in V$. In the proof of Lemma 13.1.8, we remarked that this forces $x=\lambda 1$ and $\operatorname{deg} x=0$. Thus $x^{*}=\lambda .1$, and $x x^{*}=(-1)^{\operatorname{deg} x} .1=1$ implies that $\lambda^{2}=1$, or $\lambda= \pm 1$. So ker $\rho=\mathbb{Z}_{2}$.

If we let $\left\{e_{i}\right\}_{i=1}^{n}$ be an orthonormal basis for $V=\mathbb{R}^{n}$, we note that $e_{i}^{*}=e_{i}$, and hence $e_{i} e_{i}^{*}=-1=$ $(-1)^{\operatorname{deg} e_{1}}$.1. Clearly $\rho\left(e_{i}\right) e_{i}=e_{i} e_{i} e_{i}=-e_{i}$ Also we have:

$$
\rho\left(e_{i}\right) e_{j}=e_{i} e_{j} e_{i}=-e_{i}^{2} e_{j}=e_{j} \text { for } j \neq i
$$

Thus $e_{i} V e_{i}^{*} \subset V$, and $e_{i} \in \operatorname{Pin}(n)$ for all $i$. The above calculation shows that $\rho\left(e_{i}\right)$ is orthogonal reflection about the hyperplane $\left(\mathbb{R} e_{i}\right)^{\perp}$ in $V$. Since each unit vector $v \in V$ can be completed to an orthonormal basis, it follows that every unit vector $v \in V \subset C l(V)$ is in $\operatorname{Pin}(n)$, and $\rho(v)$ is just the reflection about the hyperplane $(\mathbb{R} v)^{\perp} \subset V$. Since the group $O(n, \mathbb{R})$ is generated by relections about hyperplanes, it follows that $\rho: \operatorname{Pin}(n) \rightarrow O(n, \mathbb{R})$ is surjective. This proves the exact sequence of (i), and (i) follows.

For any $x \in \operatorname{Pin}(n)$, we have $\rho(x) \in O(n, \mathbb{R})$, and indeed we saw in the last paragraph that $\rho(x)=\rho\left(v_{1} \ldots v_{n}\right)$ for some unit vectors $v_{i}$. This means that $x= \pm v_{1} \ldots v_{k}=\left( \pm v_{1}\right) . . v_{k}$, and (ii) follows.

Since $\operatorname{deg}\left(v_{1} \ldots v_{k}\right)=k \bmod 2$, from (ii) it follows that an element $x \in \operatorname{Pin}(n)$ lies in $C l^{0}(V)$ iff $x$ can be expressed as a Clifford product of an even number of unit vectors. Since the set of elements in $O(n, \mathbb{R})$ expressible as products of an even number of reflections is precisely $S O(n)$, all the assertions of (iii) follow trivially from (i) and (ii).

Since an element $x \in \operatorname{Spin}(n)$ is expressible as a Clifford product

$$
x=v_{1} \ldots v_{2 m}
$$

where $v_{i} \in V$ are unit vectors, to connect $x$ by a path in $\operatorname{Spin}(n)$ to the identity element 1 , it is enough to connect the pairwise doublet elements $v_{2 i-1} v_{2 i}$ to 1 by a path $y_{i}(t)$ in $\operatorname{Spin}(n)$ (so that $\prod_{i=1}^{m} y_{i}(t)$ will be required path connecting $x$ to 1 ). So let $v, w$ be unit vectors in $V$, and let us find a path in $\operatorname{Spin}(n)$ connecting $v . w$ to 1 . If $w$ is linearly dependent on $v$, then $v . w= \pm 1$, and it is trivial to connect it to 1 . So assume $w$ and $v$ are linearly indpendent. Let $e$ be a unit vector in the span $\mathbb{R} v+\mathbb{R} w$ which is perpendicular to $v$. Then letting $e(t)$ be a path of unit vectors in $V$ joining $w$ to $e$, we see that $v . w$ can be joined to $v . e$ by the path $v . e(t)$ in $\operatorname{Spin}(n)$.

Hence, we may assume without loss of generality that the unit vectors $v$ and $w$ are orthogonal, and exhibit a path joining $v . w$ to 1 in $\operatorname{Spin}(n)$. Consider the path:

$$
x(t)=(\cos t) 1+(\sin t) v \cdot w
$$

we clearly have $x^{*}(t)=(\cos t) 1+(\sin t) w \cdot v$ and

$$
\begin{aligned}
x(t) x^{*}(t) & =[(\cos t) 1+(\sin t) v \cdot w][(\cos t) 1+(\sin t) w \cdot v] \\
& =\left(\cos ^{2} t+\sin ^{2} t v \cdot w \cdot w \cdot v\right) 1+\sin t \cos t(v \cdot w+w \cdot v)=\left(\cos ^{2} t+(-1)^{2} \sin ^{2} t\right) 1+\cos t \sin t(2\langle v, w\rangle) \\
& =\left(\cos ^{2} t+\sin ^{2} t\right) 1+\cos t \sin t(0)=1
\end{aligned}
$$

It is also easy to check that $x(t) V x(t)^{*} \subset V$, (In fact $\rho(v . w)$ is some planar rotation, and $\rho(x(t))$ joins that planar rotation to the identity element of $S O(n)$ ). Thus $x(t)$ is the required path joining $v . w$ to 1 , and (iv) follows.

To see (v), let $\left\{e_{i}\right\}_{i=1}^{n}$ be an orthonormal basis for $V=\mathbb{R}^{n}$. For $i \neq j$, note that

$$
\left(e_{i} e_{j}\right)^{2}=-e_{i} e_{i} e_{j} e_{j}=-(-1)(-1)=-1
$$

so that $\left(e_{i} e_{j}\right)^{2 m}=(-1)^{m}$ and $\left(e_{i} e_{j}\right)^{2 m+1}=(-1)^{m} e_{i} e_{j}$. Hence if we take the exponential:

$$
\begin{aligned}
\exp \left(t e_{i} e_{j}\right) & =1+t e_{i} e_{j}+\frac{t^{2}}{2!}\left(e_{i} e_{j}\right)^{2}+\ldots \\
& =\left(1-\frac{t^{2}}{2!}+\frac{t^{4}}{4!}+. .\right) 1+\left(t-\frac{t^{3}}{3!}+\frac{t^{5}}{5!}+. .\right) e_{i} e_{j} \\
& =(\cos t) 1+(\sin t) e_{i} e_{j}
\end{aligned}
$$

We have seen above that this is precisely the path joining 1 to $e_{i} e_{j}$ in $\operatorname{Spin}(n)$. We can compute its derivative at $t=0$ as

$$
{\left.\frac{d\left(\exp \left(t e_{i} e_{j}\right)\right.}{d t} \right\rvert\, t=0}^{d t}=\left(-\sin t .1+(\cos t) e_{i} e_{j}\right)_{\mid t=0}=e_{i} e_{j}
$$

which shows that all these elements $e_{i} e_{j}$ for $i \neq j$ lie in the Lie algebra of $\operatorname{Spin}(n)$. Since the span of $\left\{e_{i} e_{j}\right\}_{i<j}$ is of dimension $\frac{n(n-1)}{2}$, which is precisely the dimension of $\mathfrak{s o}(n)=\operatorname{Lie}(\operatorname{Spin}(n))$, it follows that this last Lie algebra is the linear span of $\left\{e_{i} e_{j}\right\}_{i<j}$. To get the isomorphism even more explicitly, note that

$$
\begin{aligned}
\rho\left(\exp \left(t e_{i} e_{j}\right)\right) e_{i} & =\left(\cos t+\sin t e_{i} e_{j}\right) e_{i}\left(\cos t+\sin t e_{j} e_{i}\right)=\left(\cos ^{2} t\right) e_{i}+\left(\sin ^{2} t\right)\left(e_{i} e_{j} e_{i} e_{j} e_{i}\right)+2 \sin t \cos t\left(e_{i} e_{j} e_{i}\right) \\
& =\left(\cos ^{2} t-\sin ^{2} t\right) e_{i}+(2 \sin t \cos t) e_{j}=(\cos 2 t) e_{i}+(\sin 2 t) e_{j}
\end{aligned}
$$

Similarly, one verifies that

$$
\rho\left(\exp \left(t e_{i} e_{j}\right)\right) e_{j}=(-\sin 2 t) e_{i}+(\cos 2 t) e_{j}
$$

and also that since $e_{i} e_{j}$ commutes with $e_{k}$ for all $k \neq i, k \neq j$, we have $\rho\left(\exp \left(t e_{i} e_{j}\right)\right) e_{k}=e_{k}$ for all $k \neq i, k \neq j$. As a consequence,

$$
\rho\left(\exp \left(t e_{i} e_{j}\right)\right)=\left(\begin{array}{cc}
\cos 2 t & -\sin 2 t \\
\sin 2 t & \cos 2 t
\end{array}\right)
$$

where the matrix on the right is a rotation in the $e_{i}, e_{j}$ plane of $V$. Thus

$$
\dot{\rho}\left(e_{i} e_{j}\right)=\frac{d\left(\rho\left(\exp \left(t e_{i} e_{j}\right)\right)\right.}{d t}{ }_{\mid t=0}=2\left(\begin{array}{cc}
0 & -1 \\
1 & 0
\end{array}\right)
$$

So that $\dot{\rho}\left(\sum_{i<j} a_{i j} e_{i} e_{j}\right)=2\left[a_{i j}\right]$ and thus

$$
\dot{\rho}\left(\frac{1}{4} \sum_{i \neq j} a_{i j} e_{i} e_{j}\right)=\left[a_{i j}\right]
$$

for a skew-symmetric real matrix $\left[a_{i j}\right]$. This proves (v), and the proposition follows.

Example 13.2.3. It is easy to check that $\operatorname{Pin}(1)=\mathbb{Z}_{2}$, and $\operatorname{Spin}(1)=\{1\}$. Note that $C l\left(\mathbb{R}^{2}\right)=\mathbb{H}$, and the operation $*$ on $C l\left(\mathbb{R}^{2}\right)$ is the map defined by $i^{*}=e_{1}^{*}=e_{1}=i, j^{*}=e_{2}^{*}=e_{2}=j$ and $k^{*}=\left(e_{1} e_{2}\right)^{*}=e_{2} e_{1}=$ $-e_{1} e_{2}=-k$. Also $C l^{0}\left(\mathbb{R}^{2}\right)=\mathbb{R} 1+\mathbb{R} k$, and $C l^{1}\left(\mathbb{R}^{2}\right)=\mathbb{R} i+\mathbb{R} j$. If $x=a 1+b k \in C l^{0}$ (resp. $x=a i+b j \in C l^{1}$ ), then $x x^{*}=a^{2}+b^{2}$ (resp. $-a^{2}-b^{2}$ ), and also $x\left(\alpha e_{1}+\beta e_{2}\right) x^{*} \in V$ in both cases. Hence $\operatorname{Pin}(2)=S^{1} \times \mathbb{Z}_{2}$, and $\operatorname{Spin}(2)=S^{1}=\left\{a 1+b k: a^{2}+b^{2}=1\right\}$. It is also verified easily that for $x=(\cos t) 1+(\sin t) k \in \operatorname{Spin}(2)$ :

$$
\rho((\cos t) 1+(\sin t) k)=\left(\begin{array}{cc}
\cos 2 t & \sin 2 t \\
-\sin 2 t & \cos 2 t
\end{array}\right)
$$

so that the map $\rho: \operatorname{Spin}(2)=S^{1} \rightarrow S O(2)=S^{1}$ is just the squaring map.
Finally, since $C l^{0}\left(\mathbb{R}^{3}\right)=\mathbb{H}=\mathbb{R} 1+\mathbb{R} i+\mathbb{R} j+\mathbb{R} k$ where $i=e_{1} e_{2}, j=e_{2} e_{3}$ and $k=e_{3} e_{1}$ (see Exercise 13.1.4), we have $1^{*}=1, i^{*}=-i, j^{*}=-j$ and $k^{*}=-k$. Thus, for a quaternion $x=a 1+b i+c j+d k \in C l^{0}$, $x^{*}=a 1-b i-c j-d k$, the conjugate quaternion, and $x x^{*}=1$ implies $a^{2}+b^{2}+c^{2}+d^{2}=1$, viz., $x$ is a unit length quaternion. It is again clear that $x V x^{*} \subset V$, so that $\operatorname{Spin}(3)$ is the group of unit quaternions, homeomorphic to $S^{3}$. Further, one easily computes that the homomorphism $\rho: \operatorname{Spin}(3) \rightarrow S O(3)$ is given by:

$$
\rho(a+b i+c j+d k)=\left(\begin{array}{ccc}
\left(a^{2}+c^{2}\right)-\left(b^{2}+d^{2}\right) & 2(c d-a b) & 2(b c+a d) \\
2(a b+c d) & \left(a^{2}+d^{2}\right)-\left(b^{2}+c^{2}\right) & 2(b d-a c) \\
2(b c-a d) & 2(b d-a c) & \left(a^{2}+b^{2}\right)-\left(c^{2}+d^{2}\right)
\end{array}\right)
$$

Note that $\rho(-x)$ and $\rho(x)$ are the same, as they should be. Also, recalling the central element $\eta=e_{1} e_{2} e_{3} \in$ $C l^{1}\left(\mathbb{R}^{3}\right)$ we see that $\eta^{*}=e_{3} e_{2} e_{1}=-\eta$, so that $\rho(\eta) e_{i}=-\eta e_{i} \eta=-\eta \eta e_{i}=-e_{i}$. Thus $\rho(\eta)=-I \in O(3, \mathbb{R})$, and also $\operatorname{Pin}(3)=\operatorname{Spin}(3) \coprod \operatorname{Spin}(3) \eta=\operatorname{Spin}(3) \times \mathbb{Z}_{2}$, since $\eta$ is central.
13.3. Spin structures on manifolds. Let $M$ be a connected oriented Riemannian manifold of dimension $n$. There is the orthonormal oriented frame bundle:

$$
S O(n) \rightarrow P \rightarrow M
$$

whose fibre is

$$
P_{x}=\left\{\text { oriented orthonormal frames in } T_{x} M\right\} \simeq S O(n)
$$

Definition 13.3.1. We say that $M$ has a spin structure if there exists a principal $\operatorname{Spin}(n)$ bundle $\widetilde{P} \rightarrow M$ and a double-covering map $\rho: \widetilde{P} \rightarrow P$ so that the following diagram commutes:

$$
\begin{array}{cccc}
\operatorname{Spin}(n) & \rightarrow & \widetilde{P} & \rightarrow M \\
\rho_{x} \downarrow & & \downarrow \rho & I d \downarrow \\
S O(n) & \rightarrow & P & \rightarrow M
\end{array}
$$

There is a handy criterion for the existence of a spin structure on $M$, as also a way of parametrising all possible spin structures on $M$. Namely,
Proposition 13.3.2. The oriented Riemannian manifold $M$ as above has a spin structure iff the second Stiefel-Whitney class $w_{2}(M)=0$. Furthermore, if there does exist a spin structure on $M$, then the set of all spin-structures on $M$ is in bijective correspondence with $H^{1}\left(M, \mathbb{Z}_{2}\right)$.

Proof: Let $S O(n) \xrightarrow{i} P \rightarrow M$ be the principal $S O(n)$ bundle as above, and consider the Serre spectral sequence of this fibration with $\mathbb{Z}_{2}$ coefficients. Then:

$$
E_{2}^{p, q}=H^{p}\left(M, H^{q}\left(S O(n), \mathbb{Z}_{2}\right)\right) \Rightarrow H^{p+q}\left(P, \mathbb{Z}_{2}\right)
$$

Note that we have the exact sequence:

$$
0 \rightarrow E_{3}^{0,1} \rightarrow E_{2}^{0,1} \xrightarrow{d_{2}} E_{2}^{2,0} \rightarrow E_{3}^{2,0} \rightarrow 0
$$

since $d_{2}: E_{2}^{-2,2} \rightarrow E_{2}^{0,1}$ and $d_{2}: E_{2}^{2,0} \rightarrow E_{2}^{4,-1}$ are zero maps, the spectral sequence being first quadrant. For the same reason, $d_{r}: E_{r}^{0,1} \rightarrow E_{r}^{r, 2-r}$ and $d_{r}: E_{r}^{2,0} \rightarrow E_{r+2,1-r}$ are zero maps for $r \geq 3$, and so $E_{3}^{0,1}=E_{\infty}^{0,1}$ and $E_{3}^{2,0}=E_{\infty}^{2,0}$. Since

$$
E_{\infty}^{0,1}=F^{0}\left(H^{1}\left(P, \mathbb{Z}_{2}\right)\right) / F^{1}\left(H^{1}\left(P, \mathbb{Z}_{2}\right)\right)
$$

and $F^{0}\left(H^{1}\left(E, \mathbb{Z}_{2}\right)\right)=H^{1}\left(E, \mathbb{Z}_{2}\right)$, we have a natural quotient map $H^{1}\left(E, \mathbb{Z}_{2}\right) \rightarrow E_{\infty}^{0,1}=E_{3}^{0,1}$. Noting that $E_{2}^{0,1}=H^{0}\left(M, H^{1}\left(S O(n), \mathbb{Z}_{2}\right)\right)=H^{1}\left(S O(n), \mathbb{Z}_{2}\right)$ and $E_{2}^{2,0}=H^{2}\left(M, H^{0}\left(S O(n), \mathbb{Z}_{2}\right)\right)=H^{2}\left(M, \mathbb{Z}_{2}\right)$, we have the exact sequence:

$$
\begin{equation*}
H^{1}\left(P, \mathbb{Z}_{2}\right) \xrightarrow{i^{*}} H^{1}\left(S O(n), \mathbb{Z}_{2}\right) \xrightarrow{\delta} H^{2}\left(M, \mathbb{Z}_{2}\right) \tag{42}
\end{equation*}
$$

The first map is $i^{*}$ by applying the functoriality of the Serre spectral sequence to the inclusion of a point into $M$, and called an "edge homomorphism". The image $\delta(1)$ of the generator $1 \in H^{1}\left(S O(n), \mathbb{Z}_{2}\right)=\mathbb{Z}_{2}$ is the Stiefel-Whitney class $w_{2}(M)$. Also, $\operatorname{ker} \delta=\operatorname{Im} i^{*}$ by exactness of (42).
$M$ has a spin structure iff there is a double cover $\widetilde{P} \xrightarrow{\rho} P$ which makes the diagram of Definition 13.3.1 commute.

Double covers of $P$ are in 1-1 correspondence with index 2 subgroups of $\pi_{1}(P)$, which is in bijective correspondence with $\operatorname{hom}_{\mathbb{Z}}\left(\pi_{1}(P), \mathbb{Z}_{2}\right)$. But this last group is precisely $H^{1}\left(P, \mathbb{Z}_{2}\right)$. Hence $\rho$ is an element of $H^{1}\left(P, \mathbb{Z}_{2}\right)$. Since the diagram of Definition 13.3 .1 commutes, the restriction of the double cover $\rho: P \rightarrow M$ to a point $x \in M$ must correspond to the nontrivial double-cover $\rho_{x}: \operatorname{Spin}(n) \rightarrow S O(n)$. Now the double cover $\rho_{x}$ is represented by the unique generating element $1 \in \operatorname{hom}_{\mathbb{Z}}\left(\pi_{1}(S O(n)), \mathbb{Z}_{2}\right)=H^{1}\left(S O(n), \mathbb{Z}_{2}\right)=\mathbb{Z}_{2}$. By functoriality, it follows that $i^{*}(\rho)=1$. Now, there exists such a $\rho \in H^{1}\left(P, \mathbb{Z}_{2}\right)$ satisfying $i^{*}(\rho)=1$ iff $\delta(1)=w_{2}(M)=0$, by the exactness of the sequence (42). This proves that $M$ has a spin structure iff $w_{2}(M)=0$, and the first part of the proposition follows.

From the previous para, it also follows that spin structures on $M$ are in 1-1 correspondence with the inverse image $\left(i^{*}\right)^{-1}(1) \in H^{1}\left(P, \mathbb{Z}_{2}\right)$, where $1 \in H^{1}\left(S O(n), \mathbb{Z}_{2}\right)=\mathbb{Z}_{2}$ is the generator. But $\left(i^{*}\right)^{-1}(1)$ is the set-theoretic complement of $\left(i^{*}\right)^{-1}(0)=\operatorname{ker} i^{*}$ in $H^{1}\left(P, \mathbb{Z}_{2}\right)$, and has the same cardinality as ker $i^{*}$. We claim that this kernel is isomorphic to $H^{1}\left(M, \mathbb{Z}_{2}\right)$.

Consider the tail-end of the homotopy exact sequence of the fibration $S O(n) \rightarrow P \rightarrow M$, we have:

$$
\pi_{1}(S O(n)) \xrightarrow{i_{*}} \pi_{1}(P) \xrightarrow{\pi_{*}} \pi_{1}(M) \rightarrow 1
$$

so that taking $\operatorname{hom}_{\mathbb{Z}}\left(-, \mathbb{Z}_{2}\right)$ of this sequence, and noting $\operatorname{hom}_{\mathbb{Z}}\left(\pi_{1}(X), \mathbb{Z}_{2}\right)=H^{1}\left(X, \mathbb{Z}_{2}\right)$ by Hurewicz and Universal Coefficient Theorems, we have the exact sequence:

$$
0 \rightarrow H^{1}\left(M, \mathbb{Z}_{2}\right) \rightarrow H^{1}\left(P, \mathbb{Z}_{2}\right) \xrightarrow{i^{*}} H^{1}\left(S O(n), \mathbb{Z}_{2}\right)
$$

This shows that $\operatorname{ker} i^{*} \simeq H^{1}\left(M, \mathbb{Z}_{2}\right)$, and the proposition is proved.

Corollary 13.3.3. Every 2-connected Riemannian manifold is an orientable spin manifold.

Example 13.3.4 (Real projective spaces). The real projective space $\mathbb{R} \mathbb{P}(n)$ is spin iff $n \equiv 3 \bmod 4$. It is well known that $T(\mathbb{R} \mathbb{P}(n)) \oplus \epsilon^{1} \simeq\left(\gamma^{1 *}\right)^{n+1}$, where $\gamma^{1}$ is the tautological line bundle on $\mathbb{R} \mathbb{P}(n)$, and $\epsilon^{1}$ the trivial line bundle. Thus the total Steifel-Whitney class of $\mathbb{R} \mathbb{P}(n)$ is given by

$$
w(\mathbb{R P}(n))=(1+x)^{n+1}
$$

where $x \in H^{1}\left(\mathbb{R} \mathbb{P}(n), \mathbb{Z}_{2}\right)$ is the generator, and the first Stiefel-Whitney class of $\gamma^{1 *}$. So

$$
w_{2}(\mathbb{R} \mathbb{P}(n))=\frac{(n+1) n}{2} x^{2}
$$

Now $\mathbb{R} \mathbb{P}(n)$ is orientable iff $n=2 k+1$, and in this event $w_{2}(\mathbb{R} \mathbb{P}(n))=(k+1)(2 k+1) x^{2}$. This is zero iff $k$ is odd, i.e. iff $n=2(2 m+1)+1=4 m+3$.

Example 13.3.5 (Complex projective spaces). The complex projective space $\mathbb{C P}(n)$ is spin iff $n$ is odd. For, there is again the equivalence of complex vector bundles:

$$
T(\mathbb{C P}(n)) \oplus \epsilon_{\mathbb{C}}^{1} \simeq\left(\gamma^{1 *}\right)^{n+1}
$$

where $\gamma^{1}$ is the complex tautological line bundle on $\mathbb{C P}(n)$. Thus the total Chern class of $\mathbb{C P}(n)$ is given by

$$
c(\mathbb{C P}(n))=(1+x)^{n+1}
$$

where $x \in H^{2}(\mathbb{C P}(n), \mathbb{Z})$ is the generator, and the first Chern class of $\gamma^{1 *}$. This shows that the first Chern class

$$
c_{1}(\mathbb{C P}(n))=(n+1) x
$$

It is a fact that $w_{2}$ of a complex vector bundle considered as a real bundle is the mod 2 reduction of its first Chern class. Hence $w_{2}(\mathbb{C P}(n))=0$ iff $(n+1)$ is even, i.e. iff $n$ is odd.

Exercise 13.3.6. Using the identity $T\left(G_{k}\left(\mathbb{R}^{n}\right)\right) \simeq \operatorname{hom}\left(\gamma^{k}, \gamma^{k, \perp}\right)$, and arguments similar to the ones above, investigate which real grassmannians are spin. Likewise for complex grassmannians.

## 14. Representations

14.1. Clifford Modules. Let $V$ be a real inner product space with positive definite inner product $\langle-,-\rangle$. Let $C l(V)$ be the corresponding Clifford algebra.

Definition 14.1.1. Let $\mathbb{F}=\mathbb{R}$ or $\mathbb{C}$. We say that an $\mathbb{F}$-vector space $S$ is an $\mathbb{F}$-Clifford module if there is a unital $\mathbb{R}$ - algebra homomorphism:

$$
\rho: C l(V) \rightarrow \operatorname{hom}_{\mathbb{F}}(S, S)
$$

Example 14.1.2. Letting $S=C l(V)$, and letting $\rho(x) y=x . y$ (left Clifford multiplication by $x)$ or $\rho(x) y=$ $y . x^{*}$ (right Clifford multiplication by $x^{*}$ ) turns $C L(V)$ into an $\mathbb{R}$ - Clifford module. These are called the left (resp. right) regular representations.

A very important $\mathbb{R}$-Clifford module over $C l(V)$ is the exterior algebra $\Lambda^{*} V$. To describe it, we let $\left\{e_{i}\right\}_{i=1}^{n}$ be an orthonormal basis of $V$ with respect to $\langle-,-\rangle$. We also have the $\mathbb{R}$-linear Hodge-star operator

$$
*: \Lambda^{k}(V) \rightarrow \Lambda^{n-k}(V)
$$

which is defined on the basis elements of $\Lambda^{k}(V)$ by:

$$
*\left(e_{i_{1}} \wedge e_{i_{2}} \wedge \ldots \wedge e_{i_{k}}\right)=(-1)^{\sigma} e_{j_{1}} \wedge e_{j_{2}} \ldots \wedge e_{j_{n-k}}
$$

where $\left\{i_{1}, \ldots, i_{k}, j_{1}, . ., j_{n-k}\right\}=\{1,2, . ., n\}$, and $(-1)^{\sigma}$ is the sign of the permutation $\sigma=\left(i_{1} i_{2} \ldots, j_{1}, . ., j_{n-k}\right)$. We note that with this definition,

$$
\alpha \wedge * \beta=\langle\alpha, \beta\rangle \omega_{V}
$$

where $\omega_{n}=e_{1} \wedge . . \wedge e_{n} \in \Lambda^{n}(V)$ is the oriented volume element of $V$, and $\langle\alpha, \beta\rangle$ is the canonical inner product on $\Lambda^{k}(V)$ induced by $\langle-,-\rangle$ on $V$ (it is the inner product which makes $\left\{e_{i_{1}} \wedge e_{i_{2}} \ldots \wedge e_{i_{k}}\right\}$ an orthonormal basis for $\Lambda^{k}(V)$ ). It is readily checked that this inner product on $\Lambda^{k}(V)$ and the oriented volume element $\omega_{V}$ do not depend on the choice of orthonormal basis, and hence the $*$-operator is invariantly defined. It is easily checked that the square $* \circ *$ is scalar multiplication by $(-1)^{k(n-k)}$ on $\Lambda^{k}(V)$.

Definition 14.1.3 (Interior multiplication). For $v \in V$, and $\alpha \in \Lambda^{k}(V)$, define the element:

$$
v\lrcorner \alpha:=(-1)^{n k+n}(*(v \wedge * \alpha)) \in \Lambda^{k-1}(V)
$$

Lemma 14.1.4. Interior multiplication above satisfies the following:
(i): $\langle v \wedge \alpha, \beta\rangle=\langle\alpha, v\lrcorner \beta\rangle$ for $\alpha \in \Lambda^{k}(V), \beta \in \Lambda^{k+1}(V)$, and $v \in V$.
(ii): The composite:

$$
\Lambda^{k}(V) \xrightarrow{v_{\dashv}} \Lambda^{k-1}(V) \xrightarrow{v_{\not}} \Lambda^{k-2}(V)
$$

is zero.
(iii): For a vector $v \in V$, and $\alpha \in \Lambda^{k}(V)$, we have:

$$
v \wedge(v\lrcorner \alpha)+v\lrcorner(v \wedge \alpha)=\langle v, v\rangle \alpha
$$

Proof: By the discussion on the $*$-operator, we have

$$
\begin{aligned}
\langle v \wedge \alpha, \beta\rangle \omega_{V} & =(v \wedge \alpha) \wedge * \beta=(-1)^{k} \alpha \wedge v \wedge * \beta \\
& =(-1)^{k}(-1)^{(n-k) k} \alpha \wedge(* *)(v \wedge * \beta)=(-1)^{n k} \alpha \wedge *[*(v \wedge * \beta)] \\
& \left.\left.=(-1)^{n k}(-1)^{n(k+1)+n} \alpha \wedge *(v\lrcorner \beta\right)=\langle\alpha, v\lrcorner \beta\right\rangle \omega_{V}
\end{aligned}
$$

which proves (i). Thus interior multiplication $v\lrcorner(-)$ is the adjoint to $v \wedge(-)$ with respect to $\langle-,-\rangle$ on $\Lambda^{*}(V)$.
Since $v \wedge(v \wedge \alpha) \equiv 0$, the adjoint formula (i) implies (ii).
We note that for the basis vector $e_{1}$, and an element $\beta \in \Lambda^{*}$ such that $\beta$ does not involve $e_{1}$ anywhere, $\left.e_{1}\right\lrcorner \beta$ is orthogonal to all $\gamma$ not involving $e_{1}$, since $\left.\left\langle e_{1}\right\lrcorner \beta, \gamma\right\rangle=\left\langle\beta, e_{1} \wedge \gamma\right\rangle=0$. Further $\left.\left\langle e_{1}\right\lrcorner \beta, e_{1} \wedge \gamma\right\rangle=$ $\left\langle\beta, e_{1} \wedge e_{1} \wedge \gamma\right\rangle=0$. It follows that $\left.e_{1}\right\lrcorner \beta=0$, if $\beta$ does not involve $e_{1}$. On the other hand $\left.e_{1}\right\lrcorner\left(e_{1} \wedge \gamma\right)=\gamma$ for all $\gamma$ not involving $e_{1}$, as is easily checked again by taking inner products of both sides with various $\tau$, and using (i). Now, for a general $\alpha \in \Lambda^{k}(V)$, write $\alpha=\gamma+e_{1} \wedge \beta$, where $\beta$ and $\gamma$ do not involve $e_{1}$. Then $e_{1} \wedge \alpha=e_{1} \wedge \gamma$. Also $\left.e_{1}\right\lrcorner \alpha=\beta$. Hence

$$
\left.\left.\left.e_{1}\right\lrcorner\left(e_{1} \wedge \alpha\right)+e_{1} \wedge\left(e_{1}\right\lrcorner \alpha\right)=e_{1}\right\lrcorner\left(e_{1} \wedge \gamma\right)+e_{1} \wedge \beta=\gamma+e_{1} \wedge \beta=\alpha
$$

Since any unit vector $v$ can be completed to an orthonormal basis, the above formula is true for all unit vectors $v \in V$. For a general $v$, apply this formula to $\frac{v}{\|v\|}$, and (iii) follows.

Proposition 14.1.5. The exterior algebra $\Lambda^{*}(V)$ is an $\mathbb{R}$-Clifford module over $C l(V)$. The action is uniquely determined by the action of $v \in V \subset C l(V)$, and that in turn is given by:

$$
v . \alpha=v \wedge \alpha-v\lrcorner \alpha \quad \text { for } \alpha \in \Lambda^{*}(V)
$$

Finally, the action of $v \in V$ above is skew-symmetric with respect to the natural inner-product $\langle-,-\rangle$ on $\Lambda^{*}(V)$.

Proof: We define the action:

$$
v . \alpha:=v \wedge \alpha-v\lrcorner \alpha
$$

To extend this action to all of $C l(V)$, by the universal property of Clifford algebras, we need to check that $v^{2} \alpha=v . v . \alpha=-\langle v, v\rangle \alpha$ for all $\alpha \in \Lambda^{*}(V)$. However:

$$
\begin{aligned}
v . v . \alpha & =v \wedge(v \wedge \alpha-v\lrcorner \alpha)-v\lrcorner(v \wedge \alpha-v\lrcorner \alpha) \\
& =-v \wedge(v\lrcorner \alpha)-v\lrcorner(v \wedge \alpha)=-\langle v, v\rangle \alpha
\end{aligned}
$$

by using (ii) and (iii) of the previous Lemma 14.1.4.
Finally, by using (i) of the previous Lemma 14.1.4, we have

$$
\langle v . \alpha, \beta\rangle=\langle v \wedge \alpha, \beta\rangle-\langle v\lrcorner \alpha, \beta\rangle=\langle\alpha, v\lrcorner \beta\rangle-\langle\alpha, v \wedge \beta\rangle=-\langle\alpha, v . \beta\rangle
$$

which shows that the action of $v \in V$ is skew symmetric with respect to $\langle-,-\rangle$. Hence the proposition.

Exercise 14.1.6. Show that:

$$
v\lrcorner\left(w_{1} \wedge w_{2} \wedge \ldots \wedge w_{k}\right)=\sum_{i=1}^{k}(-1)^{i}\left\langle v, w_{i}\right\rangle w_{1} \wedge w_{2} \wedge \ldots \widehat{w_{i}} \ldots \wedge w_{k}
$$

where the hat denotes omission. (Simplest to just use a basis, or the adjointness formula $\langle v \wedge \alpha, \beta\rangle=\langle\alpha, v\lrcorner \beta\rangle$ in (i) of 14.1.4.)

Lemma 14.1.7. The representation of $C l(V)$ on $\Lambda^{*}(V)$ above has the following further properties:
(i): The map

$$
\begin{aligned}
\sigma: C l(V) & \rightarrow \Lambda^{*}(V) \\
x & \mapsto x .1
\end{aligned}
$$

is an $\mathbb{R}$-vector space isomorphism, called the symbol map. For a multiindex $I=\left(i_{1}<i_{2}<\ldots<i_{k}\right)$, we have $\sigma\left(e_{i_{1}} e_{i_{2}} . . e_{i_{k}}\right)=e_{i_{1}} \wedge e_{i_{2}} \wedge \ldots \wedge e_{i_{k}}$.
(ii): The inverse of $\sigma$ is the $\mathbb{R}$-linear map $c: \Lambda^{*}(V) \rightarrow C l(V)$ called the quantisation map. It obeys $c\left(e_{i_{1}} \wedge e_{i_{2}} \wedge \ldots \wedge e_{i_{k}}\right)=e_{I}$.
(iii): The representation above complexifies to a representation:

$$
\mathbb{C} l(V):=C l(V) \otimes_{\mathbb{R}} \mathbb{C} \rightarrow \Lambda_{\mathbb{C}}^{*}(V)=\Lambda^{*}(V) \otimes_{\mathbb{R}} \mathbb{C}=\Lambda^{*}\left(V_{\mathbb{C}}\right)
$$

and the symbol and quantisation maps extend to the complexifications.
(iv): [Action of the volume element] Give $\mathbb{R}^{n}$ its usual euclidean inner product and denote the corresponding Clifford algebra $C l\left(\mathbb{R}^{n}\right)$ as $C l_{n}$. Let $\omega_{n}=e_{1} \ldots e_{n}$ be the volume element of $C l_{n}$. Then:
(a): $\omega_{n} v+v \omega_{n}=0$ for $n$ even and $\omega_{n} v=v \omega_{n}$ for $n$ odd and all $v \in \mathbb{R}^{n}$. Hence, for $n$ odd, $\omega_{n}$ commutes with everything and is a central element in $C l_{n}$. For $n$ even, $\omega_{n}$ commutes with $C l_{n}^{0}$ and anticommutes with $C l_{n}^{1}$.
(b): $\omega_{n}^{2}=(-1)^{p}$ where $p=\left[\frac{n+1}{2}\right]$, the integral part of $\frac{n+1}{2}$. Hence $\omega_{n}^{2}=-1$ for $n \equiv 1,2 \bmod 4$ and $\omega^{2}=1$ for $n \equiv 0,3 \bmod 4$.
(c): The action of $\omega_{n}$ on $\Lambda^{*}(V)$ is related to the Hodge-star operator by:

$$
\omega \cdot \phi=(-1)^{n k+\frac{k(k-1)}{2}} * \phi \text { for } \phi \in \Lambda^{k}(V)
$$

(v): [Chirality element] In the complexification $\mathbb{C} l_{n}=C L_{n} \otimes_{\mathbb{R}} \mathbb{C}$, define the complex volume element or chirality element:

$$
\tau_{n}:=(\sqrt{-1})^{p} \omega_{n} \quad \text { where } p=\left[\frac{n+1}{2}\right]
$$

the box brackets denoting the greatest integer part. By (b) of (iv) above

$$
\tau_{n}^{2}=1 \text { for all } n
$$

Since $\omega_{n}$ is central for all $n \equiv 1 \bmod 2$, we have $\tau_{n}$ is central for all $n \equiv 1 \bmod 2 . \tau_{n}$ is related to the Hodge-star operator by:

$$
\tau_{n} \cdot \phi=i^{p+k(2 n+k-1)} * \phi \quad \text { for } \phi \in \Lambda_{\mathbb{C}}^{k}\left(\mathbb{R}^{n}\right)
$$

In particular, if $n=4 m$ and $k=2 m$, we have $\tau_{n} \phi=* \phi$. (Chirality coincides with Hodge-star on middle dimension for $n=4 m$ ).

Proof: Note that, denoting the $C l(V)$ action with a dot, we have:

$$
\left.e_{i} \cdot 1=e_{i} \wedge 1-e_{i}\right\lrcorner 1=e_{i} \text { for } 1 \in \Lambda^{0}(V)
$$

and so for $I=\left(i_{1}<i_{2}<\ldots<i_{k}\right)$, it follows that:

$$
\begin{aligned}
e_{I} \cdot 1 & \left.=e_{i_{1}} e_{i_{2}} . . e_{i_{k}} \cdot 1=\left(e_{i_{1}} e_{i_{2}} . . e_{i_{k-1}}\right) \cdot e_{i_{k}}=\left(e_{i_{1}} e_{i_{2}} . . e_{i_{k-2}}\right) \cdot\left(e_{i_{k-1}} \wedge e_{i_{k}}-e_{i_{k-1}}\right\lrcorner e_{i_{k}}\right) \\
& =\left(e_{i_{1}} e_{i_{2}} . . e_{i_{k-2}}\right) \cdot\left(e_{i_{k-1}} \wedge e_{i_{k}}\right)=\ldots .=e_{i_{1}} \wedge e_{i_{2}} \wedge \ldots \wedge \wedge e_{i_{k}}
\end{aligned}
$$

because $\left.e_{l}\right\lrcorner\left(e_{j_{1}} \wedge e_{j_{2}} \wedge . . e_{j_{m}}\right)=0$ if $l \neq j_{i}$ for all $i$. This proves (i).
(ii) follows immediately from (i), since $c=\sigma^{-1} . c$ is called the quantisation map because all supercommutators are 0 in $\Lambda^{*}(V)$, but not in $C l(V)$, and $c$ puts a "non- supercommuting" algebra structure on the supercommutative algebra $\Lambda^{*}(V)$.
(iii) is obvious from definitions.

For (iv), note that for any $n, e_{i} e_{j}+e_{j} e_{i}=2 \delta_{i j}$ implies that

$$
e_{i} \omega_{n}=(-1)^{n-1} \omega_{n} e_{i}
$$

so we have (a) of (iv). We also have:

$$
\omega_{n}^{2}=\omega_{n-1} e_{n} \omega_{n-1} e_{n}=(-1)^{n-1} \omega_{n-1}^{2} e_{n}^{2}=(-1)^{n} \omega_{n-1}^{2}
$$

So that $\omega_{4 k+4}^{2}=\omega_{4 k+3}^{2}=-\omega_{4 k+2}^{2}=-\omega_{4 k+1}^{2}=\omega_{4 k}^{2}=\ldots . \omega_{0}^{2}=1$, and (b) follows.
To see (c) of (iv), note that if $e_{I}=e_{i_{1}} . . e_{i_{k}} \in C l_{n}$, and $J$ is any multi-index with $J=\{1,2, . ., n\} \backslash I$, then $e_{I} e_{J}=(-1)^{\sigma} \omega_{n}$ where $\sigma$ is the permutation

$$
\sigma=\left(i_{1}, i_{2}, \ldots, j_{1}, . ., j_{n-k}\right)
$$

Also note that $e_{I} e_{I}=(-1)^{\frac{k(k+1)}{2}}$. Now by (a)

$$
\begin{aligned}
\omega_{n} e_{I} & =(-1)^{k(n-1)} e_{I} \cdot \omega_{n}=(-1)^{k(n-1)+\sigma} e_{I} e_{I} e_{J} \\
& =(-1)^{k(n-1)+\frac{k(k+1)}{2}}\left[(-1)^{\sigma} e_{J}\right]=(-1)^{n k+\frac{k(k-1)}{2}}\left[(-1)^{\sigma} e_{J}\right]
\end{aligned}
$$

Now apply both sides to $1 \in \Lambda^{0}$ to get (c).
(v) follows immediately from the definition of $\tau_{n}$ and (iv). When $n=4 m, p=\left[\frac{n+1}{2}\right]=2 m$, and for $k=2 m$ the exponent

$$
p+k(2 n+k-1)=2 m+2 m(8 m+2 m-1)=16 m^{2}+4 m \equiv 0 \quad \bmod 4
$$

so that $i^{p+k(2 n+k-1)}=1$, and $\tau_{4 m} \phi=* \phi$ for $\phi \in \Lambda_{\mathbb{C}}^{2 m}\left(\mathbb{R}^{n}\right)$. This proves the lemma.

Corollary 14.1.8. If $W$ is a $\mathbb{R}$-Clifford module over $C l_{n}$, and $n \equiv 0,3 \bmod 4$, then $W=W^{+} \oplus W^{-}$as an $\mathbb{R}$-vector space, where $W^{ \pm}=\left(1 \pm \omega_{n}\right) W$ is the $( \pm 1)$-eigenspace of the volume element action $\omega_{n}$. ( ). If $n \equiv 3$ $\bmod 4$, the centrality of $\omega_{n}$ ensures that $W^{ \pm}$are both $\mathbb{R}$-Clifford submodules of $W$. If $n \equiv 0 \bmod 4$, then $W^{ \pm}$ are $\mathbb{R}$-modules over $C l_{n}^{0}$.

Analogously, since $\tau_{n}^{2}=1$, every $\mathbb{C}$-Clifford module $W$ over $C l_{n}$ splits into ( $\pm 1$ )-eigenspaces $W^{ \pm}$of the chirality element $\tau_{n}$ for all $n$, both being $\mathbb{C}$-subspaces. Again, if $n \equiv 3 \bmod 4, \tau_{n}$ is central, and $W^{ \pm}$are $\mathbb{C}$-Clifford submodules. If $n \equiv 0 \bmod 4$, the subspaces $W^{ \pm}$are modules over $C l_{n}^{0}$.

Proof: Obvious from (b) of (iv) and (v) of the Lemma 14.1.7 above. When $n \equiv 0 \bmod 4$, we note that $\omega_{n}$ (resp. $\tau_{n}$ ) anticommutes with all $v \in V$, and hence commutes with $C l^{0}$. Thus $W^{ \pm}$is a $\mathbb{R}$ (resp. $\mathbb{C}$ ) module over $C l_{n}^{0}$.
Definition 14.1.9. Let $\mathbb{F}=\mathbb{R}$ or $\mathbb{C}$. We will say that an $\mathbb{F}$-Clifford module over $C l(V)$ is irreducible if the only $\mathbb{F}$ - Clifford submodules of $S$ are $S$ and $\{0\}$.

Remark 14.1.10. If $n \equiv 3 \bmod 4$, then for an irreducible $\mathbb{R}$ (resp. $\mathbb{C}$ )-Clifford module $W$ over $C l_{n}$, the volume element $\omega_{n}$ (resp. chirality $\tau_{n}$ ) either acts as +1 (viz. $W=W^{+}$) or as -1 (viz. $W=W^{-}$), and $W^{+}$ and $W^{-}$(if both exist) are distinct irreducible modules. This is obvious because of the Corollary 14.1.8 which asserts that $W^{ \pm}$are $\mathbb{F}$-Clifford submodules of $W$ when $n \equiv 3 \bmod 4$, by the centrality of $\omega_{n}$ (resp. $\tau_{n}$ ). Also, since $C l_{n}$ module equivalence will preserve the sign of $\omega_{n} .()$ (resp. $\left.\tau_{n} .()\right)$, it follows that modules on which these operators act as +1 are not isomorphic to those on which they act by -1 .

Lemma 14.1.11 (Complete reducibility of Clifford modules). Every $\mathbb{F}$-Clifford module is a direct sum of irreducible $\mathbb{F}$-Clifford submodules.

Proof: We define the Clifford group in $C l(V)$ to be the group:

$$
\Gamma_{n}:=\left\{ \pm e_{I}: I=\left(i_{1}<i_{2}<\ldots<i_{k}\right) ; 0 \leq k \leq n\right\}
$$

which is of order $2^{n+1}$. For example, $\Gamma_{2}$ is the Hamilton group $\{ \pm 1, \pm i, \pm j, \pm k\}$. Denote the element $-1 \in \Gamma_{n}$ by $\nu$. If we let $\mathbb{R}\left[\Gamma_{n}\right]$ denote the real group algebra over $\Gamma_{n}$, we have a surjective $\mathbb{R}$ - algebra homomorphism:

$$
\begin{aligned}
\rho: \mathbb{R}\left[\Gamma_{n}\right] & \rightarrow C l(V) \\
e_{I} & \mapsto e_{I} \\
\nu & \mapsto-1
\end{aligned}
$$

Thus there is a $1-1$ correspondence between $\mathbb{F}$-modules over $C l(V)$ and $\mathbb{F}$-modules over $\mathbb{R}\left[\Gamma_{n}\right]$ on which the element $\nu$ acts as -1 . So let $S$ be a $\mathbb{F}$-module over $C l(V)$. Then, via $\rho, S$ is a $\mathbb{F}$-module over the algebra $\mathbb{R}\left[\Gamma_{n}\right]$, i.e. a $\mathbb{F}$-module over the Clifford group $\Gamma_{n}$. By averaging over the finite group $\Gamma_{n}$, there alway exists a $\Gamma_{n}$-invariant positive definite real (if $\mathbb{F}=\mathbb{R}$ ), resp. complex sesquilinear (if $\mathbb{F}=\mathbb{C}$ ) inner product $\langle-,-\rangle$. on the $\mathbb{F}$-module $S$. Thus every $\Gamma_{n} \mathbb{F}$-submodule will have a $\Gamma_{n}$-invariant orthogonal complement with respect to $\langle-,-\rangle$. It follows that $S$ decomposes into the orthogonal direct sum of finitely many irreducible $\Gamma_{n} \mathbb{F}$ submodules $S_{i}$. Thus $S_{i}$ are irreducible $\mathbb{R}\left[\Gamma_{n}\right] \mathbb{F}$-submodules. Since $\nu$ is acting as -1 on $S$, it is acting as -1 on each $S_{i}$, so each $S_{i}$ is a $C l(V) \mathbb{F}$-submodule. It is clearly irreducible over $C l(V)$ since it is irreducible over $\mathbb{R}\left[\Gamma_{n}\right]$. The lemma follows.

So it remains to identify what the irreducible $C l(V) \mathbb{F}$ - modules are. This will be addressed in the following proposition.

Proposition 14.1.12. For $n \equiv 0,1,2 \bmod 4$, there is exactly one irreducible $\mathbb{R}$-module over $C l_{n}$. For $n \equiv 3$ $\bmod 4$, there are two distinct irreducible $\mathbb{R}$-modules over $C l_{n}$. They are distinguished by the fact that on one the volume element $\omega_{n}$ acts as $(+1)$, and on the other as $(-1)$. The dimensions of these modules are readable from the following list:

$$
\begin{array}{ccccccccc}
n: & 8 k+1 & 8 k+2 & 8 k+3 & 8 k+4 & 8 k+5 & 8 k+6 & 8 k+7 & 8 k+8 \\
d_{n}: & 2^{4 k+1} & 2^{4 k+2} & 2^{4 k+2} & 2^{4 k+3} & 2^{4 k+3} & 2^{4 k+3} & 2^{4 k+3} & 2^{4 k+4}
\end{array}
$$

For $n \equiv 0 \bmod 2$, there is exactly one irreducible $\mathbb{C}$-module over $C l_{n}$, of $\mathbb{C}$-dimension $2^{n / 2}$. For $n \equiv 1$ $\bmod 2$, there are exactly two irreducible $\mathbb{C}$-modules over $C l_{n}$, each of $\mathbb{C}$-dimension $2^{\frac{n-1}{2}}$. They are distinguished by the fact that on one the chirality element $\tau_{n}$ acts as $(+1)$ and on the other as $(-1)$.

Proof: We recall the list:

$$
\begin{array}{ccccccccc}
n: & 1 & 2 & 3 & 4 & 5 & 6 & 7 & 8 \\
C l_{n}: & \mathbb{C} & \mathbb{H} & \mathbb{H} \oplus \mathbb{H} & \mathbb{H}(2) & \mathbb{C}(4) & \mathbb{R}(8) & \mathbb{R}(8) \oplus \mathbb{R}(8) & \mathbb{R}(16)
\end{array}
$$

and the fact that $C l_{n+8} \simeq C l_{n} \otimes_{\mathbb{R}} \mathbb{R}(16)$ from the Remark 13.1.5. Also note that by (i) of Exercise 13.1.4, we have $C l_{3}=\mathbb{H} \oplus \omega \mathbb{H}$, with $\omega^{2}=\left(e_{1} e_{2} e_{3}\right)^{2}=1$ and $\omega$ a central element. This algebra may be rewritten as $(1+\omega) \mathbb{H} \oplus(1-\omega) \mathbb{H}$, where $(1+\omega)(1-\omega)=0$, so that $C l_{3}=\mathbb{H} \oplus \mathbb{H}$. Note that $\omega(1 \pm \omega)=(1 \pm \omega)$, so that the two summands in $C l_{3}$ are distinguished by the sign of the action of $\omega$.

The corresponding fact is also true of $C l_{7}$, though we haven't computed it thus far. However, assuming that $C l_{6}=\mathbb{R}(8)$, it is easy to check that $C l_{7}^{0} \simeq C l_{6}$, by taking $e_{I} e_{7} \mapsto \pm e_{I}$ and $e_{J} \mapsto e_{J}$ for all subsets $I, J \subset\{1,2, . ., 6\}$. Now it is easy to check that $C l_{7}^{1}=\omega C l_{7}^{0}$, and $C l_{7}=C l_{7}^{0} \oplus \omega C l_{7}^{0}=(1+\omega) \mathbb{R}(8) \oplus(1-\omega) \mathbb{R}(8) \simeq$ $\mathbb{R}(8) \oplus \mathbb{R}(8)$. So again, the two summands of $C l_{7}$ are distinguished by the sign of Clifford multiplicaton by $\omega$.

From this list it follows that $C l_{n}$ is a matrix algebra $\mathbb{K}(k)$ over $\mathbb{K}=\mathbb{R}$ or $\mathbb{C}$ or $\mathbb{H}$ for $n=1,2,4,5,6,8$ and a sum of two copies of the same matrix algebra $\mathbb{K}(k)$ for $n=3,7$. Also, since $\mathbb{K}(k) \otimes_{\mathbb{R}} \mathbb{R}(m)=\mathbb{K}(m k)$, it follows by the 8 -periodicity above that $C l_{n}$ is a matrix algebra $\mathbb{K}(k)$ for $n \equiv 0,1,2,4,5,6 \bmod 8$, i.e. $n \not \equiv 3 \bmod 4$, and a direct sum of two identical matrix algebras $\mathbb{K}(k)$ for $n \equiv 3,7 \bmod 8$, i.e. $n \equiv 3 \bmod 4$.

It is well known that the $\mathbb{K}$-matrix algebra $\mathbb{K}(k)$ is simple, and has exactly one irreducible $\mathbb{R}$-module over it, namely $\mathbb{K}^{k}$, with the obvious left action by matrix multiplication. The direct sum of two copies of $\mathbb{K}(k)$ has two distinct irreducible modules over it, viz. $\mathbb{K}^{k}$ with one action from the first summand, and the other action from the second summand. Thus by the foregoing, the two irreducible modules for $n \equiv 3$ mod 4 are distinguished by the sign of the action of $\omega_{n}=e_{1} \ldots e_{n}$. Letting $d_{n}$ denote the $\mathbb{R}$-dimension of these irreducible modules, we have the following table:

$$
\begin{array}{ccccccccc}
n: & 1 & 2 & 3 & 4 & 5 & 6 & 7 & 8 \\
d_{n}: & 2 & 4 & 4 & 8 & 8 & 8 & 8 & 16
\end{array}
$$

It follows that $C l_{n+8}=C l_{n} \otimes \mathbb{R}(16)$ will have exactly one irreducible $\mathbb{R}$-module for $n \not \equiv 3 \bmod 4$ and two distinct ones for $n \equiv 3 \bmod 4$. The dimensions of these modules are read off from the above table, and the inductive formula $d_{n+8}=16 d_{n}$ arising out of periodicity.

Denote an irreducible $\mathbb{R}$-Clifford module over $C l_{n}$ as $W_{n}$. Then by the remarks above and the list at the top we have:

$$
\begin{array}{ccccccccc}
n & 1 & 2 & 3 & 4 & 5 & 6 & 7 & 8 \\
W_{n} & \mathbb{C} & \mathbb{H} & \mathbb{H}_{ \pm} & \mathbb{H}^{2} & \mathbb{C}^{4} & \mathbb{R}^{8} & \mathbb{R}_{ \pm}^{8} & \mathbb{R}^{16}
\end{array}
$$

where the subscript $\pm$ on $W_{3}$ and $W_{7}$ signifies two distinct irreducible modules, both isomorphic as vector spaces to the entry in that slot. This implies by the periodicity $W_{n+8}=W_{n} \otimes_{\mathbb{R}} \mathbb{R}^{16}$ that we have the following list of irreducible $\mathbb{R}$-Clifford modules $W_{n}$ over $C l_{n}$ whose $\mathbb{R}$-dimension is $d_{n}$ :

$$
\begin{array}{ccccccccc}
n: & 8 k+1 & 8 k+2 & 8 k+3 & 8 k+4 & 8 k+5 & 8 k+6 & 8 k+7 & 8 k+8 \\
W_{n}: & \mathbb{C}^{2^{4 k}} & \mathbb{H}^{2^{4 k}} & \mathbb{H}_{ \pm}^{2^{4 k}} & \mathbb{H}^{2^{4 k+1}} & \mathbb{C}^{2^{4 k+2}} & \mathbb{R}^{2^{4 k+3}} & \mathbb{R}_{ \pm}^{2^{4 k+3}} & \mathbb{R}^{2^{4 k+4}} \\
d_{n}: & 2^{4 k+1} & 2^{4 k+2} & 2^{4 k+2} & 2^{4 k+3} & 2^{4 k+3} & 2^{4 k+3} & 2^{4 k+3} & 2^{4 k+4}
\end{array}
$$

The complex modules are much simpler to describe. Noting that an $\mathbb{R}$ - algebra homomorphism:

$$
\rho: C l(V) \rightarrow \operatorname{hom}_{\mathbb{C}}(W, W)
$$

automatically extends to the complexification $\mathbb{C l}(V):=C l(V) \otimes_{\mathbb{R}} \mathbb{C}$, we see that a $\mathbb{C}$-Clifford module becomes a $\mathbb{C l}(V)$ module. Noting that $\mathbb{C} \otimes_{\mathbb{R}} \mathbb{C}=\mathbb{C} \oplus \mathbb{C}$, and $\mathbb{H} \otimes_{\mathbb{R}} \mathbb{C}=\mathbb{C}(2)$ (see (ii) of Exercise 13.1.4), we get the following list of complex Clifford algebras from the real list above:

| $n:$ | 1 | 2 | 3 | 4 | 5 | 6 | 7 | 8 |
| :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: | :---: |
| $\mathbb{C} l_{n}:$ | $\mathbb{C} \oplus \mathbb{C}$ | $\mathbb{C}(2)$ | $\mathbb{C}(2) \oplus \mathbb{C}(2)$ | $\mathbb{C}(4)$ | $\mathbb{C}(4) \oplus \mathbb{C}(4)$ | $\mathbb{C}(8)$ | $\mathbb{C}(8) \oplus \mathbb{C}(8)$ | $\mathbb{C}(16)$ |

which means we have:

$$
\begin{aligned}
\mathbb{C} l_{n} & =\mathbb{C}\left(2^{k}\right) \oplus \mathbb{C}\left(2^{k}\right) \text { for } 1 \leq n=2 k+1 \leq 8 \\
& =\mathbb{C}\left(2^{k}\right) \text { for } 1 \leq n=2 k \leq 8
\end{aligned}
$$

Note that for $n=3,7$, the two summands in $\mathbb{C} l_{n}$ are distinguished by the sign of multiplication by the central volume element $\omega_{n}$. Note also that the chirality elements (see definition in (v) of Lemma 14.1.7) are given by $\tau_{3}=-\omega_{3}$ and $\tau_{7}=\omega_{7}$. Hence, for $n=3,7$, the two summands in $\mathbb{C} l_{n}$ are distinguished by the sign of the action of multiplication by the chirality $\tau_{n}$.

Since $\mathbb{R}(16) \otimes_{\mathbb{R}} \mathbb{C}=\mathbb{C}(16)$, it follows that

$$
\begin{aligned}
\mathbb{C} l_{n+8} & =C l_{n+8} \otimes_{\mathbb{R}} \mathbb{C}=C l_{n} \otimes_{\mathbb{R}} \mathbb{R}(16) \otimes_{\mathbb{R}} \mathbb{C}=C l_{n} \otimes_{\mathbb{R}} \mathbb{C}(16)=C l_{n} \otimes_{\mathbb{R}}\left(\mathbb{C} \otimes_{\mathbb{C}} \mathbb{C}(16)\right) \\
& =\left(C l_{n} \otimes_{\mathbb{R}} \mathbb{C}\right) \otimes_{\mathbb{C}} \mathbb{C}(16)=\mathbb{C} l_{n} \otimes_{\mathbb{C}} \mathbb{C}(16)
\end{aligned}
$$

Thus there is exactly one irreducible $\mathbb{C}$-Clifford module over $C l_{n}$ when $n$ is even, and two inequivalent irreducible $\mathbb{C}$ - Clifford modules over $C l_{n}$ when $n$ is odd. Again, when $n$ is odd, the two inequivalent modules are distinguished by the sign of the action of the chirality $\tau_{n}$.

Combining the two facts above, we obtain:

$$
\begin{aligned}
\mathbb{C} l_{n} & =\mathbb{C}\left(2^{\frac{n-1}{2}}\right) \oplus \mathbb{C}\left(2^{\frac{n-1}{2}}\right) \text { for } n \equiv 1 \bmod 2 \\
& =\mathbb{C}\left(2^{\frac{n}{2}}\right) \text { for } n \equiv 0 \bmod 2
\end{aligned}
$$

Since the matrix algebra $\mathbb{C}(k)$ is simple, there is exactly one irreducible $\mathbb{C}$-module over it, viz. $\mathbb{C}^{k}$ with the obvious action by matrix multiplication. Similarly, over the direct sum algebra $\mathbb{C}(k) \oplus \mathbb{C}(k)$, there are exactly two irreducible ones (each isomorphic as a $\mathbb{C}$-vector space to $\mathbb{C}^{k}$ ), coming from the action of the two distinct summands. As noted above, the two summands are distinguished by the sign of the chirality oeprator $\tau_{n}$. This proves the proposition.

Remark 14.1.13. Note that from the proposition above, since $d_{n}<2^{n}$ for all $n \geq 3$, it follows that the $\mathbb{R}$-Clifford module $\Lambda^{*}\left(\mathbb{R}^{n}\right)$ of dimension $2^{n}$ described in Proposition 14.1.5 is irreducible iff $n=1$ or $n=2$. For the same reason the left and right regular representations of $C l_{n}$ on itself is irreducible iff $n=1$ or 2 .

To further analyse the real and complex representations of $C l_{n}$, we introduce the notion of a graded Clifford module. That is,

Definition 14.1.14. Say that $W$ is a $\mathbb{Z}_{2}$-graded $\mathbb{F}$-Clifford module over $C l(V)$ (or a $C l(V) \mathbb{F}$-supermodule) if $W=W^{0} \oplus W^{1}$, with $W^{i}$ as $\mathbb{F}$-vector subspaces satisfying:

$$
C l^{i}(V) W^{j} \subset W^{k} \text { where } k=i+j \bmod 2
$$

A $\mathbb{C}$-supermodule over $C l(V)$ can be naturally regarded as a $\mathbb{C l}(V):=C l(V) \otimes_{\mathbb{R}} \mathbb{C} \mathbb{C}$-supermodule.

Example 14.1.15. If we regard $C l(V)=C l^{0}(V) \oplus C l^{1}(V)$ as a module over itself via left regular representation (or right regular multiplication), it becomes a $C l(V)$ supermodule. Analogously, the decomposition $\mathbb{C l}(V)=$ $\mathbb{C l}(V)^{0} \oplus \mathbb{C l}(V)^{1}$ makes $\mathbb{C l}(V)$ a $\mathbb{C}$-supermodule over $\mathbb{C} l(V)$ via left or right regular representation.

Example 14.1.16 (The exterior algebra again). We noted in Proposition 14.1.5 that the exterior algebra $\Lambda^{*}(V)$ is a $C l(V)$ module. Hence the summands $\Lambda^{e}(V)=\oplus_{i=0}^{n} \Lambda^{2 i}(V)=C l^{0} V .1$ and $\Lambda^{o}(V)=\oplus_{i=0}^{n} \Lambda^{2 i+1}(V)=$ $C l^{1}(V) .1$ gives $\Lambda^{*}(V)$ the structure of a $C l(V)$-supermodule, by considering the foregoing example.

In entirely analogous fashion, $\Lambda_{\mathbb{C}}^{*}(V):=\Lambda^{*}(V) \otimes_{\mathbb{R}} \mathbb{C}$ becomes a $\mathbb{C} l(V)$ supermodule via the decomposition $\Lambda_{\mathbb{C}}^{*}(V)=\Lambda_{\mathbb{C}}^{e}(V) \oplus \Lambda_{\mathbb{C}}^{o}(V)$ into even and odd degree forms.

Example 14.1.17. Let $n \equiv 0$ or $n \equiv 3 \bmod 4$. By the Corollary 14.1 .8 the left $\mathbb{R}$-Clifford module $C l_{n}$ decomposes into the $(+1)$ and $(-1)$ eigenspaces of $\omega_{n}$. We denote this decomposition as:

$$
C l_{n}=C l_{n}^{+} \oplus C l_{n}^{-} \quad \text { where } \quad C l_{n}^{ \pm}:=\left(1 \pm \omega_{n}\right) C l_{n} \quad \text { and } n=4 m, 4 m+3
$$

This is a different $\mathbb{Z}_{2}$ grading from the earlier $C l^{0} \oplus C l^{1}$ grading. Indeed, for $n=4 m+3$, the element $\left(1+\omega_{4 m+3}\right)$ is in $C l^{+}$, but not in $C l^{0}$ or $C l^{1}$, since $1 \in C l^{0}$ and $\omega_{4 m+3} \in C l^{1}$. Similarly, for $n=4 m$, the element $\left(1-\omega_{4 m}\right) \in C l^{-}$but not in $C l^{1}$.

When $n=4 m$, we have $e_{i} \omega_{4 m}=-\omega_{4 m} e_{i}$ for all $i$, and so $a \omega_{4 m}=-\omega_{4 m} a$ for all $a \in C l^{1}$, and $a \omega_{4 m}=\omega_{4 m} a$ for all $a \in C l^{0}$. Thus, for $n=4 m$, we have that $C l_{4 m}=C l_{4 m}^{+} \oplus C l_{4 m}^{-}$is a $C l_{4 m} \mathbb{R}$-supermodule. (Unfortunately,
the corresponding fact is untrue for $n=4 m+3$ since $\omega_{4 m+3}$ is central, so $C l^{1}$ preserves both $C l^{+}$and $C l^{-}$ instead of interchanging them). However, the above grading on $C l_{4 m}$ has some bearing on $C l\left(\mathbb{R}^{4 m+1}\right)$, as we shall see soon.

Example 14.1.18 (Signature grading). Let $V=\mathbb{R}^{4 m}$, and consider the $\mathbb{R}$-Clifford module $\Lambda^{*}\left(\mathbb{R}^{4 m}\right)$ over $C l_{4 m}$. By the previous example, $C l_{4 m}=C l_{4 m}^{+} \oplus C l_{4 m}^{-}$becomes a $C l_{4 m}$ supermodule via action of left Clifford multiplication, the decomposition being determined by the sign of multiplication by $\omega_{4 m}$. Since $\Lambda^{*}\left(\mathbb{R}^{4 m}\right)=$ $C l_{4 m} .1$, it follows that:

$$
\Lambda^{*}\left(\mathbb{R}^{4 m}\right)=\Lambda^{+}\left(\mathbb{R}^{4 m}\right) \oplus \Lambda^{-}\left(\mathbb{R}^{4 m}\right)
$$

where $\Lambda^{ \pm}\left(\mathbb{R}^{4 m}\right):=C l_{4 m}^{ \pm}$. . This makes $\Lambda^{*}\left(\mathbb{R}^{4 m}\right)$ a $C l_{4 m}$ supermodule, by the previous example. Similar considerations apply to $\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{4 m}\right)$ which becomes a $\mathbb{C} l_{4 m} \mathbb{C}$-supermodule via the grading:

$$
\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{4 m}\right)=\Lambda_{\mathbb{C}}^{+}\left(\mathbb{R}^{4 m}\right) \oplus \Lambda_{\mathbb{C}}^{-}\left(\mathbb{R}^{4 m}\right)
$$

This last grading is called the signature grading because the Clifford action of $\tau_{4 m}$ coincides with the Hodge-star operator in the middle dimension $\Lambda_{\mathbb{C}}^{2 m}$, by the last statement in (v) of Lemma 14.1.7.

It is helpful to have an explicit model for the complex Clifford modules. This is the content of the next proposition.

Proposition 14.1.19 (The irreducible complex $\mathbb{C} l_{2 m}$ modules). Let $V=\mathbb{R}^{2 m}$ with the usual euclidean inner product $\langle-,-\rangle$. Extend this inner product by complex linearity to $\langle-,-\rangle$ on the complexification $V_{\mathbb{C}}=V \otimes_{\mathbb{R}}$ $\mathbb{C}=\mathbb{C}^{2 m}$ (i.e. this inner product is not positive definite on $V_{\mathbb{C}}$, being $\mathbb{C}$-linear in both variables). Let $P$ be the complex subspace of $V_{\mathbb{C}}$ defined by:

$$
P:=\mathbb{C}-\operatorname{span}\left\{e_{2 j-1}-i e_{2 j}: 1 \leq j \leq m\right\}
$$

Then set $S=\Lambda^{*}(P)$, a $\mathbb{C}$-vector space with $\operatorname{dim}_{\mathbb{C}} S=2^{m}$. Then $S=S^{+} \oplus S^{-}$is a $\mathbb{C} l_{2 m} \mathbb{C}$-supermodule which is irreducible. $S^{ \pm}$are the $\pm 1$-eigenspaces with respect to the (non-central) chirality element $\tau_{2 m}$, and turn out to be $S^{+}=\Lambda^{e v} P$ and $S^{-}=\Lambda^{o} P$. Finally:

$$
\mathbb{C} l_{2 m}=C l_{2 m} \otimes_{\mathbb{R}} \mathbb{C}=\operatorname{hom}_{\mathbb{C}}(S, S)
$$

Proof: First note that $V_{\mathbb{C}}$, being a complexification, comes with the natural complex conjugation $v \otimes \lambda \mapsto v \otimes \bar{\lambda}$ for $v \in V$. Also, $P$ is a real-form of $V_{\mathbb{C}}$, i.e.

$$
V_{\mathbb{C}}=P \oplus \bar{P}
$$

where $\bar{P}$ denotes the complex conjugation of $P$ inside $V_{\mathbb{C}}$. Now we claim that the subspace $P$ is isotropic, i.e. $\langle v, w\rangle \equiv 0$ for all $v, w \in P$. For,

$$
\left\langle e_{2 j-1}-i e_{2 j}, e_{2 k-1}-i e_{2 k}\right\rangle=\left\langle e_{2 j-1}, e_{2 k-1}\right\rangle+i^{2}\left\langle e_{2 j}, e_{2 k}\right\rangle=\delta_{j k}-\delta_{j k} \equiv 0 \text { for all } 1 \leq j \leq m
$$

Now define a basis of $P$ by:

$$
f_{j}:=\frac{1}{\sqrt{2}}\left(e_{2 j-1}-i e_{2 j}\right) \quad 1 \leq j \leq m
$$

so that

$$
\left\langle f_{j}, \bar{f}_{k}\right\rangle=\frac{1}{2}\left(\left\langle e_{2 j-1}, e_{2 k-1}\right\rangle+\left\langle e_{2 j}, e_{2 k}\right\rangle\right)=\frac{1}{2}\left(\delta_{j k}+\delta_{j k}\right)=\delta_{j k}
$$

which shows that $\left\{\bar{f}_{j}\right\}$ is a basis of $\bar{P}$ which is dual to the basis $\left\{f_{j}\right\}$ of $\bar{P}$. This identifies $\bar{P}$ with the complex dual $P^{*}$.

Define the action of $P$ on $\Lambda^{*} P$ by

$$
v \circ \phi:=\sqrt{2} v \wedge \phi \text { for } v \in P, \phi \in \Lambda^{*} P
$$

Note that $(v \circ(v \circ \phi))=2(v \wedge v \wedge \phi) \equiv 0=\langle v, v\rangle \phi$ for all $v \in P$, and $\phi \in \Lambda^{*} P$.
Define the action of $P^{*}$ on $S=\Lambda^{*} P$ by duality:

$$
\langle\bar{v} \circ \phi, \psi\rangle:=-\langle\phi, v \circ \psi\rangle=-\sqrt{2}\langle\phi, v \wedge \psi\rangle=-\sqrt{2}\langle v\lrcorner \psi, \phi\rangle
$$

Hence setting $\bar{v} \circ \phi=-\sqrt{2} v\lrcorner \phi$ defines an action of $P^{*}=\bar{P}$ on $\Lambda^{*} P$, which also satisfies $\bar{v} \circ(\bar{v} \circ \phi) \equiv 0=-\langle\bar{v}, \bar{v}\rangle \phi$.
Now, we need to verify the Clifford relations. We have already seen that $f_{i} \circ f_{i} \circ \phi \equiv 0=\left\langle f_{i}, f_{i}\right\rangle \phi$ for all $i$. Also $f_{i} \circ f_{j} \circ \phi+f_{j} \circ f_{i} \circ \phi=2\left(f_{i} \wedge f_{j}+f_{j} \wedge f_{i}\right) \wedge \phi=0=-2\left\langle f_{i}, f_{j}\right\rangle \phi$. Similar relations hold for $\bar{f}_{i}$ 's. We just need to check the mixed relations, viz.,

$$
\begin{aligned}
f_{k} \circ\left(\overline{f_{j}} \circ \phi\right)+\bar{f}_{j} \circ\left(f_{k} \circ \phi\right) & \left.\left.=-2 f_{k} \wedge\left(f_{j}\right\lrcorner \phi\right)-2 f_{j}\right\lrcorner\left(f_{k} \wedge \phi\right) \\
& \left.\left.=-\left(e_{2 k-1}-i e_{2 k}\right) \wedge\left[\left(e_{2 j-1}-i e_{2 j}\right)\right\lrcorner \phi\right]-\left(e_{2 j-1}-i e_{2 j}\right) \wedge\left[\left(e_{2 k-1}-i e_{2 k}\right)\right\lrcorner \phi\right] \\
& \left.\left.\left.\left.=-\left[e_{2 k-1} \wedge e_{2 j-1}\right\lrcorner+e_{2 j-1} \wedge e_{2 k-1}\right\lrcorner\right] \phi-\left[e_{2 k} \wedge e_{2 j}\right\lrcorner+e_{2 j} \wedge e_{2 k}\right\lrcorner\right] \phi \\
& \left.\left.\left.\left.+i\left[e_{2 k} \wedge e_{2 j-1}\right\lrcorner+e_{2 j-1} \wedge e_{2 k}\right\lrcorner\right] \phi+i\left[e_{2 j} \wedge e_{2 k-1}\right\lrcorner+e_{2 k-1} \wedge e_{2 j}\right\lrcorner\right] \phi \\
& =-\left\langle e_{2 k-1}, e_{2 j-1}\right\rangle \phi-\left\langle e_{2 j}, e_{2 k}\right\rangle \phi=-2 \delta_{k j} \phi=-2\left\langle f_{k}, \bar{f}_{j}\right\rangle \phi
\end{aligned}
$$

since the relation (iii) in Lemma 14.1.4 implies that $(v \wedge w\lrcorner \phi+w \wedge v\lrcorner \phi)=\langle v, w\rangle \phi$. Similarly one checks for $\bar{f}_{k} \circ f_{j} \circ \phi$

This shows that the action "o" makes $\Lambda^{*} \mathrm{~Pa} C l\left(V_{\mathbb{C}}\right)=\mathbb{C l}(V) \mathbb{C}$-module. This module, call it $S$, is irreducible, because its complex dimension is the complex dimension of $\Lambda^{*} P$, i.e. $2^{m}$. In the second part of Proposition 14.1.12, we saw that the dimension of the unique $\mathbb{C} l_{2 m}$ Clifford $\mathbb{C}$-module is $2^{m}$. Hence this must be that module, provided we check that the action is not trivial, and that is obvious.

We also recall that $\tau_{2 m}^{2}=1$, and this module $\Lambda^{*} P$ will split into the $( \pm 1)$-eigenspaces $\Lambda^{ \pm} P$ of the chirality element $\tau_{2 m}$. Since $\tau_{2 m}$ anticommutes with all $v \in V_{\mathbb{C}}, \tau_{2 m}$ commutes with $\mathbb{C l}_{2 m}^{0}$ and anticommutes with $\mathbb{C} l_{2 m}^{1}$. Hence $\mathbb{C} l_{2 m}^{0} \circ \Lambda^{ \pm} P \subset \Lambda^{ \pm} P$ and $\mathbb{C} l_{2 m}^{1} \circ \Lambda^{ \pm} P \subset \Lambda^{\mp} P$. In other words, the grading $\Lambda^{*} P=\Lambda^{+} P \oplus \Lambda^{-} P$ makes $\Lambda^{*} P$ a $\mathbb{C}$ - supermodule over $\mathbb{C} l_{2 m}$. That is $S=S^{+} \oplus S^{-}$, with $S^{ \pm}:=\Lambda^{ \pm} P$.

It is also useful to identify $\Lambda^{+} P$ and $\Lambda^{-} P$ explicitly. To compute the action of the chirality element $\tau_{2 m}$, first note that

$$
f_{j} \bar{f}_{j}=2^{-1}\left(e_{2 j-1}-i e_{2 j}\right)\left(e_{2 j-1}+i e_{2 j}\right)=2^{-1}\left(-1-1+2 i e_{2 j-1} e_{2 j}\right)=\left(-1+i e_{2 j-1} e_{2 j}\right)
$$

and similarly $\bar{f}_{j} f_{j}=-1-i e_{2 j-1} e_{2 j}$ it follows that $i e_{2 j-1} e_{2 j}=\frac{1}{2}\left(f_{j} \bar{f}_{j}-\bar{f}_{j} f_{j}\right)$, so that

$$
\tau_{2 m}=i^{m}\left(e_{1} e_{2}\right)\left(e_{2} e_{3}\right) . . e_{2 m-1} e_{2 m}=2^{-m} \prod_{j=1}^{m}\left(f_{j} \bar{f}_{j}-\bar{f}_{j} f_{j}\right)
$$

Write a $k$-form $\phi \in \Lambda^{k} P$ as

$$
\phi=\alpha_{1}+f_{1} \wedge \beta_{1}
$$

where $\alpha_{1}$ and $\beta_{1}$ are independent of $f_{1}$. Then $f_{1} \circ \phi=\sqrt{2} f_{1} \wedge \alpha_{1}$, and $\left.\bar{f}_{1} \circ \phi=-\sqrt{2} f_{1}\right\lrcorner\left(f_{1} \wedge \beta_{1}\right)=-\sqrt{2} \beta_{1}$. Thus:

$$
\left.\bar{f}_{1} \circ\left(f_{1} \circ \phi\right)=-2 f_{1}\right\lrcorner\left(f_{1} \wedge \alpha_{1}\right)=-2 \alpha_{1}
$$

and

$$
f_{1} \circ\left(\bar{f}_{1} \circ \phi\right)=-2 f_{1} \wedge \beta_{1}
$$

Thus

$$
\left(f_{1} \bar{f}_{1}-\bar{f}_{1} f_{1}\right)\left(\alpha_{1}+f_{1} \wedge \beta_{1}\right)=2\left(\alpha_{1}-f_{1} \wedge \beta_{1}\right)
$$

Identical formulae hold for $f_{j}$ and $\bar{f}_{j}$, so that we have the following consequence for a decomposable form $\phi=f_{I}:=f_{i_{1}} \wedge f_{i_{2}} \wedge \ldots \wedge f_{i_{k}}$ :

$$
\begin{aligned}
\left(f_{j} \bar{f}_{j}-\bar{f}_{j} f_{j}\right) f_{I} & =-2 f_{I} \quad \text { whenever } j \in I \\
\left(f_{j} \bar{f}_{j}-\bar{f}_{j} f_{j}\right) f_{I} & =2 f_{I} \quad \text { whenever } j \notin I
\end{aligned}
$$

It follows that $\tau_{2 m} \circ f_{I}=2^{-m} \cdot 2^{m}(-1)^{k}(+1)^{m-k} f_{I}=(-1)^{k} f_{I}$. Hence $\tau_{2 m}$ acts as $(-1)^{k}$ on $\Lambda^{k} P$. Thus $S^{+}=\Lambda^{+} P=\Lambda^{e v} P$ and $S^{-}=\Lambda^{-} P=\Lambda^{o} P$.

Finally, consider the map:

$$
\begin{aligned}
\rho: \mathbb{C} l_{2 m} & \rightarrow \operatorname{hom}_{\mathbb{C}}(S, S) \\
x & \mapsto x \circ()
\end{aligned}
$$

Note that:

$$
\begin{aligned}
\operatorname{hom}_{\mathbb{C}}(S, S) & =\left(\Lambda^{*} P\right)^{*} \otimes \Lambda^{*} P=\Lambda^{*} P^{*} \otimes \Lambda^{*} P \\
& =\Lambda^{*} \bar{P} \otimes \Lambda^{*} P=\Lambda^{*}(\bar{P} \oplus \bar{P})=\Lambda^{*}\left(V_{\mathbb{C}}\right)
\end{aligned}
$$

Both sides have complex dimension $2^{2 m}$, and it is easy to check that $\rho$ has no kernel (exercise!). This proves that $\rho$ is an isomorphism and the proposition follows.

By a magical occurrence, the graded pieces of the unique irreducible $\mathbb{C} l_{2 m}$ supermodule $S$ above are the two distinct irreducible modules over $\mathbb{C} l_{2 m-1}$. More precisely:

Corollary 14.1.20 (Irreducible $\mathbb{C}$-modules over $\mathbb{C} l_{2 m-1}$ ). There is an isomorphism $C l_{n+1}^{0} \simeq C l_{n}$ of $\mathbb{R}$-algebras which complexifies to an isomorphism $\mathbb{C} l_{n+1}^{0} \simeq \mathbb{C} l_{n}$. If we consider the graded pieces $S^{ \pm}$of the irreducible $\mathbb{C} l_{2 m} \mathbb{C}$-supermodule $S$ of the previous Proposition 14.1 .19 , we have that $S^{ \pm}$are both $\mathbb{C}$-modules over $\mathbb{C} l_{2 m}^{0}$. Under the isomorphism above, they are $\mathbb{C}$-modules over $\mathbb{C} l_{2 m-1}$. Their complex dimensions are $2^{m-1}$, and they are precisely the two distinct irreducible $\mathbb{C}$-modules over $\mathbb{C} l_{2 m-1}$.

Proof: The map $f: \mathbb{R}^{n} \rightarrow C l_{n+1}^{0}$ is defined by $e_{i} \mapsto e_{i} e_{n+1}$ for $i=1,2, . ., n$. Now $f\left(e_{i}\right)^{2}=\left(e_{i} e_{n+1}\right)^{2}=-1$, and for $i \neq j$

$$
f\left(e_{i}\right) f\left(e_{j}\right)+f\left(e_{j}\right) f\left(e_{i}\right)=e_{i} e_{n+1} e_{j} e_{n+1}+e_{j} e_{n+1} e_{i} e_{n+1}=e_{i} e_{j}+e_{j} e_{i}=0
$$

So by the universal property of Clifford algebras, it extends to a $\mathbb{R}$ - algebra homomorphism $C l_{n} \rightarrow C l_{n+1}^{0}$. It is an isomorphism because it is clearly injective and both sides have the same dimension. Likewise for the complexifications.

Note that under the isomorphism $f: \mathbb{C} l_{2 m-1} \rightarrow \mathbb{C} l_{2 m}^{0}$, we have for the chirality element:

$$
\begin{aligned}
f\left(\tau_{2 m-1}\right) & =i^{m} f\left(e_{1} \ldots e_{2 m-1}\right)=i^{m}\left(e_{1} e_{2 m}\right)\left(e_{2} e_{2 m}\right) \ldots\left(e_{2 m-2} e_{2 m}\right)\left(e_{2 m-1} e_{2 m}\right) \\
& =(-1)^{m-1} i^{m} e_{1} e_{2}\left(e_{2 m}\right)^{2} e_{3} e_{4}\left(e_{2 m}\right)^{2} \ldots, e_{2 m-3} e_{2 m-2}\left(e_{2 m}\right)^{2}\left(e_{2 m-1} e_{2 m}\right) \\
& =(-1)^{m-1}(-1)^{m-1} i^{m} e_{1} \ldots e_{2 m}=i^{m} e_{1} e_{2} \ldots e_{2 m}=\tau_{2 m}
\end{aligned}
$$

Hence the module $S^{+}$over $\mathbb{C} l_{2 m}^{0}$ becomes a module over $\mathbb{C} l_{2 m-1}$ via the isomorphism $f$, and since $f\left(\tau_{2 m-1}\right)=$ $\tau_{2 m}$, it follows that $\tau_{2 m-1}$ acts as +1 on $S^{+}$. Similarly, $\tau_{2 m-1}$ acts as $(-1)$ on $S^{-}$. Since $\operatorname{dim}_{\mathbb{C}} S=2^{m}$, and $\operatorname{dim}_{\mathbb{C}} S^{ \pm}=\frac{1}{2} \operatorname{dim}_{\mathbb{C}} S=2^{m-1}$, it follows that $S^{ \pm}$are the two inequivalent $\mathbb{C} l_{2 m-1}$ irreducible $\mathbb{C}$-modules. The corollary follows.

Notation: Let us denote the two distinct irreducible $\mathbb{C} l_{2 m-1} \mathbb{C}$-modules by $S_{2 m-1}^{+}$and $S_{2 m-1}^{-}$, both of complex dimension $2^{m-1}$. Let us denote the unique irreducible $\mathbb{C} l_{2 m} \mathbb{C}$-module (which is a supermodule) by $S_{2 m}$, of complex dimension $2^{m}$. We note by the Corollary 14.1.20 that the graded pieces $S_{2 m}^{ \pm}$(both of complex dimension $2^{m-1}$ ) are precisely $S_{2 m-1}^{ \pm}$as $\mathbb{C}$-vector spaces, and their module structure over $\mathbb{C} l_{2 m}^{0}$ is precisely their module structure over $\mathbb{C} l_{2 m-1}$ under the identification $\mathbb{C} l_{2 m-1} \simeq \mathbb{C} l_{2 m}^{0}$.
14.2. Complex spin representations. We first note that since $\operatorname{Spin}(n) \subset C l_{n} \subset \mathbb{C} l_{n}$, any $\mathbb{C} l_{n} \mathbb{C}$-module will give a $\mathbb{C}$-module over $\operatorname{Spin}(n)$ by restricting the action, because the group mutiplication on $\operatorname{Spin}(n)$ is the Clifford multiplication in $\mathbb{C} l_{n}$. A similar remark applies to $\operatorname{Pin}(n)$, but they are of less concern to us here.

Proposition 14.2.1. On the spinor group $\operatorname{Spin}(2 m)$, there are two inequivalent irreducible $\mathbb{C}$-modules (=complex representations). They are denoted by $\Delta_{2 m}^{ \pm}$, and are distinguished by the sign of the chirality element $i^{m} \omega_{2 m}$ (where $\omega_{2 m}$ resides in $\operatorname{Spin}(2 m)$ ). Both are of complex dimension $2^{m-1}$, and are called the half-spin representations. They do not descend to $S O(2 m)$.

On the spinor group $\operatorname{Spin}(2 m-1)$, there is exactly one irreducible $\mathbb{C}$-module, of dimension $2^{m-1}$, and is denoted $\Delta_{2 m-1}$. It does not descend to $S O(2 m-1)$.

Proof: First note that $\omega_{2 m} \in \mathbb{C} l_{2 m}^{0}$, and since it is a product of unit vectors $e_{i}$, lies in $\operatorname{Pin}(2 m)$. Thus $\omega_{2 m} \in \operatorname{Spin}(2 m)$, and for a $\mathbb{C}$-module over $\operatorname{Spin}(2 m)$, the action of $i^{m} \omega_{2 m}$ makes sense. Now, by the Proposition 14.1.19, there is the unique $\mathbb{C}$-supermodule $S_{2 m}$ over $\mathbb{C} l_{2 m}$, with graded pieces $S_{2 m}^{ \pm}$. Both of these graded pieces are $\mathbb{C}$-modules over $\mathbb{C} l_{2 m}^{0}$. Hence both are modules over $\operatorname{Spin}(2 m) \subset \mathbb{C} l_{2 m}^{0}$. Call them $\Delta_{2 m}^{ \pm}$. They are distinguished by the sign of the chirality action $i^{m} \omega_{2 m}$, (or $i^{m} \rho\left(\omega_{2 m}\right)$ to be more precise, where $\rho$ is the representation on $S_{2 m}$ ).

It is clear that $\mathbb{C l}_{2 m}^{0}$ is generated as an algebra by elements of $\operatorname{Spin}(2 m)$ (indeed all elements $e_{I}$ with $|I|$ even are in $\operatorname{Spin}(2 m)$ ), it follows that if these modules $\Delta_{2 m}^{ \pm}$are reducible as $\operatorname{Spin}(2 m)$ modules, they will be reducible as $\mathbb{C} l_{2 m}^{0}$ modules. That is $S_{2 m}^{ \pm}$will be reducible as $\mathbb{C} l_{2 m-1}$ modules. But we have seen in Corollary 14.1.20 that they are precisely the two irreducible $\mathbb{C} l_{2 m-1}$ modules. Thus $\Delta_{2 m}^{ \pm}$are both irreducible $\mathbb{C}$-modules over $\operatorname{Spin}(2 m)$. Their dimensions are given by:

$$
\operatorname{dim}_{\mathbb{C}} \Delta_{2 m}^{ \pm}=\operatorname{dim}_{\mathbb{C}} S_{2 m}^{ \pm}=2^{m-1}
$$

It is also clear from the construction of the $\mathbb{C}$-supermodule $S=\Lambda^{*} P$ over $\mathbb{C} l_{2 m}$ in Proposition 14.1.19 that $-1 \in \mathbb{C} l_{2 m}$ acts as $\left(-I d_{S}\right)$ on $S$, and hence $(-1) \in \operatorname{Spin}(2 m)$ acts as $-I d$ on both $S_{2 m}^{ \pm}$, so neither representation $\Delta_{2 m}^{ \pm}$descends to $S O(2 m)$.

For the odd spin representations, we start out with the two distinct irreducible $\mathbb{C}$-modules $S_{2 m-1}^{ \pm}$over $\mathbb{C} l_{2 m-1}$. This time around, the volume element $\omega_{2 m-1}$ is of odd parity, and lives in $C l_{2 m-1}^{1}$. Hence $\omega_{2 m-1}$ does not live in $\operatorname{Spin}(2 m-1)$. Hence the action of $\operatorname{Spin}(2 m-1)$ is completely determined by the action of $C l_{2 m-1}^{0}$ on $S_{2 m-1}^{ \pm}$.

We claim that the action of $\mathbb{C} l_{2 m-1}^{0}$ is identical on both the irreducibles $S_{2 m-1}^{ \pm}$. Indeed if we let $\alpha_{n}: \mathbb{C} l_{n} \rightarrow$ $\mathbb{C} l_{n}$ be the involution defined by extending the map $v \rightarrow(-v)$ of $V=\mathbb{R}^{n}$ to $\mathbb{C} l_{n}$, we have $\mathbb{C} l_{n}^{0}$ (resp. $\mathbb{C} l_{n}^{-}$is the $(+1)$ (resp. $(-1)$ )-eigenspace of $\alpha_{n}$. Since $\omega_{2 m-1} \in \mathbb{C} l_{2 m-1}^{1}$, it follows that $\alpha_{2 m-1}\left(\omega_{2 m-1}\right)=-\omega_{2 m-1}$. Hence $\alpha_{2 m-1}$ interchanges the +1 and -1 eigenspaces of $\tau_{2 m-1}$ on $\mathbb{C} l_{2 m-1}$, and so interchanges $\mathbb{C} l_{2 m-1}^{ \pm}$, the two summands of $\mathbb{C} l_{2 m-1}$. Thus $\mathbb{C l} l_{2 m-1}^{0}$ is the diagonal subalgebra in the direct sum $\mathbb{C} l_{2 m-1}=\mathbb{C} l_{2 m-1}^{+} \oplus \mathbb{C l}_{2 m-1}^{-}=$ $\mathbb{C}\left(2^{m-1}\right) \oplus \mathbb{C}\left(2^{m-1}\right)$. Hence the two distinct irreducible modules $\mathbb{C}^{2^{m-1}}$, coming from the action of each matrix algebra summand, will receive the same action from the diagonal $\mathbb{C} l_{2 m-1}^{0}$. Hence the claim.

So we may define $\Delta_{2 m-1}$ to be either $S_{2 m-1}^{+}$or $S_{2 m-1}^{-}$(it doesn't matter which) with $\operatorname{Spin}(2 m-1)$ action being the restriction of the $\mathbb{C} l_{2 m-1}$ action. The proofs of the other statements are similar to the even case above.

### 14.3. Inner products, orthogonality and unitarity.

Definition 14.3.1. Let $W$ be an $\mathbb{R}$-module (resp. $\mathbb{C}$-module) over $C l(V)$, and let $(-,-)$ be a positive definite inner product (resp. positive definite hermitian inner product) on $W$. (We are using a different symbol to distinguish it from the euclidean inner product $\langle-,-\rangle$ on $V$ with respect to which the Clifford algebra $C l(V)$ is defined.) We say that $W$ is a self-adjoint module over $C l(V)$ if

$$
(x . v, w)=(-1)^{\operatorname{deg} x}\left(v, x^{*} w\right) \text { for all } v, w \in W, \quad x \text { homogeneous } \in C l(V)
$$

where $*$ is the anti-isomorphism defined in Definition 13.1.6. This is clearly equivalent to

$$
(e . v, w)=-(v, e . w) \quad \text { for all } e \in V, \quad v, w \in W
$$

i.e. the Clifford action of vectors should be skew-adjoint with respect to $(-,-)$.

Example 14.3.2. By the last remark above, the second part of Proposition 14.1.5 implies that the action of $C l(V)$ on $\Lambda^{*}(V)$ is self-adjoint, with the inner product $(-,-)$ on $\Lambda^{*}(V)$ being the natural inner product $\langle-,-\rangle$, induced by the one on $V$.

Example 14.3.3. We recall the construction of the unique irreducible $\mathbb{C}$-supermodule $S_{2 m}$ over $C l_{2 m}$ (equivalently $\mathbb{C}_{2 m}$ ) in Proposition 14.1.19. Recall that $V_{\mathbb{C}}=P \oplus \bar{P}$, where $V=\mathbb{R}^{2 m}$. We already have the complexification $\langle-,-\rangle$ on $\Lambda^{*}\left(V_{\mathbb{C}}\right)$ of the real inner product $\langle-,-\rangle$ on $\Lambda^{*}(V)$ (alluded to in the foregoing example). This is an inner product on $\Lambda^{*}\left(V_{\mathbb{C}}\right)=\Lambda^{*}(V) \otimes \mathbb{C}$ which is complex linear in both slots. This inner-product satisfies:

$$
\langle\overline{\phi \otimes \lambda}, \overline{\psi \otimes \mu}\rangle=\langle\phi \otimes \bar{\lambda}, \psi \otimes \bar{\mu}\rangle=\langle\phi, \psi\rangle \bar{\lambda} \bar{\mu}=\overline{\langle\phi, \psi\rangle \lambda \mu}=\overline{\langle\phi \otimes \lambda, \psi \otimes \mu\rangle} \text { for all } \phi, \psi \in \Lambda^{*}(V)
$$

that is,

$$
\begin{equation*}
\overline{\langle\phi, \psi\rangle}=\langle\bar{\phi}, \bar{\psi}\rangle \quad \text { for all } \phi, \psi \in \Lambda^{*}\left(V_{\mathbb{C}}\right)=\Lambda^{*}(V) \otimes \mathbb{C} \tag{43}
\end{equation*}
$$

We have the complex conjugation $P \rightarrow \bar{P}$, which maps $\Lambda^{*} P \rightarrow \Lambda^{*} \bar{P}$ inside $\Lambda^{*}\left(V_{\mathbb{C}}\right)$. So define a hermitian inner product on $S_{2 m}=\Lambda^{*} P$ by:

$$
(\phi, \psi):=\langle\phi, \bar{\psi}\rangle \quad \text { for } \phi, \psi \in \Lambda^{*} P
$$

For $e \in V \subset V_{\mathbb{C}}$, we have $e=\bar{e}$. Let $\phi=w_{1} \wedge w_{2} \ldots \wedge w_{k} \in \Lambda^{*} P$, with $w_{i} \in P$. Then, by Exercise 14.1.6:

$$
\begin{aligned}
e\lrcorner \bar{\phi} & =e\lrcorner\left(\bar{w}_{1} \wedge \bar{w}_{2} \ldots \wedge \bar{w}_{k}\right)=\sum_{i}(-1)^{i}\left\langle e, \bar{w}_{i}\right\rangle\left(\bar{w}_{1} \wedge \bar{w}_{2} \ldots \wedge \widehat{\widehat{w}_{i}} \wedge \ldots \wedge \bar{w}_{k}\right) \\
& =\sum_{i}(-1)^{i} \overline{\left\langle\bar{e}, w_{i}\right\rangle} \overline{\left(w_{1} \wedge w_{2} . . \wedge \widehat{w_{i}} \wedge \ldots \wedge w_{k}\right)}=\overline{e\lrcorner \phi}
\end{aligned}
$$

using $e=\bar{e}$ and the equation (43) above. Now, using the definition of the Clifford action in Proposition 14.1.19 and (i) of Lemma 14.1.4, we compute:

$$
(e . \psi, \phi)=\sqrt{2}\langle e \wedge \psi, \bar{\phi}\rangle=\sqrt{2}\langle\psi, e\lrcorner \bar{\phi}\rangle=\sqrt{2}\langle\psi, \overline{e\lrcorner \phi}\rangle=-(\psi, e . \phi)
$$

which shows that Clifford multiplication by elements of $V$ is skew- adjoint with respect to this hermitian inner product $(-,-)$, and hence the module $S_{2 m}$ is self-adjoint over $C l_{2 m}$.

Exercise 14.3.4. Are the irreducible modules $S_{2 m-1}^{ \pm}$self- adjoint as Clifford modules over $C l_{2 m-1}$ ?

Here is an important property of self-adjoint Clifford modules.

Proposition 14.3.5. Let $W$ be a self-adjoint $\mathbb{R}$-module (resp. $\mathbb{C}$ - module) over $C l_{n}$ with respect to the positive definite real (resp. positive definite hermitian) inner product. Then if we consider $W$ as a module over $\operatorname{Spin}(n) \subset C l_{n}$, the resulting representation

$$
\rho: \operatorname{Spin}(n) \rightarrow G L(W)
$$

is orthogonal (resp. unitary).

Proof: From (iii) of the Proposition 13.2.2, we have $g \in \operatorname{Spin}(n)$ implies $\operatorname{deg} g=0$ and $g^{*} g=1$, so that by self adjointness of $W$,

$$
\left(g \cdot w_{1}, g \cdot w_{2}\right)=\left(w_{1}, g^{*} g w_{2}\right)=\left(w_{1}, w_{2}\right) \text { for } g \in \operatorname{Spin}(n), w_{i} \in W
$$

which proves the proposition.

Corollary 14.3.6. The representation of $\operatorname{Spin}(n)$ on $\Lambda^{*}\left(\mathbb{R}^{n}\right)$ is a (special) orthogonal representation. The two complex half-spin representations $\Delta_{2 m}^{ \pm}$of $\operatorname{Spin}(2 m)$ are unitary representations.
14.4. Decomposition formulae for $\operatorname{Spin}(2 m)$ representations. We would now like to relate the left $\operatorname{Spin}(2 n)$ module $\mathbb{C} l_{2 m}$, the left $\operatorname{Spin}(2 m)$ module $\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m}\right)$, as well as the lifted representations of $S O(2 m)$ modules $\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m}\right)$ and its $S O(2 m)$-submodules $\Lambda_{\mathbb{C}}^{e v}, \Lambda_{\mathbb{C}}^{o}, \Lambda_{\mathbb{C}}^{ \pm}$etc., with the irreducible half-spin representations $\Delta_{2 m}^{ \pm}$constructed in the Proposition 14.2.1.

## Proposition 14.4.1.

(i): Consider $\mathbb{C} l_{2 m}$ as a left module over itself, by left multiplication. Then $\mathbb{C} l_{2 m}$ decomposes into $2^{m}$ irreducible $\mathbb{C} l_{2 m}$-modules $V_{\epsilon}$ where $\epsilon=\left(\epsilon_{1}, \epsilon_{2}, . ., \epsilon_{m}\right)$ with each $\epsilon_{i}= \pm 1$. Each $V_{\epsilon}$ is isomorphic to the unique irreducible supermodule $S_{2 m}$ as a $\mathbb{C} l_{2 m}$-module. $V_{\epsilon}$ further decomposes into the two complex subspaces $V_{\epsilon}^{ \pm}$via the chirality left action of $\tau_{2 m}$, so that $V_{\epsilon}^{ \pm} \simeq S_{2 m}^{ \pm}$.
(ii): Consider $\mathbb{C} l_{2 m}$ as a $\operatorname{Spin}(2 m)$ complex module by the restricted left Clifford multiplication action from $\mathbb{C} l_{2 m}$. Then as a $\operatorname{Spin}(2 m)$ module we have $\mathbb{C} l_{2 m}=2^{m} \Delta_{\epsilon}^{+} \oplus 2^{m} \Delta_{\epsilon}^{-}$, where $\Delta_{\epsilon}^{ \pm}$are isomorphic to the distinct irreducible half-spin representations $\Delta_{2 m}^{ \pm}$respectively, as $\operatorname{Spin}(2 m)$ modules.
(iii): The complex exterior algebra $\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m}\right)$ considered as a $\mathbb{C} l_{2 m}$ module as in Proposition 14.1.5 has a decomposition into irreducibles analogous to (i) over $\mathbb{C} l_{2 m}$, and a decomposition analogous to (ii) above, as a $\operatorname{Spin}(2 m)$ module.

Proof: Consider the elements of $\mathbb{C} l_{2 m}$ defined by:

$$
\alpha_{j}:=i e_{2 j-1} e_{2 j} \quad j=1,2, \ldots, m
$$

Then it easily follows that:
(a) $\quad \alpha_{j} \alpha_{k}=\alpha_{k} \alpha_{j}$ for all $1 \leq k, j \leq m$
(b) $\quad \alpha_{j}^{2}=1 \quad$ for all $1 \leq j \leq m$

Now consider the right-multiplication action of $\alpha_{j}$ on $\mathbb{C l}_{2 m}$. By (a) and (b) above, $\mathbb{C} l_{2 m}$ breaks up into simultaneous eigenspaces $V_{\epsilon}$, where $\alpha_{j}$ acts by $\epsilon_{j}$ on $V_{\epsilon}$, and $\epsilon_{j}=+1$ or -1 . Since right and left multiplication commute, each $V_{\epsilon}$ is a left $\mathbb{C} l_{2 m}$-submodule of $\mathbb{C} l_{2 m}$ under left action. Noting that

$$
e_{2 j-1} \alpha_{j}=i e_{2 j-1} e_{2 j-1} e_{2 j}=-i e_{2 j-1} e_{2 j} e_{2 j-1}=-\alpha_{2 j} e_{2 j-1}
$$

it follows that right multiplication by $e_{2 j-1}$ will map $V_{\epsilon}$ isomorphically to $V_{\epsilon^{\prime}}$ as a $\mathbb{C} l_{2 m}$-module where $\epsilon_{k}^{\prime}=\epsilon_{k}$ for $k \neq j$ and $\epsilon_{j}^{\prime}=-\epsilon_{j}$. Thus all the $V_{\epsilon}$ are isomorphic to $V_{(+1,+1, \ldots,+1)}$ as $\mathbb{C}_{2 m}$-modules. Thus $\operatorname{dim}_{\mathbb{C}} V_{\epsilon}=$ $\frac{1}{2^{m}} \operatorname{dim}_{\mathbb{C}} \mathbb{C} l_{2 m}=2^{m}$. It follows for reasons of dimension that each $V_{\epsilon}$ is irreducible and $V_{\epsilon} \simeq S_{2 m}$ as a left $\mathrm{Cl}_{2 m}$-module.

Thus $V_{\epsilon}=V_{\epsilon}^{+} \oplus V_{\epsilon}^{-}$, where $V_{\epsilon}^{ \pm}$are the $( \pm 1)$-eigenspaces of left multiplication by chirality $\tau_{2 m}$. Clearly $V_{\epsilon}^{ \pm} \simeq S_{2 m}^{ \pm}$as $\mathbb{C l}_{2 m}^{0}$ - modules.

Now (ii) is clear by setting $\Delta_{\epsilon}^{ \pm}=V_{\epsilon}^{ \pm}$with $\operatorname{Spin}(2 m)$ action being restriction of $\mathbb{C} l_{2 m}^{0}$ action and the Proposition 14.2.1.
(iii) follows by noting that $\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m}\right)=\mathbb{C} l_{2 m} .1$ where $1 \in \Lambda_{\mathbb{C}}^{0}\left(\mathbb{R}^{2 m}\right)$. The proposition follows.

Definition 14.4.2 (Some $\mathbb{C} l_{2 m}$-bimodules). We note that $\mathbb{C} l_{2 m}$ has both a left $\mathbb{C} l_{2 m}$-module structure by left multiplication, and a right $\mathbb{C} l_{2 m}$-module structure by right multiplication, which can be thought of as a left module structure by $x . z:=z x^{*}$. Hence $\mathbb{C} l_{2 m}$ may be thought of as a $\mathbb{C} l_{2 m} \otimes \mathbb{C} l_{2 m}$ left-module, viz. $(x \otimes y) \circ z:=x . z . y^{*}$. Such a thing is called a $\mathbb{C} l_{2 m}$-bimodule.

Now we recall the algebra isomorphism:

$$
\mathbb{C} l_{2 m} \rightarrow \operatorname{hom}_{\mathbb{C}}\left(S_{2 m}, S_{2 m}\right)
$$

from Proposition 14.1.19. On the right side, we can again produce two $\mathbb{C l}_{2 m}$-module structures. Namely $(x . T)(w):=x T(w)$ for $x \in \mathbb{C} l_{2 m}$ and $w \in S_{2 m}$, and also $(x \circ T)(w)=T\left(x^{*} w\right)$ for $x \in \mathbb{C} l_{2 m}$ and $w \in S_{2 m}$. This is again a $\mathbb{C} l_{2 m}$ bimodule structure, or left $\mathbb{C} l_{2 m} \otimes \mathbb{C} l_{2 m}$-module structure given by $(x \otimes y) \circ T=x T\left(y^{*}-\right)$.

We now have the following proposition:

Proposition 14.4.3. The isomorphism $\mathbb{C} l_{2 m} \simeq \operatorname{hom}_{\mathbb{C}}\left(S_{2 m}, S_{2 m}\right)$ is an isomorphism of $\mathbb{C} l_{2 m} \otimes \mathbb{C} l_{2 m}$ modules (i.e. $\mathbb{C} l_{2 m}$-bimodules). In particular, by restricting to the diagonal subalgebra $\mathbb{C} l_{2 m} \subset \mathbb{C} l_{2 m} \otimes \mathbb{C} l_{2 m}$, we have that the adjoint action of $\mathbb{C} l_{2 m}$ on itself is the same as the adjoint action of $\mathbb{C l}_{2 m}$ on $\operatorname{hom}_{\mathbb{C}}\left(S_{2 m} \otimes S_{2 m}\right)$ by $x \cdot T:=x . T .\left(x^{*}-\right)$.

Proof: We note that $\mathbb{C} l_{2 m}=\mathbb{C}\left(2^{m}\right)$ as an algebra, and so $\mathbb{C} l_{2 m} \otimes \mathbb{C} l_{2 m}=\mathbb{C}\left(2^{m}\right) \otimes \mathbb{C}\left(2^{m}\right)=\mathbb{C}\left(2^{2 m}\right)$. But $\mathbb{C}\left(2^{2 m}\right)$ is precisely $\mathbb{C} l_{4 m}$. Thus a $\mathbb{C} l_{2 m} \otimes \mathbb{C} l_{2 m}$ left-module structure (or $\mathbb{C} l_{2 m}$-bimodule structure) is precisely a left $\mathbb{C} l_{4 m}$-module structure. Since $\operatorname{dim}_{\mathbb{C}} \mathbb{C} l_{2 m}=2^{2 m}=\operatorname{dim}_{\mathbb{C}}\left(S_{2 m}, S_{2 m}\right)$, and both of these $\mathbb{C} l_{4 m}$ modules are non-trivial, it follows that both modules are isomorphic as $\mathbb{C} l_{4 m}$-modules to the unique irreducible $\mathbb{C} l_{4 m^{-}}$ module $S_{4 m}$. That is, they are isomorphic as $\mathbb{C} l_{2 m} \otimes \mathbb{C} l_{2 m}$-modules, proving the first assertion. The second assertion clearly follows from the first.

Now we can consider the lifted modules from $S O(2 m)$. That is, let

$$
\rho: \operatorname{Spin}(2 m) \rightarrow S O(2 m)
$$

be the 2 -covering defined in the Proposition 13.2.2. Then the modules $\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m}\right)$ is a natural $S O(2 m)$ module by the action which is defined on decomposables in $\Lambda^{k}$ by :

$$
g \cdot\left(v_{1} \wedge v_{2} \wedge \ldots \wedge v_{k}\right)=\left(g v_{1} \wedge g v_{2} \ldots \wedge g v_{k}\right)
$$

Clearly this action preserves $\Lambda_{\mathbb{C}}^{e v}$ and $\Lambda_{\mathbb{C}}^{o}$. Also it is easily checked that this action preserves the volume element $e_{1} \wedge e_{2} \wedge \ldots \wedge e_{n}$, as well as the positive definite inner product $\langle-,-\rangle$ on $\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m}\right)$, so it commutes with the Hodge-star operator $*$. Hence $\Lambda_{\mathbb{C}}^{ \pm}$are also $S O(2 m)$ submodules of $\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m}\right)$. Thus for $W$ being any of these $S O(2 m)$-modules, the composite map:

$$
\operatorname{Spin}(2 m) \xrightarrow{\rho} S O(2 m) \rightarrow \operatorname{hom}_{\mathbb{C}}(W, W)
$$

makes $W$ into a "lifted" $\operatorname{Spin}(2 m)$-module.

Proposition 14.4.4 (Decomposition of lifted $\operatorname{Spin}(2 m)$-modules). We have the following identities:
(i): The lifted $\operatorname{Spin}(2 m)$ module $\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m}\right)$ is isomorphic to $\mathbb{C} l_{2 m}$ (with adjoint action of $\operatorname{Spin}(2 m)$ ) as a $\operatorname{Spin}(2 m)$-module. It is isomorphic to $\Delta_{2 m} \otimes \Delta_{2 m}$ (where $\operatorname{Spin}(2 m)$ acts by tensor product action $(x .(v \otimes w)):=x v \otimes x w)$. That is, the lifted module $\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m}\right)$ has a "square root" $\Delta_{2 m}$.
(ii): The isomorphism in (i) above maps the $\operatorname{Spin}(2 m)$-submodule $\Lambda_{\mathbb{C}}^{+}$(resp. $\Lambda_{\mathbb{C}}^{-}$) of the lifted module $\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m}\right)$ isomorphically to the $\operatorname{Spin}(2 m)$-submodule $\Delta_{2 m}^{+} \otimes \Delta_{2 m}\left(\right.$ resp. $\left.\Delta_{2 m}^{-} \otimes \Delta\right)$ of $\Delta_{2 m} \otimes \Delta_{2 m}$.
(iii): The isomorphism of (i) above maps the lifted $\operatorname{Spin}(2 m)$-submodule $\Lambda_{\mathbb{C}}^{e v}$ of $\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m}\right)$ isomorphically to the submodule $\left((-1)^{m} \Delta_{2 m}^{+} \otimes \Delta_{2 m}^{+}\right) \oplus\left((-1)^{m} \Delta_{2 m}^{-} \otimes \Delta_{2 m}^{-}\right)$of $\Delta_{2 m} \otimes \Delta_{2 m}$. Similarly, it maps the submodule $\Lambda_{\mathbb{C}}^{o}$ isomorphically to $\left((-1)^{m} \Delta_{2 m}^{+} \otimes \Delta_{2 m}^{-}\right) \oplus\left((-1)^{m} \Delta_{2 m}^{-} \otimes \Delta_{2 m}^{+}\right)$of $\Delta_{2 m} \otimes \Delta_{2 m}$.

Proof: We note that for the $\mathbb{C}$-basis element $e_{i_{1}} \wedge e_{i_{2}} \ldots \wedge e_{i_{k}}$ of $\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m}\right)$, the lifted action of $x \in \operatorname{Spin}(2 m)$ is given by:

$$
x .\left(e_{i_{1}} \wedge e_{i_{2}} \ldots \wedge e_{i_{k}}\right):=\rho(x) e_{i_{1}} \wedge \rho(x) e_{i_{2}} \wedge \ldots \wedge \rho(x) e_{i_{k}}=x e_{i_{1}} x^{*} \wedge x e_{i_{2}} x^{*} \ldots \wedge x e_{i_{k}} x^{*}
$$

Now, under the $\mathbb{C}$-vector space isomorphism of "quantisation" (see (ii) of Proposition 14.1.7) identifying $\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m}\right)$ with $\mathbb{C} l_{2 m}$, the element on the right goes to

$$
x e_{i_{1}} x^{*} x e_{i_{2}} x^{*} \ldots x e_{i_{k}} x^{*}=x\left(e_{i_{1}} e_{i_{2}} \ldots . e_{i_{k}}\right) x^{*}
$$

which is precisely the adjoint action of $\mathbb{C} l_{2 m}$ on itself. Hence the lifted $\operatorname{Spin}(2 m)$ module $\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m}\right)$ is isomorphic to $\mathbb{C} l_{2 m}$ with adjoint $\operatorname{Spin}(2 m)$ action.

We have seen in the second assertion of the Proposition 14.4 .3 above that the $\mathbb{C} l_{2 m}$-module $\mathbb{C} l_{2 m}$ with adjoint action is isomorphic to the $\mathbb{C} l_{2 m}$-module $\operatorname{hom}_{\mathbb{C}}\left(S_{2 m}, S_{2 m}\right)$ (also with adjoint action of $\mathbb{C} l_{2 m}$.) Restricting both modules to $\operatorname{Spin}(2 m)$ shows that $\mathbb{C} l_{2 m}$ with adjoint action is isomorphic to hom $\mathbb{C}\left(\Delta_{2 m}, \Delta_{2 m}\right)=\Delta_{2 m} \otimes \Delta_{2 m}^{*}$
as a $\operatorname{Spin}(2 m)$ module. We can identify the contragredient module $\Delta_{2 m}^{*}$ with the right action by $x^{*}\left(=x^{-1}\right)$ with the left-module $\Delta_{2 m}$ with left action by $x$. This proves (i).

We know that the chirality element $\tau_{2 m}$ commutes with all elements in $\mathbb{C} l_{2 m}^{0}$, and hence $\tau_{2 m}\left(x y x^{*}\right)=$ $x\left(\tau_{2 m} y\right) x^{*}$ for $x \in \mathbb{C} l_{2 m}^{0}$, and in particular $x \in \operatorname{Spin}(2 m)$. Also $\tau_{2 m}^{2}=1$ implies that the splitting of $\mathbb{C} l_{2 m}$ as a lifted $\operatorname{Spin}(2 m)$ module into $( \pm 1)$-eigenspaces $\mathbb{C} l_{2 m}^{ \pm}$makes $\mathbb{C} l_{2 m}^{ \pm}$into $\operatorname{Spin}(2 m)$-submodules. So we need to know the $( \pm 1)$-eigenspaces of $\Delta_{2 m} \otimes \Delta_{2 m}$ under left multiplication by $\tau_{2 m}$. By the Proposition 14.4.3, this is just the action $\tau_{2 m}(x \otimes y)=\tau_{2 m} x \otimes y$. Thus the splitting is $\Delta_{2 m}^{ \pm} \otimes \Delta_{2 m}$. So $\mathbb{C} l_{2 m}^{ \pm}=\Delta_{2 m}^{ \pm} \otimes \Delta_{2 m}$. Using the isomorphism of $\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m}\right)$ with $\mathbb{C} l_{2 m}$, we get (ii).

For (iii), note that the conjugation action of $\omega_{2 m}$ satisfies

$$
\omega_{2 m} e_{i} \omega_{2 m}^{*}=-e_{i} \omega_{2 m} \omega_{2 m}^{*}=-(-1)^{2 m} e_{i}=-e_{i}
$$

which shows that $\rho\left(\omega_{2 m}\right)$ acts as +1 on $e_{I}$ with $I$ of even cardinality and $(-1)$ on $I$ of odd cardinality. Under the identification of $\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m}\right)$ with $\mathbb{C} l_{2 m}$ by (i), we find that $\Lambda_{\mathbb{C}}^{e v}$ is the submodule corresponding to ( +1 )eigenspace of $\rho\left(\omega_{2 m}\right)$, and $\Lambda_{\mathbb{C}}^{o}$ the (-1)- eigenspace of $\rho\left(\omega_{2 m}\right)$. So it remains to identify, in view of (i), the $\pm 1$-eigenspaces of the operator $\omega_{2 m} \otimes \omega_{2 m}$ on $\Delta_{2 m} \otimes \Delta_{2 m}$. Note that since $\omega_{2 m}$ commutes with $\mathbb{C} l_{2 m}^{0}$, it commutes with all of $\operatorname{Spin}(2 m)$, and so these $\pm 1$-eigenspaces are $\operatorname{Spin}(2 m)$ - submodules of $\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m}\right)$.

Since $i^{m} \omega_{2 m}=\tau_{2 m}$, we have $\omega_{2 m} \otimes \omega_{2 m}=(-1)^{m} \tau_{2 m} \otimes \tau_{2 m}$. Using the notation $(-1)^{m} \Delta_{2 m}^{+}:=\Delta_{2 m}^{+}$for $m$ even, and $\Delta_{2 m}^{-}$for $m$ odd, (and a similar notation for $(-1)^{m} \Delta_{2 m}^{-}$) we find that the +1 -eigenspace (resp. (-1)-eigenspace) of $\omega_{2 m} \otimes \omega_{2 m}$ is clearly $\left((-1)^{m} \Delta_{2 m}^{+} \otimes \Delta_{2 m}^{+}\right) \oplus\left((-1)^{m} \Delta_{2 m}^{-} \otimes \Delta_{2 m}^{-}\right)\left(\right.$resp. $\left((-1)^{m} \Delta_{2 m}^{+} \otimes \Delta_{2 m}^{-}\right) \oplus$ $\left.\left((-1)^{m} \Delta_{2 m}^{-} \otimes \Delta_{2 m}^{+}\right)\right)$. This proves (iii) and the proposition follows.

There is a fact about the "derived" adjoint action we shall need later on:

Proposition 14.4.5. The vector subspace spanned by $\left\{e_{i} e_{j}: i<j\right\}$ inside the real Clifford algebra $C L(V)$ is denoted by $C^{2}(V)$ (Recalling the quantisation map $c$ of (ii) in Proposition 14.1.7, $C^{2}(V)=c\left(\Lambda^{2}(V)\right)$. Then
(i): $C^{2}(V)$ is a Lie algebra under the commutator $[x, y]=x y-y x$ in $C l(V)$.
(ii): The map $\tau: C^{2}(V) \rightarrow \mathfrak{s o}(V)$ defined by $\tau(a) v=[a, v]$ is an isomorphism of Lie algebras.
(iii): Define the exponential map of $C(V)$ by:

$$
\begin{aligned}
\exp _{C}: C(V) & \rightarrow C(V) \\
x & \mapsto 1+x+\frac{x^{2}}{2!}+\ldots+\frac{x^{k}}{k!}+\ldots
\end{aligned}
$$

Then $\exp _{C}\left(C^{2}(V)\right)=\operatorname{Spin}(V)$.

Proof: By directly using $e_{i}^{2}=-1$ and $e_{i} e_{j}=-e_{j} e_{i}$ we compute:

$$
\begin{aligned}
{\left[e_{i} e_{j}, e_{k} e_{l}\right] } & =0 \text { if } i<j, k<l,\{i, j\} \cap\{k, l\}=\phi \text { or }(i, j)=(k, l) \\
& =-2 e_{i} e_{l} \text { if } i<j=k<l \\
& =2 e_{j} e_{l} \text { if } i=k, j \neq l
\end{aligned}
$$

which shows that $C^{2}(V)$ is a Lie algebra and (i) follows.
Note also that

$$
\begin{aligned}
{\left[e_{i} e_{j}, e_{k}\right] } & =-2 e_{i} \text { for } i<j=k \\
& =2 e_{j} \text { for } k=i<j \\
& =0 \text { for } k \neq i, k \neq j
\end{aligned}
$$

This clearly shows that $\tau\left(e_{i} e_{j}\right)$ for $i<j$ preserves $V=\operatorname{span}_{\mathbb{R}}\left\{e_{i}\right\}$, and hence maps to $g l(V)$. Since $\tau\left(e_{i} e_{j}\right)=$ $2\left(E_{j i}-E_{i j}\right),\left(E_{i j}\right.$ being the matrix with 1 in the $(i j)$ - spot and zeros elsewhere), and since the combinations
the set $\left\{\left(E_{j i}-E_{i j}\right): i<j\right\}$ constitutes a basis of $\mathfrak{s o}(V)$, it follows that $\tau$ is a vector space isomorphism. That it is a Lie algebra isomorphism follows easily from the fact that

$$
[\tau(x), \tau(y)] v=\tau(x)([y, v])-\tau(y)([x, v])=[x,[y, v]]-[y,[x, v]]==-[v,[x, y]]=\tau([x, y]) v
$$

by the Jacobi identity. This proves (ii).

In the course of proving (v) of Proposition 13.2.2, we found that $\exp _{C}\left(t e_{i} e_{j}\right)=\cos t .1+\sin t\left(e_{i} e_{j}\right)$ for $i \neq j$, and consequently Lie $(\operatorname{Spin}(V))$ was precisely $C^{2}(V)$. Now the exponential of $C^{2}(V)$ is going to be a connected Lie- subgroup $G \subset C l^{\times}$, and of dimension $\frac{n(n-1)}{2}$. Also its Lie algebra is $C^{2}(V)$. Since a connected compact Lie group is precisely the exponential of its Lie algebra, it follows that $G=\exp _{C} V=\operatorname{Spin}(V)$, and (iii) follows.

There is another crucial proposition which allows us to recover any $\mathbb{C} l_{2 m}$-supermodule as a tensor product with the irreducible $\mathbb{C} l_{2 m}$ supermodule $S_{2 m}$.

Proposition 14.4.6. Let $W$ be any $\mathbb{C}_{2 m}$-module with chirality grading $W^{ \pm}$. Then there exists a $\mathbb{C}$-vector space $V$ such that $W \simeq S_{2 m} \otimes_{\mathbb{C}} V$ as a $\mathbb{C} l_{2 m}$-supermodule. This $V$ is uniquely determined by $W$, and is called the twisting space for the supermodule $W$.

Proof: In the statement, we are treating $V$ as an ungraded $\mathbb{C}$-vector space, and equipping $S_{2 m} \otimes_{\mathbb{C}} V$ with the obvious left $\mathbb{C} l_{2 m}$-module structure defined by. $x .(s \otimes v)=x s \otimes v$. The supermodule structure on $S_{2 m} \otimes_{\mathbb{C}} V$ is defined by $\left(S_{2 m} \otimes_{\mathbb{C}} V\right)^{0}:=S_{2 m}^{+} \otimes V$ and $\left(S_{2 m} \otimes_{\mathbb{C}} V\right)^{1}:=S_{2 m}^{-} \otimes V$ (chirality grading). That this is a left $\mathbb{C} l_{2 m}$-supermodule structure on $S_{2 m} \otimes_{\mathbb{C}} V$ follows from the corresponding fact about $S_{2 m}$.

Consider the functor $\mathcal{F}$ from the category $\mathcal{C}$ of finite dimensional $\mathbb{C} l_{2 m}$-supermodules to itself, defined by $W \mapsto S_{2 m} \otimes_{\mathbb{C}} \operatorname{hom}_{\mathbb{C}_{2 m}}\left(S_{2 m}, W\right)$. Here $\operatorname{hom}_{\mathbb{C}_{2 m}}\left(S_{2 m}, W\right)$ is the ungraded $\mathbb{C}$-vector space of $\mathbb{C} l_{2 m}$-module morphisms of $S_{2 m} \rightarrow W$, and the tensor product $S_{2 m} \otimes_{\mathbb{C}} \operatorname{hom}_{\mathbb{C} l_{2 m}}\left(S_{2 m}, W\right)$ is made into a $\mathbb{C} l_{2 m}$-supermodule as in the last paragraph. There is the natural transformation of functors $\phi: \mathcal{F} \rightarrow I d_{\mathcal{C}}$ defined by

$$
\begin{aligned}
\phi_{W}: \mathcal{F}(W)=S_{2 m} \otimes_{\mathbb{C}} \operatorname{hom}\left(S_{2 m}, W\right) & \rightarrow W \\
s \otimes T & \mapsto T(s)
\end{aligned}
$$

Note that both functors $\mathcal{F}$ and $I d_{\mathcal{C}}$ are additive with respect to direct sums in $\mathcal{C}$. Also, on an irreducible $\mathbb{C l}_{2 m^{-}}$ supermodule $W$, we have $\operatorname{hom}_{\mathbb{C} l_{2 m}}\left(S_{2 m}, W\right) \simeq \mathbb{C} \phi_{W}$, where $\phi_{W}: S_{2 m} \rightarrow W$ is the unique $\mathbb{C}_{2 m}$-supermodule isomorphism between $S_{2 m} \rightarrow W$, since $\mathbb{C} l_{2 m}$ has a unique irreducible module $S_{2 m}$, and the only $\mathbb{C} l_{2 m}$-module maps between these finite dimensional irreducibles are $\left\{\lambda \phi_{W}\right\}_{\lambda \in \mathbb{C}}$ (these statements follow from the Schur lemma). Thus the natural transformation of functors $\mathcal{F} \rightarrow I d_{\mathcal{C}}$ is a natural equivalence on the full subcategory of irreducibles.

By Lemma 14.1.11 asserting complete reducibility of all $\mathbb{C} l_{2 m}$-modules, and the additivity of both functors $\mathcal{F}$ and $I d_{\mathcal{C}}$, it follows that $\phi_{W}$ is a natural equivalence of functors on all of $\mathcal{C}$. Also, the isomorphism $\phi_{W}$ : $\mathcal{F}(W) \rightarrow W$ is explicitly given by $s \otimes T \mapsto T(w)$, by the definition of $\phi_{W}$.

Example 14.4.7. For instance, we saw in Proposition 14.4.3 that as a left $\mathbb{C} l_{2 m}$-module, $\mathbb{C} l_{2 m} \simeq \operatorname{hom}_{\mathbb{C}}\left(S_{2 m}, S_{2 m}\right)$. This last module may be rewritten as $S_{2 m} \otimes_{\mathbb{C}} S_{2 m}^{*}$, so that the twisting space in this case is $S_{2 m}^{*}$. By the Proposition 14.4.6 above, there follows the curious fact that $\operatorname{hom}_{\mathbb{C l}_{2 m}}\left(S_{2 m}, \mathbb{C} l_{2 m}\right) \simeq S_{2 m}^{*}=\operatorname{hom}_{\mathbb{C}}\left(S_{2 m}, \mathbb{C}\right)$ as a $\mathbb{C}$-vector space.
14.5. Supertraces. A useful book-keeping device, which walks the bridge between an index and a trace, is the supertrace.

Definition 14.5.1. Let $W$ be a $\mathbb{C} l_{2 m}$-module. Recall the chirality element $\tau_{2 m} \in \mathbb{C} l_{2 m}$ defined by $i^{m} \omega_{2 m}$.
Give $W$ the $\mathbb{Z}_{2}$-grading $W^{ \pm}=( \pm 1)$ - eigenspace of $\tau_{2 m}$, which is the same grading as in its supermodule structure over $\mathbb{C} l_{2 m}$. We have seen that $\mathbb{C} l_{2 m}^{0} W^{ \pm} \subset W^{ \pm}$since $\tau_{2 m}(a v)=a \tau_{2 m}(v)$ for all $v \in W, a \in \mathbb{C} l_{2 m}^{0}$. Similarly, $\mathbb{C} l_{2 m}^{1} W^{ \pm} \subset W^{\mp}$, since $\tau_{2 m}(a v)=-a \tau_{2 m} v$ for all $v \in W, a \in \mathbb{C} l_{2 m}^{1}$. For $a \in \mathbb{C} l_{2 m}$, consider the endomorphism $a .(-)$ of $W$, and define the supertrace

$$
\begin{aligned}
\operatorname{str}_{W}(a)=\operatorname{tr}_{W}\left(\tau_{2 m} a\right) & =\operatorname{tr}_{W^{+}} a-\operatorname{tr}_{W^{-}} a \text { if } a \in \mathbb{C} l_{2 m}^{0} \\
& =0 \text { if } a \in \mathbb{C} l_{2 m}^{1}
\end{aligned}
$$

The formulas on the right for homogeneous elements in $\mathbb{C} l_{2 m}^{0}$ or $\mathbb{C} l_{2 m}^{1}$ follow from the fact that for $a \in \mathbb{C} l_{2 m}^{+}$, $\tau_{2 m} a$ acts as $a: W^{+} \rightarrow W^{+}$, and as $(-a): W^{-} \rightarrow W^{-}$, whereas for $a \in \mathbb{C} l_{2 m}^{-}, \tau_{2 m} a$ acts as $(-a): W^{+} \rightarrow W^{-}$ and $a: W^{-} \rightarrow W^{+}$, and is "off-diagonal".

Note that for any $\mathbb{C} l_{2 m}$-module $W$, the supertrace $\operatorname{str}_{W}$ gives a linear functional on $\mathbb{C} l_{2 m}$.

The following lemma characterises all supertraces on $\mathbb{C} l_{2 m}$. We define $T: \Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m}\right) \rightarrow \mathbb{C}$ be the projection into the top degree forms (as a multiple of $\left.\omega_{2 m}\right)$. Also recall the symbol map $\sigma: \mathbb{C} l_{2 m} \rightarrow \Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m}\right)$

Lemma 14.5.2. Let $W$ be any $\mathbb{C} l_{2 m}$-module. If $\psi: \mathbb{C} l_{2 m} \rightarrow \mathbb{C}$ is any linear functional which vanishes on all supercommutators in $\mathbb{C} l_{2 m}$, then $\psi=\lambda\left(\operatorname{str}_{W}\right)$ for some $\lambda \in \mathbb{C}$. Finally:

$$
\operatorname{str}_{W}(a)=(-i)^{m}\left(\operatorname{dim}_{\mathbb{C}} W\right)(T \circ \sigma(a))
$$

Proof: Recall that the grading on the module $W$ is given by the $( \pm 1)$-eigenspaces of $\tau_{m}$, viz. $W^{ \pm}$. Since $\tau_{2 m}$ commutes with $\mathbb{C} l_{2 m}^{0}$ and anticommutes with $\mathbb{C} l_{2 m}^{1}$, it follows that this grading makes $W$ a supermodule. The supercommutator of $a, b \in \mathbb{C l}_{2 m}$ was defined in Definition 13.1.7. Since $\operatorname{str}_{W}$ is linear, it suffices to show that $\operatorname{str}_{W}$ vanishes on supercommutators of homogeneous elements. If $a \in \mathbb{C} l_{2 m}^{0}$ (resp. $b \in \mathbb{C} l_{2 m}^{1}$ ), we can write it as a block-matrix in the $W^{+} \oplus W^{-}$decomposition as:

$$
a=\left(\begin{array}{cc}
a^{+} & 0 \\
0 & a^{-}
\end{array}\right) \quad \text { resp. } \quad b=\left(\begin{array}{cc}
0 & b^{-} \\
b^{+} & 0
\end{array}\right)
$$

Now if $a \in \mathbb{C} l_{2 m}^{0}$ and $b \in \mathbb{C} l_{2 m}^{1}$, then the supercommutator $[a, b]_{s} \in \mathbb{C} l_{2 m}^{1}$, and will have supertrace 0 , by the definitions above. Similarly for $a \in \mathbb{C} l_{2 m}^{1}$ and $b \in \mathbb{C} l_{2 m}^{0}$. So assume both $a, b \in \mathbb{C} l_{2 m}^{0}$, or both $a, b \in \mathbb{C} l_{2 m}^{1}$. Then, in the first case, $[a, b]_{s}=a b-b a$, which has the block matrix expression:

$$
[a, b]_{s}=\left(\begin{array}{cc}
{\left[a^{+}, b^{+}\right]} & 0 \\
0 & {\left[a^{-}, b^{-}\right]}
\end{array}\right)
$$

which implies $\operatorname{str}_{W}[a, b]_{s}=\operatorname{tr}_{W^{+}}\left(\left[a^{+}, b^{+}\right]\right)-\operatorname{tr}_{W^{-}}\left(\left[a^{-}, b^{-}\right]\right)=0$. In the second case, when both $a, b \in \mathbb{C} l_{2 m}^{1}$, then $[a, b]_{s}=a b+b a$, which has the matrix expression:

$$
[a, b]_{s}=\left(\begin{array}{cc}
a^{-} b^{+}+b^{-} a^{+} & 0 \\
0 & a^{+} b^{-}+b^{+} a^{-}
\end{array}\right)
$$

so that:

$$
\begin{aligned}
\operatorname{str}_{W}[a, b]_{s} & =\operatorname{tr}_{W^{+}}\left(a^{-} b^{+}+b^{-} a^{+}\right)-\operatorname{tr}_{W^{-}}\left(a^{+} b^{-}+b^{+} a^{-}\right) \\
& =\operatorname{tr}_{W^{+}}\left(a^{-} b^{+}\right)-\operatorname{tr}_{W^{-}}\left(b^{+} a^{-}\right)+\operatorname{tr}_{W^{+}}\left(b^{-} a^{+}\right)-\operatorname{tr}_{W^{-}}\left(a^{+} b^{-}\right)=0
\end{aligned}
$$

noting that both $W^{+}$and $W^{-}$are isomorphic as $\mathbb{C}$-vector spaces. (Left action by any $e_{i}$ interchanges $W^{+}$and $W^{-}$.)

Thus

$$
\operatorname{str}_{W}[a, b]_{s} \equiv 0 \quad \text { for all } \quad a, b \in \mathbb{C} l_{2 m}
$$

Define $C_{k}:=\sum_{i=0}^{k} c\left(\Lambda_{\mathbb{C}}^{k}\left(\mathbb{R}^{2 m}\right)\right)$. That is, $C_{k}$ is the subspace of $\mathbb{C} l_{2 m}$ spanned by all basis elements $e_{I}$ with $|I| \leq k$. We now claim that $C_{2 m-1} \subset\left[\mathbb{C}_{2 m}, \mathbb{C} l_{2 m}\right]_{s}$. For if $e_{I}$ is any basis element with $|I| \leq 2 m-1$, then there exists a $j$ such that $j \notin I$. Letting $|I|=k$, we compute:

$$
\left[e_{j}, e_{j} e_{I}\right]_{s}=e_{j}^{2} e_{I}-(-1)^{1 .(k+1)} e_{j} e_{I} e_{j}=-e_{I}-(-1)^{2 k+1} e_{I} e_{j}^{2}=-2 e_{I}
$$

which shows that every $e_{I}$ with $|I| \leq 2 m-1$ is a supercommutator, and the claim follows. Hence the supertrace satisfies $\operatorname{str}\left(C_{2 m-1}\right) \equiv 0$. Since the quotient $\mathbb{C l}_{2 m} / C_{2 m-1} \simeq \mathbb{C}$ is one dimensional, it follows that str ${ }_{W}$ descends to this 1-dimensional quotient.

Since $\operatorname{str}_{W} \tau_{2 m}=\operatorname{tr}_{W}\left(\tau_{2 m}^{2}\right)=\operatorname{dim} W \neq 0$, it follows that the supertrace $\operatorname{str}_{W}$ gives an isomorphism of $\mathbb{C} l_{2 m} / C_{2 m-1} \rightarrow \mathbb{C}$. It also follows, since $\operatorname{str}_{W}$ is not the zero map, that the dimension of $\mathbb{C} l_{2 m} /\left[\mathbb{C} l_{2 m}, \mathbb{C} l_{2 m}\right]_{s}$ cannot be zero, and since it is $\leq \operatorname{dim}_{\mathbb{C}} \mathbb{C} l_{2 m} / C_{2 m-1}=1$, must be 1 . Thus $\left[\mathbb{C} l_{2 m}, \mathbb{C} l_{2 m}\right]=C_{2 m-1}$.

The second assertion of the statement is now clear, since any linear functional annihilating all supercommutators descends to the 1 -dimensional space $\mathbb{C} l_{2 m} /\left[\mathbb{C} l_{2 m}, \mathbb{C} l_{2 m}\right]_{s}$.

Now note that $T \circ \sigma: \mathbb{C l}_{2 m} \rightarrow \mathbb{C}$ is a linear functional on $\mathbb{C l}_{2 m}$, and annihilates $C_{2 m-1}$, since ker $T=$ $\sum_{i<2 m-1} \Lambda^{i}=\sigma\left(C_{2 m-1}\right)$. Hence it annihilates $\left[\mathbb{C l}_{2 m}, \mathbb{C} l_{2 m}\right]_{s}$, and by the last para $T \circ \sigma$ and $\operatorname{str}_{W}$ are scalar multiples of each other on $\mathbb{C} l_{2 m}$. Indeed, by evaluating both on $\tau_{2 m}$, we saw that

$$
\operatorname{str}_{W}\left(\tau_{2 m}\right)=\operatorname{tr}_{W}\left(\tau_{2 m}^{2}\right)=\operatorname{dim} W
$$

whereas $T \circ \sigma\left(\tau_{2 m}\right)=i^{m}$. This implies that $\operatorname{str}_{W}=(\operatorname{dim} W)(-i)^{m}(T \circ \sigma)$. The lemma follows.

Corollary 14.5.3. The proof above showed that $\mathbb{C}_{2 m-1}=\left[\mathbb{C} l_{2 m}, \mathbb{C} l_{2 m}\right]_{s}$.

## 15. Clifford Bundles and Dirac operators

From now on, let $M$ be a compact oriented Riemannian manifold of dimension $2 m . P_{S O} \rightarrow M$ will denote its oriented orthonormal frame bundle, with structure group $S O(2 m)$.

### 15.1. Clifford bundles, Clifford modules and the Spinor bundle.

Definition 15.1.1. The Clifford bundle of $M$ is the complex vector bundle $\pi: \mathbb{C l}(M) \rightarrow M$ whose fibre at $x \in M$ is the complex Clifford algebra $\mathbb{C l}\left(T_{x}^{*} M\right)$, where $T_{x}^{*}(M)$ is given the real positive definite inner product $\langle-,-\rangle_{x}$ from the Riemannian metric induced on the cotangent bundle. It can be viewed as the associated vector bundle:

$$
P_{S O} \times_{S O(2 m)} \mathbb{C l}\left(\mathbb{R}^{2 m}\right)
$$

where $S O(2 m)$ has the obvious action on $\mathbb{C l}\left(\mathbb{R}^{2 m *}\right)$ (defined by $e_{i}^{*} \mapsto f_{i}^{*}:=g \cdot e_{i}^{*}$ for $g \in S O(2 m)$, where $e_{i}^{*}$ is an orthonormal basis for $\left.\mathbb{R}^{2 m *}\right)$.

Since the vector bundle $\Lambda_{\mathbb{C}}^{*}\left(T^{*} M\right) \rightarrow M$ is the associated bundle:

$$
P_{S O} \times_{S O(2 m)} \Lambda^{*}\left(\mathbb{R}^{2 m *}\right) \rightarrow M
$$

and the symbol map $\sigma$ and quantisation map $c$ are $S O(2 m)$ - equivariant, we get global vector bundle maps:

$$
\sigma: \mathbb{C} l(M) \rightarrow \Lambda_{\mathbb{C}}^{*}\left(T^{*} M\right)
$$

called the symbol map of $M$, and

$$
c: \Lambda_{\mathbb{C}}^{*}\left(T^{*} M\right) \rightarrow \mathbb{C l}(M)
$$

called the quantisation map of $M$.

Remark 15.1.2. We note that the action of $S O(2 n)$ on $\mathbb{C} l_{2 m}$ is the descended action from the $\operatorname{Spin}(2 m)$ action on $\mathbb{C} l_{2 m}$ by conjugation. Hence the fibre of $\mathbb{C l}(M)$ is the module $\Delta_{2 m} \otimes \Delta_{2 m}$, by (i) of Proposition 14.4.4.

In the light of the Proposition 14.4.6, it is desirable to have a bundle $\Delta \rightarrow M$ on $M$ with the fibre $\Delta_{2 m}$ (or what is the same thing, $S_{2 m}$ ), so that any bundle of Clifford modules on $M$ (such as $\left.\Lambda_{\mathbb{C}}^{*}, \Lambda_{\mathbb{C}}^{ \pm}, \Lambda_{\mathbb{C}}^{e v}, \Lambda_{\mathbb{C}}^{o}\right)$ can be written as a tensor product $\Delta \otimes_{\mathbb{C}} V$, where $V$ is a twisting bundle.

Unfortunately, this cannot be done unless we assume a $\operatorname{Spin}(2 m)$ structure on $M$, because the representation $\Delta_{2 m}$ of $\operatorname{Spin}(2 m)$ (or for that matter the representation $S_{2 m}$ of $\mathbb{C l}_{2 m}$ ) does not descend to a representation of $S O(2 m)$. Hence there is no way to start with the principal bundle $P_{S O}$ and get an associated bundle with fibre $\Delta_{2 m}$ or $S_{2 m}$.

Definition 15.1.3. Let $M$ be a Riemannian oriented manifold of dimension $2 m$, and assume it has a spin structure. Let $P_{\text {spin }} \rightarrow M$ denote the principal $\operatorname{Spin}(2 m)$-bundle over $M$, (see Definition 13.3.1). Then consider the associated complex vector bundle of rank $2^{m}$ :

$$
P_{\text {spin }} \times{ }_{\operatorname{Spin}(2 m)} \Delta_{2 m} \rightarrow M
$$

where $\Delta_{2 m}=\Delta_{2 m}^{+} \oplus \Delta_{2 m}^{-}$is the irreducible $\mathbb{C} l_{2 m}$ supermodule $S_{2 m}$ with restricted action of $\operatorname{Spin}(2 m)$, and $\Delta_{2 m}^{ \pm}$are the two irreducible half-spin representations (see Proposition 14.2.1). This is called the spin bundle over $M$, and denoted $\mathcal{S}(M) \rightarrow M$. It is the direct sum of the half spin bundles $\mathcal{S}^{ \pm}(M) \rightarrow M$, which are analogously defined as the associated rank $2^{m-1}$ complex vector bundles:

$$
P_{\text {spin }} \times \operatorname{Spin}(2 m) \Delta_{2 m}^{ \pm} \rightarrow M
$$

respectively.

Proposition 15.1.4. We have the following facts about the spin bundles:
(i): There exists a bundle map $c: \mathbb{C} l(M) \otimes_{\mathbb{C}} \mathcal{S}(M) \rightarrow \mathcal{S}(M)$ called Clifford multiplication whose restriction to fibres is the natural map $\mathbb{C l}\left(T^{*} M_{x}\right) \otimes S_{2 m, x} \rightarrow S_{2 m, x}$ defining the $\mathbb{C l}\left(T^{*} M_{x}\right)$-module structure on $S_{2 m, x}$. For notational simplicity, we denote $c(a, v)$ as a.v. Finally $\mathbb{C} l^{0}(M) \cdot \mathcal{S}^{ \pm}(M) \mapsto \mathcal{S}^{ \pm}(M)$ and $\mathbb{C} l^{1}(M) \cdot \mathcal{S}^{ \pm}(M) \mapsto \mathcal{S}^{\mp}(M)$.
(ii): The spin bundle $\mathcal{S}(M) \rightarrow M$ is a hermitian vector bundle with a natural hermitian metric (,-- ). The direct sum decomposition $\mathcal{S}(M)=\mathcal{S}^{+}(M) \oplus \mathcal{S}^{-}(M)$ is orthogonal with respect to $(-,-)$.
(iii): The Clifford action defined in (i) above is self-adjoint in the sense of Definition 14.3.1. In particular, we have:

$$
\left(\alpha_{x} \cdot v, w\right)=-\left(v, \alpha_{x}^{*} w\right) \text { for } \alpha_{x} \in T_{x}^{*}(M), \quad v, w \in \mathcal{S}(M)_{x}
$$

Proof: First note that if we let $\operatorname{Spin}(2 m)$ act by conjugation on $\mathbb{C l}_{2 m}$ (call this representation $\tau$ ), then $\rho(-1)=\rho(+1)$, and so $\tau=\mu \circ \rho$ where $\rho: \operatorname{Spin}(2 m) \rightarrow S O(2 m)$ is the double covering homomorphism, and $\mu$ is the representation of $S O(2 m)$ on $\mathbb{C} l_{2 m}$ described in Definition 15.1.1. Thus

$$
P_{\text {spin }} \times_{\tau} \mathbb{C} l_{2 m}=P_{\text {spin }} \times_{\rho} S O(2 m) \times_{\mu} \mathbb{C} l_{2 m}=P_{S O} \times_{\mu} \mathbb{C} l_{2 m}=\mathbb{C l}(M)
$$

Now, there is the map defining Clifford module action on $S_{2 m}$ :

$$
\mathbb{C} l_{2 m} \otimes_{\mathbb{C}} S_{2 m} \rightarrow S_{2 m}
$$

which is $\operatorname{Spin}(2 m)$-equivariant (since $\left.g(x \otimes v)=g x g^{*} \otimes g v \mapsto g x g^{*} . g v=g . v\right)$. Hence there is a natural map of vector bundles:

$$
\left(P_{\text {spin }} \times_{\tau} \mathbb{C} l_{2 m}\right) \otimes\left(P_{\text {spin }} \times{ }_{\operatorname{Spin}(2 m)} S_{2 m}\right) \rightarrow P_{\text {spin }} \times{ }_{\operatorname{Spin}(2 m)} S_{2 m}
$$

i.e. a bundle map $\mathbb{C l}(M) \otimes \mathcal{S}(M) \rightarrow \mathcal{S}(M)$. It clearly restricts on fibres to what we claimed, by its definition. Also the last statement of (i) follows since $S_{2 m}$ is a $\mathbb{C} l_{2 m^{-}}$supermodule. This proves (i).

For (ii), construct the metric on each fibre by taking the hermitian metric $(-,-)$ constructed on $S_{2 m}$ in the Example 14.3.3. This makes the representation of $\operatorname{Spin}(2 m)$ on $S_{2 m}=\Delta_{2 m}$ unitary, by Proposition
14.3.5. Hence the associated bundle $P_{\text {spin }} \times{ }_{\operatorname{Spin}(2 m)} S_{2 m}$ is a hermitian vector bundle. Also note that since $\omega_{2 m} \in \operatorname{Spin}(2 m)$, we have $\omega_{2 m} v=i^{-m} \tau_{2 m} v=i^{-m} v$ for $v \in S_{2 m}^{+}$, and $\omega_{2 m} w=i^{-m} \tau_{2 m} w=-i^{-m} w$ for $w \in S_{2 m}^{-}$. Thus, by the unitarity of $\operatorname{Spin}(2 m)$ action on $S_{2 m}$, we have:

$$
(v, w)=\left(\omega_{2 m} v, \omega_{2 m} w\right)=\left(i^{-m} v,-i^{-m} w\right)=-i^{-m}(-i)^{-m}(v, w)=-(v, w)
$$

which implies $(v, w)=0$ for $v \in S_{2 m}^{+}$and $w \in S_{2 m}^{-}$. Then (ii) follows, because the representation of $\operatorname{Spin}(2 m)$ on $S_{2 m}$ is unitary.
(iii) is a direct consequence of the Example 14.3.3, which showed that $S_{2 m}$ is a self-adjoint $\mathbb{C} l_{2 m}$ module with respect to $(-,-)$.

Now we are ready to abstract all the facts proved above into a definition.

Definition 15.1.5. Let $\mathcal{E} \rightarrow M$ be complex vector bundle over an oriented Riemannian manifold of dimension $2 m$, with a hermitian metric $(-,-)$. Say that this bundle is a Clifford module over $M$ if:
(i): There is a $(-,-)$-orthogonal decomposition $\mathcal{E}=\mathcal{E}^{+} \oplus \mathcal{E}^{-}$into two complex sub-bundles.
(ii): There is a vector bundle Clifford multiplication or Clifford action map:

$$
c: \mathbb{C l}(M) \otimes \mathcal{E} \rightarrow \mathcal{E}
$$

such that for each point $x$, the restriction $c_{x}: \mathbb{C l}\left(T_{x}^{*} M\right) \otimes \mathcal{E}_{x} \rightarrow \mathcal{E}_{x}$ gives $\mathcal{E}_{x}$ the structure of a $\mathbb{C l}\left(T_{x}^{*} M\right)$ supermodule, with graded pieces $\mathcal{E}_{x}^{ \pm}$. (In particular, the ranks of $\mathcal{E}^{+}$and $\mathcal{E}^{-}$are equal, and $\mathcal{E}$ is a bundle of even rank).
(iii): The action of the Clifford algebra $\mathbb{C l}\left(T_{x}^{*} M\right)$ on $\mathcal{E}_{x}$ is self-adjoint with respect to the hermitian inner product $(-,-)$ on $\mathcal{E}_{x}$.

Example 15.1.6. Clearly, for $M$ a spin manifold of dimension $2 m$, the spin bundle $\mathcal{S}(M) \rightarrow M$ is a Clifford module over $M$, by Proposition 15.1.4

Example 15.1.7. Let $M$ be an oriented Riemannian manifold, not ncessarily spin. The complexified exterior algebra bundle $\Lambda_{\mathbb{C}}^{*}\left(T^{*}\right)(M) \rightarrow M$ is a Clifford module over $M$. For, we define the Clifford action fibre by fibre as the action which extends the action:

$$
\begin{aligned}
T_{x}^{*}(M) \otimes \Lambda_{\mathbb{C}}^{*}\left(T^{*} M\right) & \rightarrow \Lambda_{\mathbb{C}}^{*}\left(T^{*} M\right) \\
\alpha \otimes \phi & \mapsto \alpha \wedge \phi-\alpha\lrcorner \phi
\end{aligned}
$$

That this extends to an action of $\mathbb{C l}\left(T_{x}^{*} M\right)$ is the content of Proposition 14.1.5. One makes the natural hermitian extension of the Riemannian inner product $\langle-,-\rangle$ on the real exterior algebra $\Lambda^{*}\left(T^{*} M\right)$, setting $(\phi \otimes \lambda, \psi \otimes \mu)=\langle\phi, \psi\rangle \lambda \bar{\mu}$, and appeals to the last part of Proposition 14.1.5 to show that the Clifford action is self-adjoint.

As expected, there are two possible gradings $\mathcal{E}^{ \pm}$available on this bundle $\mathcal{E}=\Lambda_{\mathbb{C}}^{*}\left(T^{*} M\right)$. There is the global volume element $\omega_{M} \in C^{\infty}(M, \mathbb{C l}(M))$, given in a coordinate chart $U$ of $x$ by $\omega_{M, x}:=e_{1}(x) . e_{2}(x) \ldots . e_{2 m}(x)$ where $\left\{e_{i}\right\}$ is a local orthonormal frame for $T^{*} M$ on $U$ (this definition is independent of coordinate charts, indeed $\omega_{M}$ corresponds to the Riemannian volume form on $M$ under the symbol isomorphism). Similarly, there is the global chirality element $\tau_{M}:=i^{m} \omega_{M}$.

The first grading then is the even-odd grading, in which the graded pieces $\Lambda_{\mathbb{C}}^{e v}$ and $\Lambda_{\mathbb{C}}^{o}$ come from the pointwise action of conjugation by $\omega_{M, x} \in \operatorname{Spin}(2 m) \subset \mathbb{C l}(M)_{x}$ (see the proof of (iii) in Proposition 14.4.4. Another example comes from taking the graded pieces $\Lambda_{\mathbb{C}}^{+}$and $\Lambda_{\mathbb{C}}^{-}$corresponding to the $\pm 1$ eigenspaces of $i^{m+k(k-1)} *$ (or pointwise left action by $\tau_{M, x}$, if we identify $\mathbb{C} l_{2 m}$ with $\Lambda_{\mathbb{C}}^{*}$ ).

Example 15.1.8. The Clifford bundle of $M, \operatorname{viz} \mathbb{C l}(M) \rightarrow M$ is a Clifford module, with Clifford action being left multiplication.

Note also that the Clifford bundle $\mathbb{C l}(M)$ has two possible $\mathbb{Z}_{2}$-gradings as a Clifford bundle (see Example 14.1.17, both of which equip the typical fibre $\mathbb{C l}(M)_{x}$ with the structure of a $\mathbb{C l}\left(T_{x}^{*}\right)$-supermodule. The obvious is the chirality $\mathbb{Z}_{2}$-grading $\mathbb{C l}(M)^{ \pm}$which corresponds under the symbol isomorphism to the decomposition $\Lambda_{\mathbb{C}}^{ \pm}$fo the exterior algebra bundle (see previous Example). This grading coincides with the chirality coming from left multiplication by $\tau_{M}$. On the other hand, there is the parity $\mathbb{Z}_{2}$-grading $\mathbb{C l}(M)=\mathbb{C} l^{0}(M) \oplus \mathbb{C} l^{1}(M)$ (corresponding to $\Lambda_{\mathbb{C}}^{e v}$ and $\Lambda_{\mathbb{C}}^{o}$ under symbol isomorphism), which comes from conjugation by $\omega_{M}$.

Here is the reason for introducing the spin bundle $\mathcal{S}(M) \rightarrow M$

Proposition 15.1.9. Let $M$ be an oriented spin manifold of dimension $2 m$, and let $\mathcal{S}(M) \rightarrow M$ be the spin bundle on it. Then, for any Clifford module $\mathcal{W} \rightarrow M$ on $M$, there is a hermitian complex twisting vector bundle $\mathcal{V} \rightarrow M$ such that $\mathcal{W} \simeq \mathcal{S}(M) \otimes_{\mathbb{C}} \mathcal{V}$. Note that this isomorphism is an isomorphism of Clifford modules on $M$, i.e. the graded structure and hermitian structure is also preserved.

Proof: Define the bundle $\mathcal{V}=\operatorname{hom}_{\mathbb{C} l(M)}(\mathcal{S}(M), \mathcal{W})$ and appeal to Proposition 14.4.6. That the Clifford action matches follows from that proposition, because the map of vector spaces:

$$
\begin{aligned}
\phi_{W}: S_{2 m} \otimes_{\mathbb{C}} \operatorname{hom}_{\mathbb{C} l_{2 m}}\left(S_{2 m}, W\right) & \rightarrow W \\
s \otimes T & \mapsto T(s)
\end{aligned}
$$

being an isomorphism of $\mathbb{C} l_{2 m}$-modules, is in particular $\operatorname{Spin}(2 m)$ equivariant. Thus it globalises to a vector bundle isomorphism $\phi_{\mathcal{W}}$.

Recall the hermitian metric $(-,-)_{\mathcal{S}}$ on $\mathcal{S}_{2 m}$, which was defined in Proposition 15.1.4. We just need to put a bundle metric $(-,-)_{\mathcal{V}}$ on $\mathcal{V}$ so that when the tensor product $\mathcal{S}(M) \otimes_{\mathbb{C}} \mathcal{V}$ is equipped with the tensor product hermitian metric $(-,-)_{\mathcal{S}} \otimes(-,-)_{\mathcal{V}}$, the Clifford module isomorphism $\phi_{\mathcal{W}}$ is an isometry with the given hermitian metric $(-,-) \mathcal{W}$ on $\mathcal{W}$.

We note that for a vector space $V$ with any hermitian inner product $(-,-)_{V}$ on it, the tensor product hermitian inner product on $S_{2 m} \otimes V$, defined by:

$$
(s \otimes S, t \otimes T)_{S_{2 m} \otimes V}:=(s, t)_{S}(S, T)_{V}
$$

automatically obeys self-adjointness with respect to Clifford action, because the Clifford action is self-adjoint with respect to the natural metric $(-,-)_{S}$ on $S_{2 m}$, by the Example 14.3.3.

We note that if $W$ is an irreducible $\mathbb{C l}_{2 m}$-module with a positive definite hermitian inner-product $(-,-)_{W}$ with respect to which the $\mathbb{C} l_{2 m}$ action is self-adjoint, then we claim that this self-adjointness property determines $(-,-)_{W}$ uniquely upto a non-zero complex scalar. For if $(-,-)^{\prime}$ is another hermitian inner-product with respect to which the $\mathbb{C} l_{2 m}$ action is self-adjoint, then we have a $\mathbb{C}$-linear isomorphism $A: W \rightarrow W$ such that:

$$
\left(w_{1}, w_{2}\right)^{\prime}=\left(A w_{1}, w_{2}\right)_{W} \quad \text { for all } w_{1}, w_{2} \in W
$$

Also for $c \in \mathbb{C} l_{2 m}$, we have

$$
\left(c A w_{1}, w_{2}\right)_{W}=(-1)^{\operatorname{deg} c}\left(A w_{1}, c^{*} w_{2}\right)=(-1)^{\operatorname{deg} c}\left(w_{1}, c^{*} w_{2}\right)^{\prime}=\left(c w_{1}, w_{2}\right)^{\prime}=\left(A\left(c w_{1}\right), w_{2}\right)_{W}
$$

Thus $A: W \rightarrow W$ is a map of $\mathbb{C} l_{2 m}$-modules, and by irreducibility of $W$, must be a scalar (Schur Lemma), and since $A$ is an isomorphism, the scalar must be non-zero.

If $W$ is irreducible, and $\phi_{W}: S_{2 m} \rightarrow W$ is an isomorphism of $\mathbb{C} l_{2 m}$-modules, it follows that there is a scalar $\alpha_{W} \neq 0$

$$
\alpha_{W}\left(w_{1}, w_{2}\right)_{W}=\left(\phi_{W}^{-1}\left(w_{1}\right), \phi_{W}^{-1}\left(w_{2}\right)\right)_{S} \quad \text { for all } \quad w_{1}, w_{2} \in W
$$

or equivalently

$$
(s, t)_{S}=\alpha_{W}\left(\phi_{W}(s), \phi_{W}(t)\right)_{W} \quad \text { for all } s, t \in S_{2 m}
$$

Note that $\alpha_{W}$ gets determined by the equation:

$$
\operatorname{tr} \phi_{W}^{*} \phi_{W}=\sum_{i=1}^{2^{m}}\left(\phi_{W} e_{i}, \phi_{W} e_{i}\right)_{W}=\alpha_{W}^{-1} \sum_{i=1}^{2^{m}}\left(e_{i}, e_{i}\right)_{S}=\alpha_{W}^{-1} 2^{m}
$$

where $\left\{e_{i}\right\}$ is an orthonormal basis of $S_{2 m}$ with respect to $(-,-)_{S}$.

Now let $S=\lambda \phi_{W}$ and $T=\mu \phi_{W} \in V=\operatorname{hom}_{\mathbb{C l}_{2 m}}\left(S_{2 m}, W\right)=\mathbb{C} \phi_{W}$. Define the hermitian inner product on $V$ given by:

$$
(T, S)_{V}:=\alpha_{W}^{-1} \lambda \bar{\mu}=2^{-m} \operatorname{tr}\left(\phi_{W}^{*} \phi_{W}\right) \lambda \bar{\mu}=2^{-m} \operatorname{tr} S^{*} T
$$

Then we have:

$$
\begin{aligned}
(s \otimes S, t \otimes T)_{S_{2 m} \otimes V} & =(s, t)_{S}(S, T)_{V}=\alpha_{W}\left(\phi_{W}(s), \phi_{W}(t)\right)_{W} \alpha_{W}^{-1} \lambda \bar{\mu} \\
& =\left(\lambda \phi_{W}(s), \mu \phi_{W}(t)\right)_{W}=(S(s), T(t))_{W}
\end{aligned}
$$

which shows that the isomorphism $S_{2 m} \otimes V \rightarrow W$ given by $s \otimes S \rightarrow S(s)$ is an isometry.

For a general $W$, break it into irreducibles $W_{i}$, and note that $V=\operatorname{hom}_{\mathbb{C} l_{2 m}}\left(S_{2 m}, W\right)=\oplus_{i} V_{i}$, and equip each summand $V_{i}:=\operatorname{hom}_{\mathbb{C l} l_{2 m}}\left(S_{2 m}, W_{i}\right)$ with the hermitian inner product above. To globalise to Clifford modules $\mathcal{W}$ is obvious, since the inner product $(T, S)=2^{-m} \operatorname{tr} S^{*} T$ (the normalised Hilbert-Schmidt norm) is invariantly defined, independent of frames.

Example 15.1.10. Let $M$ be a spin manifold, with the spin bundle $\mathcal{S}(M) \rightarrow M$, and its half-spin sub-bundles $\mathcal{S}^{ \pm}(M) \rightarrow M$. Then, as a direct consequence of the module identities of Proposition 14.4.4, and the fact that the isomorphisms there are isomorphisms of $\operatorname{Spin}(2 m)$-modules (i.e. $\operatorname{Spin}(2 m)$-equivariant isomorphisms), there are the following bundle identities of associated vector bundles, indeed, of Clifford modules:
(i): $\mathbb{C} l(M) \simeq \Lambda_{\mathbb{C}}^{*}\left(T^{*} M\right) \simeq \mathcal{S}(M) \otimes \mathcal{S}(M)$.
(ii): $\Lambda_{\mathbb{C}}^{ \pm}(M) \simeq \mathcal{S}^{ \pm}(M) \otimes \mathcal{S}(M)$.
(iii): $\Lambda_{\mathbb{C}}^{e v}(M) \simeq(-1)^{m} \mathcal{S}^{+}(M) \otimes \mathcal{S}^{+}(M) \oplus(-1)^{m} \mathcal{S}^{-}(M) \otimes \mathcal{S}^{-}(M)$ and $\Lambda_{\mathbb{C}}^{o}(M) \simeq(-1)^{m} \mathcal{S}^{+}(M) \otimes \mathcal{S}^{-}(M) \oplus(-1)^{m} \mathcal{S}^{-}(M) \otimes \mathcal{S}^{+}(M)$.

## Remark 15.1.11.

(i): Note that the identity (i) above says that the spin bundle $\mathcal{S}(M)$ is in some sense the "square-root" bundle of the exterior algebra (or Clifford) bundle on $M$, if $M$ is a spin manifold.
(ii): The chirality grading on $\mathcal{S}(M) \otimes \mathcal{S}(M)$ comes from left action of $\tau_{M}$ on the first factor, and predictably leads to the grading $\mathbb{C l}(M)^{ \pm}$discussed in (ii) of Example 15.1.8. The other grading, which also restricts fibrewise to a $\mathbb{C l}\left(T_{x}^{*} M\right)$-supermodule structure corresponds to the parity or $\mathbb{C} l^{0}, \mathbb{C} l^{1}(M)$ grading (coming from conjugation by $\omega_{M}$, see Example 15.1 .8 above), and has no simple relation to the chirality grading, as is evidenced by the complicated formula in (iii) of Example 15.1.10 above.

### 15.2. Clifford connections.

Definition 15.2.1 (Levi-Civita connection). If $M$ is an oriented Riemannian manifold of dimension $2 m$, there is an $S O(2 m)$-connection on the principal $S O(2 m)$ bundle $P_{S O} \rightarrow M$. This means that:
(i): There is a $\mathfrak{s o}(2 m)$-valued 1-form $\left[\omega_{i j}\right] \in \Lambda^{1}(P) \otimes \mathfrak{s o}(2 m):=C^{\infty}\left(\Lambda^{1}\left(T^{*} P\right) \otimes \mathfrak{s o}(2 m)\right)$. This merely means that $\left[\omega_{i j}\right]$ is a $2 m \times 2 m$ skew-symmetric matrix of 1-forms $\omega_{i j}$ on $P$.
(ii): If we think of $P_{S O}$ has having right $S O(2 m)$-action, then the matrix of 1-forms $\omega:=\left[\omega_{i j}\right]$ must satisfy:

$$
R_{g}^{*} \omega=g \omega g^{-1}=(\operatorname{Ad} g) \omega \quad \text { for all } g \in S O(2 m)
$$

(iii): [Torsion-free condition] Let $\sigma: U \rightarrow P_{S O \mid U}$ be a smooth local section of $P$ over an open set $U \subset M$. For $x \in U, \sigma(x)$ is an orthonormal frame at $x$. So $\sigma$ is a local orthonormal frame field over $U$, and can be regarded as a (2m)-row vector of 1-forms $\sigma=\left(\sigma_{1}, . ., \sigma_{2 m}\right)$, with $\sigma_{i, x}(X)=X_{i}$ where $X_{i}$ the $i$-th component of $X \in T_{x} M$ in the frame $\sigma(x)$. Then we require the following identity of 2-forms on $U$ :

$$
d \sigma_{i}+\sum_{j}\left(\sigma^{*} \omega_{i j}\right) \wedge \sigma_{j}=0
$$

for each smooth section $\sigma: U \rightarrow P_{S O \mid U}$ over $U$. This connection is called the Levi-Civita connection on $P_{S O}$.

Definition 15.2.2 (Covariant differentiation in associated bundles). Let $\rho: S O(2 m) \rightarrow G L_{\mathbb{C}}(V)$ be a complex representation of $S O(2 m)$ on a complex vector space $V$. Let $\mathcal{V}:=P \times{ }_{\rho} V$ be the associated complex vector bundle. For a connection on $P_{S O} \rightarrow M$ as above, one gets a covariant differentiation operator for every open set $U \subset M$ :

$$
\nabla: C^{\infty}(U, \mathcal{V}) \rightarrow C^{\infty}\left(U, T^{*} M \otimes \mathcal{V}\right)
$$

which is a $\mathbb{C}$-linear map satisfying the Leibnitz Rule:

$$
\nabla(f s)=f \nabla s+d f \otimes s \quad \text { for all } f \in C^{\infty}(U), s \in C^{\infty}(\mathcal{V})
$$

To define the above covariant differentiation, it is enough to do it on trivialising neighbourhoods $U \subset M$ for $P_{S O}$ (and of course check that the definition is independent of trivialisations). If we fix a basis $\left\{e_{i}\right\}$ of the vector space $V$, then for each smooth local section $\sigma: U \rightarrow P_{S O}$ over a trivialising neighbourhood $U \subset M$ for $P_{S O}$, we get a local framing $\widetilde{e}_{i}:=\rho(\sigma) e_{i}$ of the vector bundle $\mathcal{V}_{\mid U}$. In view of the Leibnitz Rule, it is enough to define $\nabla \widetilde{e}_{j}$, and these are defined by:

$$
\nabla \widetilde{e}_{j}:=\sum_{i} \dot{\rho}\left(\sigma^{*} \omega\right)_{i j} \otimes \widetilde{e}_{i}
$$

which is often abbreviated to $\nabla \widetilde{e}_{j}:=\sum_{i} \omega_{i j} \otimes \widetilde{e}_{i}$, where $\omega:=\dot{\rho}(\omega)$ is a $\mathfrak{g l}(V)$-valued 1-form on $U$, called the Cartan connection 1-form. If $X \in T_{x}^{*}(M)$ is a (real) tangent vector at $x$, and $s$ a section of $\mathcal{V}$, we can define;

$$
\left.\nabla_{X} s:=X\right\lrcorner \nabla s
$$

In a local trivialising neighbourhood we have: $\nabla_{X} \widetilde{e}_{j}=\sum_{i} \omega_{i j}(X) \widetilde{e}_{i}$. We can also define $\nabla_{X}$ for $X$ a real tangent vector field on $M$.

We finally note that if $e_{x} \mapsto g_{\alpha \beta}(x) e_{x}$ is a coordinate change on $U_{\alpha} \cap U_{\beta}$ for the principal bundle $P_{S O}$, where $g_{\alpha \beta}: U_{\alpha} \cap U_{\beta} \rightarrow S O(2 m)$, then if $\omega^{\alpha}$ and $\omega^{\beta}$ are the matrix-valued Cartan 1- forms on $U_{\alpha}$ and $U_{\beta}$ respectively, then there is the transformation formula:

$$
\omega^{\alpha}=\operatorname{Ad}\left(g_{\alpha \beta}\right) \omega^{\beta}+d g_{\alpha \beta} \cdot g_{\alpha \beta}^{-1}
$$

where the product in the second term on the right is a matrix product.

Lemma 15.2.3. If the representation $\rho: S O(2 m) \rightarrow V$ is unitary with respect to a hermitian inner product $(-,-)_{V}$ on $V$, the associated bundle $\mathcal{V}:=P_{S O} \times{ }_{\rho} \mathcal{V}$ is a hermitian vector bundle, with hermitian inner product denoted $(-,-)$. The covariant derivative on $\mathcal{V}$ associated to the Levi-Civita connection on $P_{S O}$ is a unitary connection. It satisfies:

$$
X(s, t)=\left(\nabla_{X} s, t\right)+\left(s, \nabla_{X} t\right) \quad \text { for } s, t \in C^{\infty}(M, \mathcal{V}), \quad X \in C^{\infty}(M, T M)
$$

Proof: As noted in Example 15.1.7 above, the hermitian inner product $(-,-)$ on $\mathcal{V}$ is defined as follows. Let $[e, v],\left[e^{\prime}, w\right] \in \mathcal{V}_{x}$, with $e=e^{\prime} g$ and $g \in S O(2 m)$. Then define:

$$
\left([e, v],\left[e^{\prime}, w\right]\right)=([e, v],[e, \rho(g) w]):=(v, \rho(g) w)_{V}
$$

To check this is well-defined, we choose a different representative $\left[e h^{-1}, \rho(h) v\right]$ for $[e, v]$, with $h \in S O(2 m)$, then $e^{\prime}=\left(e h^{-1}\right) \cdot(h g)$, and so

$$
\left(\left[e h^{-1}, \rho(h) v\right],\left[e^{\prime}, w\right]\right)=(\rho(h) v, \rho(h g) w)_{V}=(\rho(h) v, \rho(h) \rho(g) w)_{V}=(v, \rho(g) w)_{V}=\left([e, v],\left[e^{\prime}, w\right]\right)
$$

since $\rho(h)$ is a unitary automorphism of $V$. This shows the definition of $(-,-)$ is independent or representatives in the first slot. Similarly for the second slot.

To check the second fact, note that if we start with the $(-,-)_{V}$-orthonormal frame $\left\{e_{j}\right\}$ for $V$, and $s \rightarrow \sigma(x) .1$ a local section on on some trivialising neighbourhood $U$ for $P_{S O}$, then the frame $\left\{\widetilde{e}_{j, x}\right\}=\left\{\left[1, \rho(\sigma(x)) e_{j}\right]\right\}$ is orthonormal in $\mathcal{V}_{x}$ for all $x \in U$ (since $\sigma(x) \in S O(2 m)$ and hence $\left.\rho(\sigma(x)) \in U(V)\right)$. Hence, for a smooth vector field $X \in C^{\infty}(U)$, we have:

$$
\begin{aligned}
\left(\nabla_{X} \widetilde{e}_{i}, \widetilde{e}_{j}\right)+\left(\widetilde{e}_{i}, \nabla_{X} \widetilde{e}_{j}\right) & =\sum_{k}\left(\left(\dot{\rho}(\omega(X))_{k i} \widetilde{e}_{k}, \widetilde{e}_{j}\right)+\left(\widetilde{e}_{i}, \dot{\rho}(\omega(X))_{k j} \widetilde{e}_{k}\right)\right) \\
& =\sum_{k}\left(\dot{\rho}(\omega(X))_{k i} \delta_{k j}+\overline{\dot{\rho}(\omega(X))_{k j}} \delta_{k i}\right) \\
& =\dot{\rho}(\omega(X))_{j i}+\overline{\dot{\rho}(\omega(X))_{i j}}=0=X\left(\delta_{i j}\right)=X\left(\widetilde{e}_{i}, \widetilde{e}_{j}\right)
\end{aligned}
$$

since $\dot{\rho}(\omega(X))_{i j}$ is skew-hermitian $(\dot{\rho}: \mathfrak{s o}(2 m) \rightarrow \mathfrak{u}(V))$. Now write a section $s \in C^{\infty}(U, \mathcal{V})$ as $s=\sum_{i} s_{i} \widetilde{e}_{i}$ and $t \in C^{\infty}(U, \mathcal{V})$ as $t=\sum_{j} t_{j} \widetilde{e}_{j}$ for smooth functions $s_{i}, t_{j} \in C^{\infty}(U)$ and use Leibnitz's Rule to conclude the result on $U \subset M$, and hence globally.

Corollary 15.2.4. Let $M$ be a compact oriented Riemannian manifold of dimension $2 m$. Then all of the complex vector bundles associated to the principal bundle $P_{S O}$, namely $\Lambda_{\mathbb{C}}^{*}\left(T^{*} M\right), \mathbb{C}(M), \Lambda_{\mathbb{C}}^{e v}, \Lambda_{\mathbb{C}}^{o}, \Lambda_{\mathbb{C}}^{ \pm}$carry a natural associated connection or covariant derivative, called the Levi-Civita connection. This Levi-Civita connection is a unitary connection with respect to the hermitian inner product $(-,-)$ introduced on them as above (see Example 15.1.7), by the foregoing Lemma 15.2.3.

In the sequel, when we write $\nabla$ or $\nabla_{X}$ for any of these bundles without any further decorations, it is understood to mean covariant derivative with respect to the Levi-Civita connection on them.

Remark 15.2.5. There are the following immediate observations:
(i): The volume form $\omega:=d V \in \Lambda_{\mathbb{C}}^{n}(M)$ is covariantly constant, where $\omega=e_{1} \wedge e_{2} \wedge \ldots \wedge e_{n}$ in a local orthonormal frame $\left\{e_{i}\right\}$. This is because the Levi-Civita connection is compatible with the extended hermitian metric on $\Lambda_{\mathbb{C}}^{*} T^{*}(M)$. Indeed, by definition of $(-,-)$ on $\Lambda_{\mathbb{C}}^{*}\left(T_{x}^{*} M\right)$, we have $(\omega(x), \omega(x))_{x} \equiv 1$ for all $x \in M$, and also $\omega=\bar{\omega}$, since it is a real differential form. So, for any real tangent vector field $X \in C^{\infty}(M, T M), \nabla_{X} \omega$ is also a real differential $n$-form (i.e. equal to its conjugate). Hence $\nabla_{X} \omega=f(x) \omega$ for some smooth real valued function $f$ on $M$. Unitarity of the Levi-Civita connection gives:

$$
0=X(\omega(x), \omega(x))_{x}=\left(\nabla_{X} \omega(x), \omega(x)\right)_{x}+\left(\omega(x), \nabla_{X} \omega(x)\right)_{x}=2 f(x)
$$

which implies $\nabla_{X} \omega \equiv 0$.
(ii): We have noted following the Definition 15.1 .1 the vector bundle isomorphisms given by the symbol and quantisation maps between $\mathbb{C l}(M)$ and $\Lambda_{\mathbb{C}}^{*}\left(T^{*} M\right)$, which arise out of the $S O(2 m)$ - equivariant symbol and quantisation maps of $\mathbb{C} l_{2 m}$ and $\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m}\right)$. Since this last map is an isometry between the hermitian inner products $\langle-,-\rangle$, on both sides, it follows that the quantisation and symbol maps of bundles are bundle isometries with respect to $(-,-)$. The proof of (i) above can then be repeated verbatim for $\mathbb{C l}(M)$, to show that $\nabla_{X}\left(\omega_{M}\right) \equiv 0$ and $\nabla_{X}\left(\tau_{M}\right) \equiv 0$, where $\omega_{M}$ is the global volume element in $\mathbb{C l}(M)$ and $\tau_{M}$ the global chirality element as defined in Example 15.1.7.
(iii): From the fact that the derived representation

$$
\dot{\rho}: \mathfrak{s o}(2 m) \rightarrow \operatorname{End}_{\mathbb{C}}\left(\Lambda_{\mathbb{C}}^{k}\left(\mathbb{R}^{m}\right)\right.
$$

is a derivation, and satisfies:

$$
\dot{\rho}(X)\left(v_{1} \wedge v_{2} \wedge \ldots \wedge v_{k}\right)=\sum_{i}\left(v_{i_{1}} \ldots \wedge \dot{\rho}(X) v_{i} \wedge \ldots \wedge v_{k}\right)
$$

one immediately obtains that for $X \in T_{x}(M)$ :

$$
\nabla_{X}\left(\omega_{1} \wedge \ldots \wedge \omega_{k}\right)=\sum_{i}\left(\omega_{1} \wedge \ldots \wedge\left(\nabla_{X} \omega_{i}\right) . . \wedge \omega_{k}\right)
$$

Analogously, for the Clifford bundle $\mathbb{C l}(M)$, we have the derivation formula:

$$
\left.\nabla_{X}\left(\omega_{1} \ldots \omega_{k}\right)=\sum_{i} \omega_{1} \ldots\left(\nabla_{X} \omega_{i}\right) \ldots \omega_{k}\right)
$$

Definition 15.2.6 (Clifford connections). Let $\mathcal{E} \rightarrow M$ be a Clifford module over $M$, in the sense of Definition 15.1.5. We say that a connection (i.e. covariant differentiation) $\nabla^{\mathcal{E}}$ on $\mathcal{E}$ is a Clifford connection if:
(i): (Metric compatibility) It is a unitary connection with respect to the given hermitian inner product $(-,-)$ on $\mathcal{E}$, and
(ii): (Clifford compatibility) For all smooth sections $c$ of $\mathbb{C l}(M)$ and $s$ of $\mathcal{E}$, we have:

$$
\nabla_{X}^{\mathcal{E}}(c . s)=\left(\nabla_{X} c\right) . s+c . \nabla_{X}^{\mathcal{E}} s
$$

and another way of saying it is that the commutator of the covariant derivative and Clifford multiplication operators:

$$
\left[\nabla_{X}^{\mathcal{E}}, c .(-)\right]=\left(\nabla_{X} c\right) \cdot(-)
$$

where the right hand side denotes Clifford action by the (Levi-Civita) covariant derivative $\nabla_{X} c$.

Remark 15.2.7. Note that a Clifford connection as above on $\mathcal{E}$ will preserve $\mathcal{E}^{ \pm}$, if $\mathcal{E}^{ \pm}$are the $( \pm 1)$-eigenspaces from left action by the global chirality $\tau_{M} \in C^{\infty}(M, \mathbb{C} l(M))$. For if $s \in \mathcal{E}$ is a smooth section, then by (ii) in the above definition:

$$
\nabla_{X}^{\mathcal{E}}\left(\tau_{M} \cdot s\right)=\left(\nabla_{X} \tau_{M} \cdot s\right)+\tau_{M} \cdot\left(\nabla^{\mathcal{E}}\right) s=\tau_{M} \cdot\left(\nabla^{\mathcal{E}} s\right)
$$

since $\nabla_{X} \tau_{M} \equiv 0$ by (ii) of Remark 15.2 .5 above. Thus covariant differentiation $\nabla_{X}^{\mathcal{E}}$ commutes with left Clifford action by $\tau_{M}$, and thus maps the $( \pm 1)$-eigenspaces $\mathcal{E}^{ \pm}$of $\tau_{M}$. In particular, it restricts to connections on $\mathcal{E}^{ \pm}$, and these connections are also Clifford compatible.

Example 15.2.8. Regarding the bundle $\mathbb{C l}(M) \rightarrow M$ as a Clifford module via left multiplication (with either the chirality grading $\mathbb{C} l^{ \pm}$, or the parity grading $\left.\mathbb{C} l^{0}, \mathbb{C} l^{1}\right)$, the Levi-Civita connection defined in Corollary 15.2.3 above is a Clifford connection. The property (i) is metric compatible, as remarked there. The Clifford compatibility comes from the last statement in (iii) of Remark 15.2.5 above. Similarly, the Levi-Civita conection on all of the other Clifford modules discussed in Corollary 15.2.3 is a Clifford connection.

Example 15.2.9 (The Spin-connection on $\mathcal{S}(M)$ ). We recall the spin bundle $\mathcal{S}(M)$, and the half-spin bundles $\mathcal{S}^{ \pm}$, which were introduced in Definition 15.1.3. These are not bundles associated to the principal bundle $P_{S O}$, as we noted earlier. To put a connection on them, we need a connection on the principal $\operatorname{Spin}(2 m)$-bundle $P_{\text {spin }}$. So assume in this example that $M$ is a compact Riemannian manifold of dimension $2 m$ with a spin structure, and let $P_{\text {spin }} \rightarrow M$ be its principal spin bundle.

In (ii) and (iii) of Proposition 14.4.5, we noted that Lie( $\operatorname{Spin}(2 m))=C^{2}(V)=\operatorname{span}_{\mathbb{R}}\left\{e_{i} e_{j}: i<j\right\}$. Also we saw that the map $\tau: C^{2}(V) \rightarrow \mathfrak{s o}(2 m)$ satisfies:

$$
\tau\left(e_{i} e_{j}\right)=2\left(E_{j i}-E_{i j}\right)
$$

Indeed, this $\tau$ is precisely the derivative $\dot{\rho}$ of the map $\rho: \operatorname{Spin}(2 m) \rightarrow S O(2 m)$, because

$$
\begin{aligned}
\dot{\rho}\left(e_{i} e_{j}\right) & \left.=\frac{d}{d t} \right\rvert\, t=0 \\
& \left.=\frac{d}{d t}_{t=0} R_{2 t}^{i j}=2\left(\exp _{C}\left(t e_{i} e_{j}\right)\right)=E_{i j}\right)
\end{aligned}
$$

where $R_{\theta}^{i j}$ is the counter-clockwise rotation by $\theta$ in the 2 -plane $\mathbb{R} e_{i} \oplus \mathbb{R} e_{j}$, by using the last paragraph in the proof of (iv) Proposition 13.2.2.

Since $\rho \circ R_{g}=R_{\rho(g)} \circ \rho: P_{\text {spin }} \rightarrow P_{S O}$, we have:

$$
\begin{equation*}
\rho_{*} \circ R_{g *}=R_{\rho(g) *} \circ \rho_{*}: T P_{s p i n} \rightarrow T P_{S O} \tag{44}
\end{equation*}
$$

Denote the map $x \mapsto g x g^{-1}$ on a Lie group $G$ as $\operatorname{Ad}^{G} g$, and its derivative at $1 \in G$ simply as $\operatorname{Ad} g: \mathfrak{g} \rightarrow \mathfrak{g}$ where $\mathfrak{g}:=\operatorname{Lie}(G)=T_{1}(G)$. Now, recalling the homomorphism $\rho: \operatorname{Spin}(2 m) \rightarrow S O(2 m)$ (also denoted by the same symbol $\rho$, in keeping with the definition of a spin structure), we see that the homomorphism:

$$
\rho \circ \mathrm{Ad}^{\operatorname{Spin}(2 m)} g: \operatorname{Spin}(2 m) \rightarrow S O(2 m)
$$

is the same as the homomorphism:

$$
\operatorname{Ad}^{S O} \rho(g) \circ \rho: \operatorname{Spin}(2 m) \rightarrow S O(2 m)
$$

for all $g \in \operatorname{Spin}(2 m)$. By equating the derivative at the identity $1 \in \operatorname{Spin}(2 m)$ of both these maps and noting that $\dot{\rho}=D \rho(1)=\tau$, we have:

$$
\begin{equation*}
\tau \circ \operatorname{Ad} g=\operatorname{Ad}(\rho(g)) \circ \tau \tag{45}
\end{equation*}
$$

Now define a $C^{2}(V)$-valued 1-form on $P_{\text {spin }}$ by:

$$
\widetilde{\omega}:=\tau^{-1}\left(\rho^{*} \omega\right)
$$

where $\omega \in \Lambda^{1}(P, \mathfrak{s o}(2 m))$ is the Levi-Civita connection 1-form on $P_{S O}$, and $\rho: P_{\text {spin }} \rightarrow P_{S O}$ is the double covering map. We need to check $\widetilde{\omega}$ satisfies the correct translation property. Using equations (44) and (45), we have:

$$
\begin{aligned}
R_{g}^{*} \widetilde{\omega}(v) & =\widetilde{\omega}\left(R_{g *} v\right)=\tau^{-1}\left[\left(\rho^{*} \omega\right)\left(R_{g *} v\right)\right]=\tau^{-1}\left[\omega\left(\rho_{*} R_{g *} v\right)\right] \\
& =\tau^{-1}\left[\omega\left(R_{\rho(g) *} \rho_{*} v\right)\right]=\tau^{-1}\left[\left(R_{\rho(g)}^{*} \omega\right)\left(\rho_{*} v\right)\right] \\
& =\tau^{-1} \operatorname{Ad} \rho(g)\left[\omega\left(\rho_{*} v\right)\right]=\operatorname{Ad}(g) \tau^{-1}\left[\omega\left(\rho_{*} v\right)\right] \\
& =\operatorname{Ad}(g)\left[\left(\tau^{-1} \rho^{*} \omega\right)(v)\right]=\operatorname{Ad}(g) \widetilde{\omega}(v)
\end{aligned}
$$

This connection form on $P_{\text {spin }}$ is called the spin connection.
If $\widetilde{\sigma}: U \rightarrow P_{\text {spin } \mid U}$ is a local section of $P_{\text {spin }}$ on a coordinate chart $U \subset M$, then $\sigma:=\rho \circ \widetilde{\sigma}: U \rightarrow P_{S O \mid U}$ will be a local section for $P_{S O}$. Then

$$
\tilde{\sigma}^{*} \widetilde{\omega}=\tau^{-1} \widetilde{\sigma}^{*} \rho^{*} \omega=\tau^{-1} \sigma^{*} \omega
$$

Let $\sigma^{*} \omega$ be given by the Cartan connection matrix of 1-forms (see Definition 15.2.2) [ $\omega_{i j}$ ] on $U$, so we can write $\sigma^{*} \omega:=\sum_{i<j} \omega_{i j} E_{i j}$, where $\omega_{i j}$ are 1-forms on $U$. Since $\tau^{-1}\left(E_{i j}\right)=\frac{1}{2} e_{i} e_{j}$, it follows that the Cartan connection 1-form on $U$ for the spin connection $\widetilde{\omega}$ is given on $U$ by:

$$
\begin{equation*}
\omega^{s p}=\widetilde{\sigma}^{*} \widetilde{\omega}=\frac{1}{2} \sum_{i<j} \omega_{i j} e_{i} e_{j} \tag{46}
\end{equation*}
$$

as an element of $C^{\infty}\left(U, C^{2}(V)\right)$ where we are making the identification $C^{2}(V)=\operatorname{Lie}(\operatorname{Spin}(2 m))$.
Now that we have a $\operatorname{Spin}(2 m)$-connection on $P_{\text {spin }}$, all associated vector bundles get a connection by the same procedure as before. That is, if $\mu: \operatorname{Spin}(2 m) \rightarrow G L(V)$ is any representation, and $\mathcal{V}=P_{\text {spin }} \times_{\mu} V$ is the associated bundle, then giving the Cartan 1-forms on a trivialising coordinate neighbourhood $U \subset M$ for $P_{\text {spin }}$ by

$$
\omega_{i j}^{\mathcal{V}}:=\dot{\mu}\left(\omega^{s p}\right)_{i j}
$$

will define covariant differentiation $\nabla^{\mathcal{V}}$ on $\mathcal{V}$. Again if the representation is unitary, the bundle will be hermitian, and $\dot{\mu}: C^{2}(V) \rightarrow \mathfrak{u}(n)$ implies that the connection $\nabla^{\mathcal{V}}$ will be compatible with this hermitian metric, i.e. will be a unitary connection.

Letting $V:=\Delta_{2 m}=S_{2 m}$, the complex spin-representation defined in Proposition 14.2.1, we have the spin-bundle $\mathcal{S}(M)$ defined in Definition 15.1.3 as the associated bundle $P \times{ }_{\mu} \Delta_{2 m}$. Here we are denoting this representation by $\mu$, and since we saw in (ii) and (iii) of Proposition 15.1.4 that the representation of $\mathbb{C} l_{2 m}$ was self-adjoint with respect to the hermitian inner-product of $\Delta_{2 m}$, and that this representation is unitary, viz. $\mu: \operatorname{Spin}(2 m) \rightarrow U\left(\Delta_{2 m}\right)$. Since the connection form $\dot{\mu}(\widetilde{\omega})$ takes values in $\mathfrak{u}\left(\Delta_{2 m}\right)$, the spin connection on $\mathcal{S}(M)$ is a unitary connection, and metric compatibility follows by definition.

To check Clifford compatibility, we need compute the commutator of the Cartan coefficients on an open set $U$, i.e. the skew-hermitian matrix $\left[\omega^{\mathcal{S}}\right]=\left[\dot{\mu}\left(\omega^{s p}\right)_{i j}\right]$ and Clifford multiplication by $c \in \mathbb{C} l(M)_{U}$. Note that $\mu: \operatorname{Spin}(2 m) \rightarrow U\left(\Delta_{2 m}\right)$ is the restriction of the Clifford action $\mu: \mathbb{C l}_{2 M} \rightarrow \operatorname{End}_{\mathbb{C}}\left(\Delta_{2 m}\right)$. Also with the identification of $\operatorname{Lie}(\operatorname{Spin}(2 m))=C^{2}(V)$, it follows that:

$$
\begin{aligned}
{[\dot{\mu}(x), \mu(c)] } & =\left[\frac{d\left(\mu\left(\exp _{C} t x\right)\right.}{d t}, \mu(c)\right]=\frac{d}{d t}{ }_{\mid t=0}\left(\mu\left[\exp _{C} t x, c\right]\right)=\mu\left[\frac{d \exp _{C} t x}{d t}{ }_{\mid t=0}, c\right] \\
& =\mu([x, c]) \text { for } x \in C^{2}(V), c \in \mathbb{C} l_{2 m}
\end{aligned}
$$

Now, for a section $s \in \mathcal{S}(M)_{\mid U}, c$ a smooth section for $\mathbb{C l}(M)_{\mid U}$ and $X$ a smooth real vector-field on $U$, where $U$ is a trivialising neighbourhood for $P_{\text {spin }}$, we have :

$$
\begin{aligned}
{\left[\nabla_{X}^{\mathcal{S}}, c\right] s } & =\nabla_{X}^{\mathcal{S}}(c . s)-c . \nabla_{X}^{\mathcal{S}} s=\omega^{\mathcal{S}}(X)(c . s)-c . \omega^{\mathcal{S}}(X) s \\
& =\dot{\mu}\left(\omega^{s p}(X)\right) \mu(c) s-\mu(c)\left(\dot{\mu}\left(\omega^{s p}(X)\right) s=\left(\left[\dot{\mu} \omega^{s p}(X), \mu(c)\right]\right) s\right. \\
& =\left(\mu\left[\omega^{s p}(X), c\right]\right) s=\mu\left(\tau\left(\omega^{s p}(X)\right) c\right) s \quad(\text { by }(\mathrm{ii}) \text { of 14.4.5) } \\
& =\mu\left(\left[\omega^{S O}(X)\right] c\right) s=\left(\nabla_{X} c\right) . s
\end{aligned}
$$

which shows Clifford compatibility of the spin connection $\nabla^{\mathcal{S}}$.

Thus the spin connection on $\mathcal{S}(M)$ is a Clifford connection.

Proposition 15.2.10. Let $M$ be a spin manifold of dimension $2 m$, and Let $\mathcal{V} \rightarrow M$ be any hermitian complex vector bundle with the inner-product $(-,-)_{\mathcal{V}}$, and a unitary connection $\nabla^{\mathcal{V}}$ on it. Then the tensor product bundle:

$$
\mathcal{E}:=\mathcal{S}(M) \otimes_{\mathbb{C}} \mathcal{V}
$$

equipped with the natural Clifford action, and the natural hermitian inner product $(-,-)_{\mathcal{E}}:=(-,-)_{\mathcal{S}} \otimes(-,-)_{\mathcal{V}}$ is a Clifford module on $M$. The tensor product connection $\nabla^{\mathcal{E}}$ of the spin connection $\nabla^{\mathcal{S}}$ and $\nabla^{\mathcal{V}}$ is a Clifford connection on this bundle.

Proof: The Clifford action is given by:

$$
c .(s \otimes v)=c . s \otimes v \quad \text { for } c \in \mathbb{C} l(M), s \in \mathcal{S}(M), v \in \mathcal{V}
$$

Clearly the supermodule structure is $\mathcal{E}^{ \pm}=\mathcal{S}^{ \pm}(M) \otimes_{\mathbb{C}} \mathcal{V}$. The tensor product hermitian inner product is given on decomposable elements by:

$$
(s \otimes v, t \otimes w)_{\mathcal{E}}:=(s, t)_{\mathcal{S}}(v, w)_{\mathcal{V}}
$$

Since $S^{+}(M)$ and $S^{-}(M)$ are orthogonal under $(-,-)_{\mathcal{S}}$, it easily follows that $\mathcal{E}^{+}$and $\mathcal{E}^{-}$are orthogonal under $(-,-)_{\mathcal{E}}$.

To check self-adjointness of Clifford action, it is enough to do it on decomposable elements, and for that we have:
 by the self-adjointness of the Clifford module $\mathcal{S}(M)$.

The tensor product connection $\nabla^{\mathcal{E}}$ is unitary because we again check on decomposable sections that:

$$
\begin{aligned}
\left(\nabla_{X}^{\mathcal{E}}(s \otimes v), t \otimes w\right)_{\mathcal{E}} & +\left(s \otimes v, \nabla^{\mathcal{E}}(t \otimes w)\right)_{\mathcal{E}}=\left(\nabla_{X}^{\mathcal{S}} s \otimes v+s \otimes \nabla^{\mathcal{V}} v, t \otimes w\right)_{\mathcal{E}}+\left(s \otimes v, \nabla_{X}^{\mathcal{S}} t \otimes w+t \otimes \nabla^{\mathcal{V}} w\right)_{\mathcal{E}} \\
& =\left[\left(\nabla_{X}^{\mathcal{S}} s, t\right)_{\mathcal{S}}+\left(s, \nabla_{X}^{\mathcal{S}} t\right)_{\mathcal{S}}\right](v, w)_{V}+(s, t)_{\mathcal{S}}\left[\left(\nabla_{X}^{\mathcal{V}} v, w\right)_{\mathcal{V}}+\left(v, \nabla_{X}^{\mathcal{V}} w\right)_{\mathcal{V}}\right] \\
& =\left[X(s, t)_{\mathcal{S}}\right](v, w)_{\mathcal{V}}+(s, t)_{\mathcal{S}}\left[X(v, w)_{\mathcal{W}}\right]=X\left((s \otimes v, t \otimes w)_{\mathcal{E}}\right)
\end{aligned}
$$

To check Clifford compatibility, again:

$$
\begin{aligned}
& \nabla_{X}^{\mathcal{E}}(c .(s \otimes v))=\nabla_{X}^{\mathcal{E}}(c . s \otimes v)=\nabla_{X}^{\mathcal{S}}(c . s) \otimes v+c . s \otimes \nabla^{\mathcal{V}} v=\left(\nabla_{X} c . s+c . \nabla_{X}^{\mathcal{S}} s\right) \otimes v+c . s \otimes \nabla^{\mathcal{V}} v \\
& =\left(\nabla_{X} c . s \otimes v\right)+c .\left(\nabla_{X}^{\mathcal{S}} s \otimes v+s \otimes \nabla^{\mathcal{V}} v\right)=\left(\nabla_{X} c\right) .(s \otimes v)+c . \nabla_{X}^{\mathcal{E}}(s \otimes v)
\end{aligned}
$$

using the Clifford compatibility of $\nabla^{\mathcal{S}}$ proved in Example 15.2.9 above. This proves the proposition.

Corollary 15.2.11. Let $M$ be a spin manifold of dimension $2 m$, and let $\mathcal{E} \rightarrow M$ be a Clifford module (over the Clifford bundle $\mathbb{C l}(M) \rightarrow M)$. Then there exists a hermitian complex vector bundle $\mathcal{V} \rightarrow M$ with a compatible unitary connection $\nabla^{\mathcal{V}}$ such that the bundle $\mathcal{E} \simeq \mathcal{S} \otimes \mathcal{V}$ as Clifford modules over $M$, and $\mathcal{E}$ becomes a Clifford module with a Clifford connection given by the tensor-product connection of the spin connection $\nabla^{\mathcal{S}}$ and $\nabla^{\mathcal{V}}$. Indeed, every Clifford connection on $\mathcal{E}$ is obtained in this way.

Proof: By the Proposition 15.1.9, we have a complex hermitian vector bundle $\mathcal{V} \rightarrow M$ such that $\mathcal{E} \simeq \mathcal{S}(M) \otimes \mathcal{V}$ as Clifford modules. Since $V$ has a hermitian metric, it has a compatible unitary connection $\nabla^{\mathcal{V}}$ (by using partitions of unity, for example).

Define the connection $\nabla^{\mathcal{E}}$ as the tensor-product connection of spin-connection $\nabla^{\mathcal{S}}$ and $\nabla^{\mathcal{V}}$. Then we are done by the Proposition 15.2.10 above.

To see the last assertion, note that for any finite dimensional module $E$ over $\mathbb{C l}_{2 m}$, we have an isomorphism of $\mathbb{C} l_{2 m}$-modules by Proposition 14.4.6:

$$
E \simeq S_{2 m} \otimes_{\mathbb{C}} V \text { for } V:=\operatorname{hom}_{\mathbb{C} l_{2 m}}\left(S_{2 m}, E\right)
$$

By breaking up $E$ into irreducibles $E_{i} \simeq S_{2 m}$ as before, and noting that hom $\mathbb{C l}_{2 m}\left(E_{i}, E_{j}\right)$ is one dimensional, it is trivial to check that the natural map:

$$
\begin{aligned}
\operatorname{hom}_{\mathbb{C}}(V, V) & \rightarrow \operatorname{hom}_{\mathbb{C} l_{2 m}}(E, E)=\operatorname{hom}_{\mathbb{C} l_{2 m}}\left(S_{2 m} \otimes_{\mathbb{C}} V, S_{2 m} \otimes_{\mathbb{C}} V\right) \\
\Lambda & \mapsto I d_{S_{2 m}} \otimes \Lambda
\end{aligned}
$$

is an isomorphism of complex vector-spaces. Since this isomorphism is canonical (basis-independent), we have an isomorphism of complex vector bundles:

$$
\operatorname{hom}_{\mathbb{C} l_{2 m}}(\mathcal{E}, \mathcal{E}) \simeq \operatorname{hom}_{\mathbb{C}}(\mathcal{V}, \mathcal{V})
$$

Now if $\nabla^{\mathcal{E}}$ denotes the tensor product connection defined above, and $\widetilde{\nabla}^{\mathcal{E}}$ is another Clifford connection, it follows by Leibnitz's rule that:

$$
\left(\nabla^{\mathcal{E}}-\widetilde{\nabla}^{\mathcal{E}}\right)(f s)=f\left(\nabla^{\mathcal{E}}-\widetilde{\nabla}^{\mathcal{E}}\right) s \text { for all } f \in C^{\infty}(M), s \in C^{\infty}(M, \mathcal{E})
$$

which shows that $\left(\nabla^{\mathcal{E}}-\widetilde{\nabla}^{\mathcal{E}}\right)=\alpha$ for some smooth section $\alpha \in C^{\infty}\left(T^{*} M \otimes \operatorname{hom}_{\mathbb{C}}(\mathcal{E}, \mathcal{E})\right)$. The Clifford compatibility condition shows that $[\alpha, c] \equiv 0$ for all smooth sections $c \in C^{\infty}(M, \mathbb{C} l(M))$, i.e. $\alpha \in C^{\infty}\left(M, T^{*} M \otimes\right.$ $\operatorname{hom}_{\mathbb{C l}_{2 m}}(\mathcal{E}, \mathcal{E})$. By the above, this last space is isomorphic to $C^{\infty}\left(M, T^{*} M \otimes \operatorname{hom}_{\mathbb{C} l_{2 m}}(\mathcal{V}, \mathcal{V})\right)$, so that $\alpha=1 \otimes \beta$ for some section $\beta \in C^{\infty}\left(M, T^{*} M \otimes \operatorname{hom}_{\mathbb{C} l_{2 m}}(\mathcal{V}, \mathcal{V})\right)$. This shows that $\widetilde{\nabla}^{\mathcal{E}}$ is given by:

$$
\widetilde{\nabla}^{\mathcal{E}}=\nabla^{\mathcal{E}}-\alpha=\left(\nabla^{\mathcal{S}}\right) \otimes 1+1 \otimes \nabla^{\mathcal{V}}-1 \otimes \beta=\nabla^{\mathcal{S}} \otimes 1+1 \otimes\left(\nabla^{\mathcal{V}}-\beta\right)
$$

which is the tensor product connection of $\nabla^{\mathcal{S}}$ and $\widetilde{\nabla}^{\mathcal{V}}:=\nabla^{\mathcal{V}}-\beta$. This proves the last assertion.

Definition 15.2.12. We say that a Clifford module $\mathcal{E} \rightarrow M$ over $M$ is a Dirac Bundle if it has a compatible Clifford connection.

Example 15.2.13. The bundles $\mathbb{C l}(M) \rightarrow M, \Lambda_{\mathbb{C}}^{*}\left(T^{*} M\right) \rightarrow M$, on an oriented Riemannian manifold of dimension $2 m$ are all Dirac bundles, by Example 15.2.8. The spin bundle $\mathcal{S}(M) \rightarrow M$ on a spin manifold of dimension $2 m$ is a Dirac bundle, by Example 15.2.9 above.

The above Corollary 15.2 .11 says that to generate any Dirac bundle on a spin manifold $M$ of dimension 2 m , it is enough to start with the prototypical spinor bundle $\mathcal{S}(M) \rightarrow M$ with its natural structure as a Dirac bundle, and then twist it with various hermitian bundles $\mathcal{V} \rightarrow M$ (with compatible unitary connections).

### 15.3. Dirac operator on a Dirac bundle.

Definition 15.3.1. Let $\mathcal{E} \rightarrow M$ be a Dirac bundle on $M$, an Riemannian manifold of dimension $2 m$, with chirality grading by $\mathcal{E}^{ \pm}$. Let $\mathbb{C} l(M) \rightarrow M$ be the Clifford bundle of $M$. Denote by $c$ the Clifford action on $\mathcal{E}$ :

$$
T^{*}(M) \otimes \mathcal{E} \xrightarrow{c} \mathcal{E}
$$

Let $\nabla^{\mathcal{E}}$ denote the Clifford connection on $\mathcal{E}$. The Dirac operator on $\mathcal{E}$ is the operator $D$ defined by the composite:

$$
C^{\infty}(M, \mathcal{E}) \xrightarrow{\nabla^{\mathcal{E}}} C^{\infty}\left(M, T^{*} M \otimes \mathcal{E}\right) \xrightarrow{c} C^{\infty}(M, \mathcal{E})
$$

Since $c\left(T^{*} M \otimes \mathcal{E}^{ \pm}\right) \subset \mathcal{E}^{\mp}$, and by definition the Clifford connection preserves the subbundles $\mathcal{E}^{ \pm}$, it follows that the Dirac operator is also $\mathbb{Z}_{2}$-graded, and $D=D^{+} \oplus D^{-}$, where

$$
D^{ \pm}: C^{\infty}\left(M, \mathcal{E}^{ \pm}\right) \rightarrow C^{\infty}\left(M, \mathcal{E}^{\mp}\right)
$$

Remark 15.3.2 (Dirac operator in local coordinates). In a local coordinate chart, we may choose an orthonormal frame $\left\{e_{i}\right\}_{i=1}^{2 m}$ of the cotangent bundle $T^{*} M$. Then for a smooth section $s \in C^{\infty}(M, \mathcal{E})$, we have

$$
\nabla^{\mathcal{E}} s=\sum_{i=1}^{2 m} e_{i} \otimes \nabla_{e_{i}}^{\mathcal{E}} s
$$

so that the Dirac operator is expressed as

$$
D s=\sum_{i=}^{2 m} e_{i} \cdot \nabla_{e_{i}}^{\mathcal{E}} s
$$

where the dot denotes Clifford action. Since $\nabla_{e_{i}}^{\mathcal{E}}$ are 1st order differential operators, it follows that $D$ is a 1 st order ifferential operator.

Proposition 15.3.3 (Self-adjointness of the Dirac operator). Let $M$ be a compact oriented Riemannian manifold of dimension $2 m$, and let $(-,-)$ denote the given hermitian inner-product on a Dirac bundle $\mathcal{E} \rightarrow M$. Define the global $L^{2}$-inner product on $C^{\infty}(M, \mathcal{E})$ by $(s, t)_{M}:=\int_{M}(s(x), t(x))_{x} d V(x)$. Then Dirac operator $D: C^{\infty}(M, \mathcal{E}) \rightarrow C^{\infty}(M, \mathcal{E})$ is formally self-adjoint with respect to $(-,-)_{M}$. In particular $D^{+}: C^{\infty}\left(M, \mathcal{E}^{+}\right) \rightarrow$ $C^{\infty}\left(M, \mathcal{E}^{-}\right)$and $D^{-}: C^{\infty}\left(M, \mathcal{E}^{-}\right) \rightarrow C^{\infty}\left(M, \mathcal{E}^{+}\right)$are adjoints of each other.

Proof: Fix a point $x \in M$, and fix a synchronous orthonormal frame in a nieghbourhood $U$ of $x$, i.e. for the Levi-Civita connection we have

$$
\text { (a) } \nabla e_{i}(x) \equiv 0 \quad \text { (b) } \quad e_{i, x}=\partial_{i, x}=\frac{\partial}{\partial x_{i} \mid x} \quad \text { for all } i=1, . ., 2 m
$$

for some coordinate system $\left(x_{1}, \ldots, x_{2 m}\right)$ on $U$. By the self-adjointness of Clifford multiplication with respect to the pointwise hermitian inner product, unitarity and Clifford compatibility of the connection, and synchronicity of the frame $\left\{e_{i}\right\}$, we have:

$$
\begin{aligned}
\left(e_{i} \nabla_{e_{i}}^{\mathcal{E}} s, t\right)_{x} & =-\left(\nabla_{e_{i}}^{\mathcal{E}} s, e_{i} \cdot t\right)_{x}=-e_{i}\left(s, e_{i} \cdot t\right)_{x}+\left(s, \nabla_{e_{i}}^{\mathcal{E}}\left(e_{i} \cdot t\right)_{x}\right. \\
& =-e_{i}\left(s, e_{i} \cdot t\right)_{x}+\left(s,\left(\nabla_{e_{i}} e_{i}\right) \cdot t\right)_{x}+\left(s, e_{i} \cdot \nabla_{e_{i}}^{\mathcal{E}} t\right)_{x} \\
& =-\partial_{i}\left(s, \partial_{i} \cdot t\right)_{x}+\left(s, e_{i} \cdot \nabla_{e_{i}}^{\mathcal{E}} t\right)_{x}
\end{aligned}
$$

Summing over $i$ we find:

$$
(D s, t)_{x}-(s, D t)_{x}=-\delta \sigma(x)
$$

where $\sigma$ is the 1-form $v \mapsto(s, v . t)_{x}=\sum_{i}\left(s, e_{i} . t\right) e_{i}^{*}$ on $U$, and $\delta \sigma(x)=\sum_{i} \partial_{i}\left(s, e_{i} . t\right)_{x}= \pm(* d * \sigma)(x)$ (i.e. the divergence of $\sigma)$. Integrating over $M$, and noting that $\int_{M}(\delta \sigma) d V=\int_{M} \delta \sigma \wedge(* 1)== \pm \int_{M} \sigma \wedge d(* 1)=0$, we have:

$$
(D s, t)_{M}=(s, D t)_{M}
$$

and our assertion follows. The last statement is clear from the fact that the restriction of $D$ to $C^{\infty}\left(M, \mathcal{E}^{ \pm}\right)$ are $D^{ \pm}$respectively.

Corollary 15.3.4. With the hypothesis of the previous proposition, the second order differential operator:

$$
D^{2}: C^{\infty}(M, \mathcal{E}) \rightarrow C^{\infty}(M, \mathcal{E})
$$

is formally self-adjoint with respect to $(-,-)_{M}$. Its restrictions, namely the composites $D^{+} D^{-}$and $D^{-} D^{+}$:

$$
C^{\infty}\left(M, \mathcal{E}^{ \pm}\right) \xrightarrow{D^{ \pm}} C^{\infty}\left(M, \mathcal{E}^{\mp}\right) \xrightarrow{D^{\mp}} C^{\infty}\left(M, \mathcal{E}^{ \pm}\right)
$$

are self-adjoint.
15.4. Weitzenbock Formulas. We need to assert that the square of the Dirac operator on a Dirac bundle is a generalised Laplacian. To this end, we have the following.

Definition 15.4.1. Let $\mathcal{E} \rightarrow M$ be a Dirac bundle, and let $\nabla^{\mathcal{E}}$ be its Clifford connection. Then for two real tangent vector fields $X, Y \in C^{\infty}\left(M, T_{x}(M)\right)$, we define a 2 -form $\Omega^{\mathcal{E}} \in C^{\infty}\left(\Lambda^{2} T^{*} M \otimes \operatorname{hom}_{\mathbb{C}}(\mathcal{E}, \mathcal{E})\right)$ by:

$$
\Omega^{\mathcal{E}}(X, Y) \otimes s=\nabla_{X}^{\mathcal{E}} \nabla_{Y}^{\mathcal{E}} s-\nabla_{Y}^{\mathcal{E}} \nabla_{X}^{\mathcal{E}} s-\nabla_{[X, Y]}^{\mathcal{E}} s
$$

The 2-form $\Omega^{\mathcal{E}}$ is called the curvature of the Clifford connection $\nabla^{\mathcal{E}}$ or just the Clifford curvature of $\mathcal{E}$.

That the object on the right side of the definition defines a 2-form follows by changing $X$ to $f X$ and $Y$ to $g Y$ where $f$ and $g$ are two smooth functions, and calculating by Leibnitz's rule that:

$$
\left[\nabla_{f X}^{\mathcal{E}}, \nabla_{g Y}^{\mathcal{E}}\right]-\nabla_{[f X, g Y]}^{\mathcal{E}}=f g\left(\left[\nabla_{X}^{\mathcal{E}}, \nabla_{Y}^{\mathcal{E}}\right]-\nabla_{[X, Y]}^{\mathcal{E}}\right)
$$

Exercise 15.4.2 (Clifford curvature in local coordinates). Let $\left\{s_{i}\right\}$ be an orthonormal frame of $\mathcal{E}_{\mid U}$ with respect to $(-,-)$, the hermitian inner prroduct on $\mathcal{E}$, where $U \subset M$ is a trivialising neighbourhood of $\mathcal{E}$. Then we may write:

$$
\nabla^{\mathcal{E}} s_{j}=\sum_{i=1}^{\mathrm{rk}_{\mathbb{C}} \mathcal{E}} \omega_{i j} s_{i}
$$

where $\omega_{i j}$ is the skew-hermitian matrix of Cartan connection 1-forms on $U$. Apply the definitions to show that $\Omega^{\mathcal{E}}$ is another skew hermitian matrix of 2-forms given by:

$$
\Omega_{i j}^{\mathcal{E}}=d \omega_{i j}+\sum_{l=1}^{\mathrm{rk}_{\mathbb{C}} \mathcal{E}}\left(\omega_{i l} \wedge \omega_{l j}-\omega_{l j} \wedge \omega_{l i}\right)
$$

which is often abbreviated in the notation:

$$
\Omega=d \omega+[\omega, \omega]
$$

Proposition 15.4.3 (Weitzenbock for a Dirac Bundle). Let $\mathcal{E} \rightarrow M$ be a Dirac bundle on an oriented compact Riemannian manifold of dimension 2 m . Then the square of the Dirac operator is given by:

$$
D^{2}=\nabla^{\mathcal{E} *} \nabla^{\mathcal{E}}+\frac{1}{2} \Omega^{\mathcal{E}}
$$

Proof: We again fix a point $x \in M$ and choose a synchronous frame $\left\{e_{i}\right\}$ for $T M_{\mid U}$ for some neighbourhood $U$ of $x$. Then note that we have $\left[e_{i}, e_{j}\right]_{x}=\left[\partial_{i, x}, \partial_{j, x}\right]=0$ for all $1 \leq i, j \leq 2 m$, and also $\left(\nabla_{e_{i}} e_{j}\right)(x)=0$ for the Levi-Civita connection on $T M$. Then $\operatorname{denote} \nabla_{e_{j}}^{\mathcal{E}}$ by $\nabla_{j}^{\mathcal{E}}$, and similarly for the Levi-Civita covariant derivative $\nabla_{e_{j}}$ by $\nabla_{j}$. We first note that:

$$
\nabla^{\mathcal{E}}: C^{\infty}(M, \mathcal{E}) \rightarrow C^{\infty}\left(M, T^{*} M \otimes \mathcal{E}\right)
$$

has a global $L_{2}$-adjoint:

$$
\nabla^{\mathcal{E} *}: C^{\infty}\left(M, T^{*} M \otimes \mathcal{E}\right) \rightarrow C^{\infty}(M, \mathcal{E})
$$

which satisfies:

$$
\left(\nabla^{\mathcal{E}} s, \omega \otimes t\right)_{M}=\left(s, \nabla^{\mathcal{E} *}(\omega \otimes t)\right)_{M} \text { for all } \omega \in \Lambda^{1}(M, \mathbb{C}), \quad s, t \in C^{\infty}(M, \mathcal{E})
$$

(Note that the hermitian inner product on $T_{\mathbb{C}}^{*} M \otimes \mathcal{E}$ is taken to be the tensor product hermitian inner-product of the hermitian inner products on the two factors.) In fact, we claim that in terms of the synchronous frame at $x$ defined above, we have:

$$
\begin{equation*}
\nabla^{\mathcal{E} *}\left(e_{i}^{*} \otimes t\right)(x)=-\left(\nabla_{i}^{\mathcal{E}} t\right)(x) \tag{47}
\end{equation*}
$$

To verify this, we write $\nabla^{\mathcal{E}} s=\sum_{j} e_{j}^{*} \otimes \nabla_{j}^{\mathcal{E}} s$, then:

$$
\begin{aligned}
\left(\nabla^{\mathcal{E}} s, e_{i}^{*} \otimes t\right)_{x} & =\sum_{j}\left(e_{j}^{*} \otimes \nabla_{j}^{\mathcal{E}} s, e_{i}^{*} \otimes t\right)_{x}=\sum_{j}\left(e_{j}, e_{i}\right)_{x}\left(\nabla_{j}^{\mathcal{E}} s, t\right)_{x} \\
& =\left(\nabla_{i}^{\mathcal{E}} s, t\right)_{x}=e_{i}(s, t)_{x}-\left(s, \nabla_{i}^{\mathcal{E}} t\right) \\
& \left.=\sum_{j} e_{j}\left(s, e_{j}\right\lrcorner\left(e_{i}^{*} \otimes t\right)\right)_{x}+\left(s, \nabla^{\mathcal{E} *}\left(e_{i}^{*} \otimes t\right)\right)_{x} \\
& =\delta \sigma(x)+\left(s, \nabla^{\mathcal{E} *}\left(e_{i}^{*} \otimes t\right)\right)_{x}
\end{aligned}
$$

where $\sigma$ is the 1 -form defined by $\left.v_{x} \mapsto(s, v\lrcorner\left(e_{i}^{*} \otimes t\right)\right)_{x}$. Again, integrating over $M$ and noting that $\int_{M} \delta \sigma d V=0$ by Stokes Theorem, we have our assertion.

Now we compute for a section $s \in C^{\infty}(U, \mathcal{E})$ that:

$$
\begin{aligned}
D^{2} s(x) & =\sum_{i} e_{i} \nabla_{i}^{\mathcal{E}}\left(\sum_{j} e_{j} \nabla_{j}^{\mathcal{E}} s\right)=\sum_{i, j}\left(e_{i} \nabla_{i} e_{j}\right) \cdot \nabla_{j}^{\mathcal{E}} s+\sum_{i, j} e_{i} e_{j} \nabla_{i}^{\mathcal{E}} \nabla_{j}^{\mathcal{E}} s \\
& =\sum_{i, j} e_{i} e_{j} \nabla_{i}^{\mathcal{E}} \nabla_{j}^{\mathcal{E}} s \text { since }\left(\nabla_{i} e_{j}\right)(x)=0 \\
& \left.=-\sum_{i} \nabla_{i}^{\mathcal{E}} \nabla_{i}^{\mathcal{E}} s+\sum_{i<j} e_{i} \cdot e_{j} \cdot\left(\nabla_{i}^{\mathcal{E}} \nabla_{j}^{\mathcal{E}}-\nabla_{j}^{\mathcal{E}} \nabla_{i}^{\mathcal{E}} s\right) \quad \text { (by Clifford relations on } e_{i}\right) \\
& =\sum_{i} \nabla^{\mathcal{E} *}\left(e_{i}^{*} \otimes \nabla_{i}^{\mathcal{E}} s\right)+\frac{1}{2} \Omega^{\mathcal{E}} s \quad\left(\text { since } \quad\left[e_{i}, e_{j}\right](x)=0\right. \text { and by using (47) above) } \\
& =\nabla^{\mathcal{E} *} \nabla^{\mathcal{E}} s+\frac{1}{2} \Omega^{\mathcal{E}} s
\end{aligned}
$$

where $\Omega^{\mathcal{E}} s(x)=\left[\sum_{i, j} e_{i} e_{j} \cdot\left(\nabla_{i}^{\mathcal{E}} \nabla_{j}^{\mathcal{E}}-\nabla_{j}^{\mathcal{E}} \nabla_{i}^{\mathcal{E}}\right) s\right](x)$ in our synchronous frame $\left\{e_{i}\right\}$ around $x$. This proves the assertion.

Proposition 15.4.4 (Weitzenbock formula for the Spin Bundle). Let $M$ be a compact spin manifold of dimension $2 m$, and let $\mathcal{S}(M) \rightarrow M$ be the spin bundle on $M$ with its natural structure as a Dirac bundle with its spin connection $\nabla^{\mathcal{S}}$ (see Examples 15.2 .9 and 15.2.13). Then for its Dirac operator $D$, we have:

$$
D^{2}=\nabla^{\mathcal{S} *} \nabla^{\mathcal{S}}+\frac{1}{4} \sum_{i, j} k
$$

where $k$ is the scalar curvature function of the Riemannian metric on $M$.

Proof: In view of the Proposition 15.4 .3 above, we need to calculate the curvature operator $\Omega^{\mathcal{S}}$ of the spin connection $\nabla^{\mathcal{S}}$. We recall from the Example 15.2 .9 that the spin connection 1-form $\widetilde{\omega} \in C^{\infty}\left(P_{\text {spin }}, T^{*} P_{\text {spin }} \otimes C^{2}(V)\right)$ is related to the connection 1-form $\omega \in C^{\infty}\left(P_{S O}, T^{*} P_{S O} \otimes \mathfrak{s o}(2 m)\right)$ by:

$$
\widetilde{\omega}=\tau^{-1}\left(\rho^{*} \omega\right)
$$

Since pullbacks commute with exterior differentiation and wedge products, and $\tau$ is an isomorphism of Lie algebras, it follows that the curvature of $\widetilde{\omega}$ is related to the curvature $\Omega$ of $\omega$ by:

$$
\widetilde{\Omega}=\tau^{-1}\left(\rho^{*} \Omega\right)
$$

Now pulling back everything onto a trivialising neighbourhood $U$ for $P_{\text {spin }}$ (resp. $P_{S O}$ ) via a section $\sigma: U \rightarrow$ $P_{\text {spin } \mid U}$ (resp. $\left(\rho \circ \sigma: U \rightarrow P_{S O \mid U}\right)$, and using that $\tau\left(e_{i} e_{j}\right)=2\left(E_{j i}-E_{i j}\right)$, we have:

$$
\begin{aligned}
\Omega^{\text {spin }} & =\sigma^{*}\left(\tau^{-1} . \rho^{*} \Omega\right)=\tau^{-1}\left((\rho \circ \sigma)^{*} \Omega\right) \\
& =\tau^{-1}\left(\sum_{i, j} \Omega_{i j}^{S O} E_{i j}\right)=\tau^{-1}\left(\sum_{i<j} \Omega_{i j}^{S O}\left(E_{j i}-E_{i j}\right)\right)=\frac{1}{2} \sum_{i<j} \Omega_{i j}^{S O} e_{i} e_{j} \\
& =\frac{1}{4} \sum_{i \neq j} \Omega_{i j}^{S O} e_{i} e_{j}
\end{aligned}
$$

where $\Omega_{i j}^{S O}$ is the Cartan curvature 2-form on $U$ for $P_{S O}$.
In terms of the synchronous frame $\left\{e_{i}\right\}$ of $T^{*} M$ around a point $x$ one knows that the curvature form of the Riemannian connection is related to the Riemannian curvature tensor by:

$$
\Omega_{i j}^{S O}=-\sum_{k \neq l} R_{k l i j} e_{k} e_{l}=-R_{k l i j} e_{k} e_{l}
$$

where we have used the Einstein repeated summation convention. (The minus sign comes from the fact that the Riemannian connection of the principal bundle $P_{S O}$ of frames in the cotangent bundle $T^{*} M$ is the negative of that of the tangent bundle). It follows that:

$$
\begin{aligned}
\Omega^{\mathcal{S}} & =\Omega^{s p i n}=-\frac{1}{4} R_{k l i j} e_{i} e_{j} e_{k} e_{l} \\
& =\frac{1}{4}\left(e_{i} e_{j} e_{l} R_{k l i j}\right) e_{k} \text { since } R_{l l i j}=0 \text { and } e_{k} e_{l}=-e_{l} e_{k} \text { for } k \neq l
\end{aligned}
$$

If $i, j, l$ are distinct indices, $e_{i} e_{j} e_{l}=e_{l} e_{i} e_{j}=e_{j} e_{l} e_{i}$, and also by the Bianchi identity, $R_{k l i j}+R_{k j l i}+R_{k i j l}=0$. SO all such terms will drop out of the sum above. Terms with $i=j$ also vanish since $R_{k l i j}$ is antisymmetric in $i, j$. So the only terms remaining are those with $i=l \neq j$ and $i \neq l=j$. The sum becomes:

$$
\begin{aligned}
\Omega^{\mathcal{S}} & =\frac{1}{4}\left(e_{l} e_{j} e_{l} R_{k l l j}+e_{i} e_{l} e_{l} R_{k l i l}\right) e_{k}=\frac{1}{4}\left(e_{j} R_{k l l j}-e_{i} R_{k l i l}\right) e_{k} \\
& =\frac{1}{4} e_{j}\left(R_{k l l j}-R_{k l j l}\right) e_{k}=-\frac{1}{2}\left(R_{k l j l}\right) e_{j} e_{k} \\
& =-\frac{1}{2} R_{k j} e_{j} e_{k}=-\frac{1}{2} R_{i i} e_{i}^{2}=\frac{1}{2} k \quad \text { (since Ricci curvature } R_{i j} \text { is symmetric) }
\end{aligned}
$$

which proves the proposition by substituting into the Weitzenbock formula in Proposition 15.4.3.

Corollary 15.4.5 (Bochner-Lichnerowicz). Let $M$ be a compact spin manifold of dimension $2 m$, and with everywhere strictly positive scalar curvature. Then the kernel of the Dirac operator on $C^{\infty}(M, \mathcal{S}(M))$ is trivial. (That is, $M$ has no "harmonic spinors").

Proof: Let $s \in C^{\infty}(M, \mathcal{S}(M))$, with $D s=0$. By the Weitzenbock formula of Proposition 15.4.4, we have

$$
0=(D s, D s)_{M}=\left(D^{2} s, s\right)=\left(\nabla^{\mathcal{S} *} \nabla^{\mathcal{S}} s, s\right)+\frac{1}{4}(k s, s)_{M}
$$

If $s \neq 0$, the fact that $k>0$ everwhere implies the right hand side is strictly positive, and we have a contradiction.

Corollary 15.4.6 (Weitzenbock for a Dirac bundle on a spin manifold). Let $M$ be a spin manifold of dimension $2 m$, and let $\mathcal{E} \rightarrow M$ be a Dirac bundle on $M$ with Clifford connection $\nabla^{\mathcal{E}}$. By the Corollary 15.2.11, we have that $\mathcal{E}=\mathcal{S} \otimes_{\mathbb{C}} \mathcal{V}$, where $\mathcal{S} \rightarrow M$ is the spin bundle on $M$, and $\nabla^{\mathcal{E}}$ is the tensor product connection of the spin connection $\nabla^{\mathcal{S}}$ on $S$, and a unitary connection $\nabla^{\mathcal{V}}$ on $\mathcal{V}$. Then, for the Dirac operator $D^{\mathcal{E}}$ we have the Weitzenbock formula:

$$
\left(D^{\mathcal{E}}\right)^{2}=\nabla^{\mathcal{E} *} \nabla^{\mathcal{E}}+\frac{1}{4} k+R^{\mathcal{V}}
$$

where $R^{\mathcal{V}}$ is the curvature operator of $\mathcal{V}$ defined by

$$
R^{\mathcal{V}}(s \otimes \sigma)=\sum_{i<j} e_{i} \cdot e_{j} . s \otimes \Omega^{\mathcal{V}}\left(e_{i}, e_{j}\right) \sigma=\sum_{i<j} c\left(e_{i}\right) c\left(e_{j}\right) R\left(e_{i}, e_{j}\right)(s \otimes \sigma)
$$

and $k$ is the scalar curvature function of $M$.

Proof: By Proposition 15.4.3 above, we just have to compute the Clifford curvature $\Omega^{\mathcal{E}}$ in terms of the curvatures $\Omega^{\mathcal{S}}$ and $\Omega^{\mathcal{V}}$. First note that by definition of the tensor product connection:

$$
\nabla_{X}^{\mathcal{E}}(s \otimes \sigma)=\nabla_{X}^{\mathcal{S}} s \otimes \sigma+s \otimes \nabla_{X}^{\mathcal{V}} \sigma \quad \text { for all } X \in T_{x} M, s \in C^{\infty}(M, \mathcal{S}), \sigma \in C^{\infty}(M, \mathcal{V})
$$

For real vector fields $X, Y$ on $M$, the commutators

$$
\left[\nabla_{X}^{S} \otimes 1,1 \otimes \nabla_{Y}^{\mathcal{V}}\right]=0=\left[\nabla_{Y}^{S} \otimes 1,1 \otimes \nabla_{X}^{\mathcal{V}}\right]
$$

Hence

$$
\begin{aligned}
\Omega^{\mathcal{E}}(X, Y)(s \otimes \sigma) & =\left(\left[\nabla_{X}^{\mathcal{E}}, \nabla_{Y}^{\mathcal{E}}\right]-\nabla_{[X, Y]}^{\mathcal{E}}\right)(s \otimes \sigma) \\
& =\left(\left[\nabla_{X}^{\mathcal{S}}, \nabla_{Y}^{\mathcal{S}}\right] s\right) \otimes \sigma+s \otimes\left(\left[\nabla_{X}^{\mathcal{V}}, \nabla_{Y}^{\mathcal{V}}\right] \sigma\right)-\nabla_{[X, Y]}^{\mathcal{S}} s \otimes \sigma-s \otimes \nabla_{[X, Y]}^{\mathcal{V}} \sigma \\
& =\Omega^{S}(X, Y) s \otimes \sigma+s \otimes \Omega^{\mathcal{V}}(X, Y) \sigma
\end{aligned}
$$

Now, in a local orthonormal frame $\left\{e_{i}\right\}$ :

$$
\begin{aligned}
\Omega^{\mathcal{E}}(s \otimes \sigma) & =\sum_{i, j} c\left(e_{i}\right) c\left(e_{j}\right) \Omega^{\mathcal{E}}\left(e_{i}, e_{j}\right)(s \otimes \sigma) \\
& =\left(\sum_{i, j} e_{i} \cdot e_{j} \Omega_{i j}^{\mathcal{S}} s\right) \otimes \sigma+\sum_{i, j} e_{i} \cdot e_{j} \cdot s \otimes \Omega_{i j}^{\mathcal{V}} \sigma \\
& \left.=\frac{k}{2} s \otimes \sigma+2 \sum_{i<j}\left(e_{i} \cdot e_{j} \cdot s\right) \otimes \Omega_{i j}^{\mathcal{V}} \sigma\right)=\frac{k}{2} s \otimes \sigma+2 R^{\mathcal{V}}(s \otimes \sigma)
\end{aligned}
$$

by using the Bochner-Lichnerowicz formula for the spin bundle $\mathcal{S}$ deduced in Corollary 15.4.5. Our corollary now follows from the Weitzenbock formula 15.4.3.

Corollary 15.4.7 (Bochner's Theorem). Let $M$ be a compact oriented Riemannian manifold of dimension $2 m$. For the Dirac bundle $\Lambda_{\mathbb{C}}^{*} M \rightarrow M$, the Dirac operator $D$ is the operator $d+\delta$, (viz. the Dirac operator of the elliptic deRham complex). Furthermore:
(i): On 1-forms, we have the Bochner formula:

$$
\Delta \phi=\nabla^{T^{*} M *} \nabla^{T^{*} M} \phi+R \phi \quad \text { for } \phi \in \Lambda^{1}(M, \mathbb{C})
$$

where $R$ is the Ricci-curvature operator of $M$.
(ii): If $M$ has everwhere positive Ricci curvature (viz. $R$ is a real positive definite symmetric matrix at each point of $M$ ), then the first Betti number $\beta_{1}(M):=\operatorname{dim}_{\mathbb{C}} H^{1}(M, \mathbb{C})$ vanishes (that is $M$ has no nontrivial harmonic 1-forms).

Proof: By definition, in a local orthonormal frame $e_{i}$ of $T^{*} M$, we have:

$$
D=\sum_{i} c\left(e_{i}\right) \nabla_{i}
$$

where $\nabla_{i}=\nabla_{e_{i}}$ is with respect to the Levi-Civita connection. If we further assume the frame is synchronous at $x$, then $\nabla_{e_{i}, x}=\partial_{i, x}$. The operator $c\left(e_{i}\right)$ of Clifford multiplication by $e_{i}$ on the Clifford module $\Lambda_{\mathbb{C}}^{*} T^{*} M$ is given by (see Example 15.1.7):

$$
\left.c\left(e_{i}\right) \alpha=e_{i} \wedge \alpha-e_{i}\right\lrcorner \alpha \quad \alpha \in \Lambda^{*}(M, \mathbb{C})
$$

So, in a synchronous frame at $x$, the Dirac operator reads as:

$$
\left.D \alpha=\sum_{i} e_{i} \wedge \partial_{i, x} \alpha-\sum_{i} e_{i}\right\lrcorner \partial_{i, x} \alpha=d \alpha+\delta \alpha
$$

(Note the minus sign appears because the " $L^{2}$-adjoint of $\partial_{i}$ is $-\partial_{i}$ " from integration by parts.) This proves the first assertion.

To see the Bochner formula in (i), we appeal to the Weitzenbock formula from Proposition 15.4.3, and apply it to the Dirac bundle $\mathcal{E}:=\Lambda_{\mathbb{C}}^{*}\left(T^{*} M\right)$. We note that by the above, $D^{2}=(d+\delta)^{2}=d \delta+\delta d=\Delta$, the Laplace-Beltrami operator. For the right side, we need to compute the Clifford curvature $\frac{1}{2} \Omega^{\mathcal{E}}$. We continue with the synchronous frame above, and for the sake of convenience, we denote the operator $e_{k} \wedge(-)$ by $e_{k}$, and the operator $\left.e_{k}\right\lrcorner(-)$ by $i_{k}$. Note that:

$$
e_{k} i_{l}\left(e_{m}\right)+i_{l} e_{k}\left(e_{m}\right)=e_{k} \delta_{l m}+i_{l}\left(e_{k} \wedge e_{m}\right)=e_{k} \delta_{l m}+\delta_{k l} e_{m}-e_{k} \delta_{l m}=\delta_{k l} e_{m}
$$

so that $e_{k} i_{l}=-i_{l} e_{k}$ for $k \neq l$. Now, the Clifford curvature operator on a 1-form $\phi=\phi_{k} e_{k}$ (using the repeated summation convention) is given by:

$$
\begin{aligned}
\Omega^{\mathcal{E}} \phi & =c\left(e_{k}\right) c\left(e_{l}\right) \Omega^{\mathcal{E}}\left(e_{k}, e_{l}\right) \phi=-\left(e_{k}-i_{k}\right)\left(e_{l}-i_{l}\right) R_{k l r s} \phi_{r} e_{s} \\
& =\left(e_{k} i_{l}+i_{k} e_{l}\right) R_{k l r s} \phi_{r} e_{s}=\left(e_{k} i_{l}-i_{l} e_{k}\right) R_{k l r s} \phi_{r} e_{s} \quad\left(\text { since } R_{k l r s}=-R_{l k r s}\right) \\
& =2 e_{k} i_{l} R_{k l r s} \phi_{r} e_{s}=2 R_{k l r s} \phi_{r} e_{k} \delta_{l s}=2 R_{k l r l} \phi_{r} e_{k}=2\left(R_{k r} \phi_{r}\right) e_{k}=2 R \phi
\end{aligned}
$$

so that $\frac{1}{2} \Omega^{\mathcal{E}} \phi=R \phi$ and the Bochner formula (i) follows.
To see (ii), note that if $\phi \in \Lambda^{1}(M, \mathbb{C})$ is a harmonic form with $\phi \neq 0$, then $\Delta \phi=0$ so that by the Bochner formula:

$$
0=(\Delta \phi, \phi)=(\nabla \phi, \nabla \phi)+(R \phi, \phi)>0
$$

by the hypothesis on $R$, a contradiction. Now, by (i) of the Hodge Theorem,

$$
\beta_{1}(M)=\operatorname{dim}_{\mathbb{C}} H^{1}(M, \mathbb{C})=\operatorname{dim}_{\mathbb{C}} \operatorname{ker} \Delta_{\Lambda^{1}}
$$

so (ii) and the Corollary follow.

Corollary 15.4.8. Let $\mathcal{E} \rightarrow M$ be a Dirac bundle, with associated Dirac operator $D$. Then the operator (called the Dirac Laplacian of $\mathcal{E}$ ):

$$
D^{2}: C^{\infty}(M, \mathcal{E}) \rightarrow C^{\infty}(M, \mathcal{E})
$$

is a generalised laplacian in the sense of Definition 12.2.1. In particular, the operators $D^{-} D^{+}: C^{\infty}\left(M, \mathcal{E}^{+}\right) \rightarrow$ $C^{\infty}\left(M, \mathcal{E}^{+}\right)$and $D^{+} D^{-}: C^{\infty}\left(M, \mathcal{E}^{-}\right) \rightarrow C^{\infty}\left(M, \mathcal{E}^{-}\right)$are both generalised laplacians. The two term complexes $D^{ \pm}: C^{\infty}\left(M, \mathcal{E}^{ \pm}\right) \rightarrow C^{\infty}\left(M, \mathcal{E}^{\mp}\right)$ are both elliptic 2- term complexes in the sense of Definition 9.4.1, and the two operators $D^{+} D^{-}$and $D^{-} D^{+}$are the Dirac Laplacians of this 2-term elliptic complex.

Proof: By the Weitzenbock formula Proposition 15.4.3, we have:

$$
D^{2}=\nabla^{\mathcal{E} *} \nabla^{\mathcal{E}}+\frac{1}{2} \Omega^{\mathcal{E}}
$$

The last term $\frac{1}{2} \Omega^{\mathcal{E}}$ is a zero-th order operator, locally given as $\frac{1}{2} \sum_{i, j} e_{i} e_{j} . \Omega_{i j}^{\mathcal{E}}$, where each $\Omega_{i, j}$ is a 2 -form. In the proof of Weitzenbock's formula, we also computed the adjoint of $\nabla^{\mathcal{E}}$ in a synchronous frame $\left\{e_{i}\right\}$ at $x$ to be:

$$
\nabla^{\mathcal{E} *}\left(e_{i} \otimes s\right)(x)=-\nabla_{i} s
$$

Consider the tensor product of the Levi-Civita connection $\nabla$ and $\nabla^{\mathcal{E}}$, and call it the connection $\nabla^{T^{*} M \otimes \mathcal{E}}$. We claim that the composite:

$$
C^{\infty}\left(M, T^{*} M \otimes \mathcal{E}\right) \stackrel{\nabla^{T^{*} M \otimes \mathcal{E}}}{\rightarrow} C^{\infty}\left(M, T^{*} M \otimes T^{*} M \otimes \mathcal{E}\right) \xrightarrow{-\operatorname{tr}} C^{\infty}(M, \mathcal{E})
$$

is the same as

$$
\nabla^{\mathcal{E} *}: C^{\infty}\left(M, T^{*} M \otimes \mathcal{E}\right) \rightarrow C^{\infty}(M, \mathcal{E})
$$

(see Lemma 12.2.4). For, if we compute with the synchronous frame at $x$ used in the proof of the Weitzenbock formula in 15.4.3, we have:

$$
\nabla^{T^{*} M \otimes \mathcal{E}}\left(e_{i}^{*} \otimes s\right)(x)=\nabla e_{i}^{*} \otimes s+e_{i}^{*} \otimes \nabla^{\mathcal{E}} s=e_{i}^{*} \otimes \sum_{j} e_{j}^{*} \nabla_{j}^{\mathcal{E}} s
$$

since $\left(\nabla e_{i}\right)(x)=0$. Thus:

$$
-\operatorname{tr} \nabla^{T^{*} M \otimes \mathcal{E}}\left(e_{i}^{*} \otimes s\right)(x)=-\operatorname{tr}\left(\sum_{j} e_{i}^{*} \otimes e_{j}^{*} \nabla_{j}^{\mathcal{E}} s\right)(x)=-\left(\sum_{j} \delta_{i j} \nabla_{j}^{\mathcal{E}} s\right)(x)=-\left(\nabla_{i}^{\mathcal{E}} s\right)(x)
$$

We computed in the proof of Weitzenbock that:

$$
\nabla^{\mathcal{E} *}\left(e_{i}^{*} \otimes s\right)(x)=-\left(\nabla_{i}^{\mathcal{E}} s\right)(x)
$$

Hence our assertion follows. Thus, in the notation of Lemma 12.2.4,

$$
\nabla^{\mathcal{E} *} \nabla^{\mathcal{E}}=-\operatorname{tr} \nabla^{T^{*} M \otimes \mathcal{E}} \nabla^{\mathcal{E}}=\Delta^{\mathcal{E}}
$$

By the Lemma $12.2 .4, \Delta^{\mathcal{E}}$ is a generalised laplacian. Hence $D^{2}=\Delta^{E}+\frac{1}{2} \Omega^{\mathcal{E}}$ is also a generalised laplacian. Since $D^{+} D^{-}$and $D^{-} D^{+}$are restrictions of $D^{2}$ to $C^{\infty}\left(M, \mathcal{E}^{-}\right)$and $C^{\infty}\left(M, \mathcal{E}^{+}\right)$respectively, they are also elliptic second order differential operators.

Since $D^{+}$and $D^{-}$are adjoints of each other by Proposition 15.3.3, the 2-term complex is an elliptic omplex by 9.4.2.

## 16. The Atiyah-Singer Index Theorem

The goal now is to write down a formula for the index of a Dirac operator on a Dirac bundle. The idea of the proof of the index theorem is to carefully examine the coefficient of the term independent of $t$ in the asymptotic expansion of the (super)trace of heat kernel for the Dirac laplacian, since integrating this over $M$ would compute the index of $D$, in view of Proposition 10.2.1. To handle the Dirac operator by bare hands is quite an effort, and was carried out by Patodi, Atiyah-Bott-Patodi and Gilkey for the various classical Dirac bundles. There is however a simple proof due to Getzler, following ideas of the physicists Alvarez-Gaume and Witten, which replaces the Dirac laplacian by a much simpler operator by a scaling procedure.

Before we get into the proof of the index theorem, let us study this simpler operator.

### 16.1. The Quantum Harmonic Oscillator and Mehler's Formula.

Definition 16.1.1. The quantum harmonic oscillator is the Schrodinger operator defined on $C^{\infty}(\mathbb{R})$ by:

$$
H:=-\frac{d^{2}}{d x^{2}}+x^{2}
$$

Proposition 16.1.2 (Facts about the Harmonic Oscillator). $H$ defined above is formally self-adjoint (on compactly supported functions), and has a discrete positive spectrum $\lambda_{n}=\left(n+\frac{1}{2}\right)$, corresponding to smooth eigenfunctions $\phi_{n}$ in the Schwartz class $\mathcal{S}(\mathbb{R})$ (which are defined in terms of the Hermite functions). Finally, $\phi_{n}$ form an orthonormal Hilbert space basis for $L_{2}(\mathbb{R})$.

Proof: The formal self-adjointness on $C_{c}(\mathbb{R})$ is clear since $H=\Delta+x^{2}$, and both operators on the right are formally self- adjoint. Also since

$$
(H \phi, \phi)=\left(\partial_{x} \phi, \partial_{x} \phi\right)+(x \phi, x \phi)
$$

for the $L_{2}(\mathbb{R})$ inner-product $(-,-)$, it follows that the eigenvalues (if any) of $H$ are non-negative.

To get the discreteness of the spectrum, one uses the annihilation operator $A=x+\partial_{x}$ and its adjoint, the creation operator $A^{*}=x-\partial_{x}$. It is easily checked that:

$$
\begin{aligned}
A A^{*} & =H+I, \quad A^{*} A=H-I \\
{\left[A, A^{*}\right] } & =-2 I \\
{[H, A] } & =-2 A, \quad\left[H, A^{*}\right]=2 A^{*}
\end{aligned}
$$

Then one defines the ground state of the oscillator as the function $\phi_{0}$ satisfying $A \phi_{0}=0$ and $\left\|\phi_{0}\right\|=1$. That is,

$$
\left(\partial_{x}+x\right) \phi=0
$$

But this is a simple ODE, and by using the integrating factor of $e^{x^{2} / 2}$, and using the $L_{2}$-normalisation, we have

$$
\phi_{0}=\pi^{-1 / 4} e^{-x^{2} / 2}
$$

Now all the other eigenfunctions are given inductively by applying the creation operator $A^{*}$ and normalising. More precisely:

$$
\phi_{k}=(2 k)^{-1 / 2} A^{*} \phi_{k-1}
$$

Note that if $\phi_{k-1}$ corresponds to eigenvalue $\lambda_{k-1}$, then

$$
\begin{aligned}
H \phi_{k} & =(2 k)^{-1 / 2} H A^{*} \phi_{k-1}=(2 k)^{-1 / 2}\left(A^{*} \lambda_{k-1} \phi_{k-1}+\left[H, A^{*}\right] \phi_{k-1}\right) \\
& =\left(\lambda_{k-1}+2\right) \phi_{k}
\end{aligned}
$$

So it remains to compute the eigenvalue of the ground state $\phi_{0}$. But

$$
\left(-\partial_{x}^{2}+x^{2}\right) e^{-x^{2} / 2}=\partial_{x}\left(x e^{-x^{2}}\right)+x^{2} e^{-x^{2} / 2}=e^{-x^{2} / 2}
$$

So $H \phi_{0}=\phi_{0}$, and $\lambda_{0}=1$. This shows that $\lambda_{k}=(2 k+1)$. Since $\phi_{0}$ is is the Schwartz class, so is $\phi_{k}=C\left(A^{*}\right)^{k} \phi_{0}$.
We will skip the proof of the fact that $\phi_{k}$ form an orthonormal basis of $L_{2}(\mathbb{R})$. See standard texts on Quantum Mechanics, which prove that $\phi_{k}$ is a polynomial times the Hermite function $H_{k}$.

Corollary 16.1.3. The associated heat operator $\left(H+\partial_{t}\right)$ on $\mathbb{R}$ has a smooth integral kernel $p_{t}(x, y)$ which satisfies:
(i): $\left(H_{x}+\partial_{t}\right) p_{t}(x, y)=0$ for all $t>0$, and $x, y \in \mathbb{R}$.
(ii): For all $\phi \in L_{2}(\mathbb{R})$, and $t>0$ the function $F(x, t):=e^{-t H} \phi$ is a smooth function, given by the integral $\int_{\mathbb{R}} p_{t}(x, y) \phi(y) d y$. It satisfies:

$$
\lim _{t \rightarrow 0} F(x, t)=\lim _{t \rightarrow 0} \int_{\mathbb{R}} p_{t}(x, y) \phi(y) d y=\phi(x)
$$

Proof: Follows by defining :

$$
p_{t}(x, y)=\sum_{k=0}^{\infty} e^{-t \lambda_{k}} \phi_{k}(x) \phi_{k}(y)
$$

which is a convergent series for all $t>0$ since the coefficients $e^{-t \lambda_{k}}$ die faster than all powers of $t$, and $\phi_{k}$ are in the Schwartz class. The proof of the other assertions are analogous to the case of a positive elliptic operator on a compact manifold (see (iii) of the Proposition 10.1.3 and (iii) of Proposition 10.1.6).

Now we can explicitly compute $u(x, t)=p_{t}(x, 0)$. Note that by definition, this is a fundamental solution to the heat equation, satisfying:

$$
\left(H+\partial_{t}\right) u(x, t)=0, \quad \lim _{t \rightarrow 0} u(x, t)=\delta_{x}
$$

where $\delta_{x}$ is the Dirac distribution massed at $x$.

Proposition 16.1.4 (Mehler's Formula). The function $u(x, t)$ defined by:

$$
u(x, t)=(2 \pi \sinh 2 t)^{-1 / 2} \exp \left(-\frac{x^{2} \operatorname{coth} 2 t}{2}\right)
$$

is a fundamental solution to the heat equation $\left(H+\partial_{t}\right) u(x, t)=0$.

Proof: By taking one's cue from the Gaussian, we try:

$$
u(x, t)=\alpha(t) \exp \left(-\frac{\beta(t) x^{2}}{2}\right)
$$

Then compute derivatives:

$$
\begin{aligned}
-\partial_{x} u(x, t) & =\alpha \beta x \exp \left(-\frac{\beta x^{2}}{2}\right) \\
-\partial_{x}^{2} u(x, t) & =\alpha \beta\left(1-\beta x^{2}\right) \exp \left(-\frac{\beta x^{2}}{2}\right)=\beta\left(1-\beta x^{2}\right) u(x, t) \\
H u(x, t) & =\left(-\partial_{x}^{2}+x^{2}\right) u(x, t)=\left(\beta+\left(1-\beta^{2}\right) x^{2}\right) u(x, t) \\
\partial_{t} u(x, t) & =\left[\alpha^{\prime}(t)-\frac{\alpha \beta^{\prime}(t) x^{2}}{2}\right] \exp \left(-\frac{\beta x^{2}}{2}\right)=\left(\frac{\alpha^{\prime}(t)}{\alpha(t)}-\frac{\beta^{\prime}(t) x^{2}}{2}\right) u(x, t)
\end{aligned}
$$

So if we arrange that:

$$
\frac{\alpha^{\prime}}{\alpha}+\beta=0, \quad \beta^{\prime} / 2=\left(1-\beta^{2}\right)
$$

Then $u(x, t)$ would be a solution to the required heat equation $\left(H+\partial_{t}\right) u(x, t)=0$. The second equation leads to:

$$
\left(\frac{1}{\beta+1}-\frac{1}{\beta-1}\right) d \beta=4 d t
$$

so that $\log \left(\frac{\beta+1}{\beta-1}\right)=4 t+C$, which implies (by taking $C=0$ ) that

$$
\beta(t)=\operatorname{coth} 2 t
$$

The other equation now becomes:

$$
\alpha^{\prime}(t)+\alpha(t) \operatorname{coth} 2 t=0
$$

which is the same as: $\sinh 2 t \alpha^{\prime}(t)+\cosh 2 t \alpha(t)=0$, which is rewritten as $2 \sinh 2 t \alpha^{\prime}(t) \alpha(t)+2 \cosh 2 t \alpha^{2}(t)=0$. But this implies:

$$
\frac{d}{d t}\left(\alpha(t)^{2} \sinh 2 t\right)=0
$$

Thus $\alpha(t)=C(\sinh 2 t)^{-1 / 2}$, for some constant $C$. Now as $t \rightarrow 0, \alpha(t) \sim C(2 t)^{-1 / 2}$ and $\beta(t) \sim 1 / 2 t$, so that $u(x, t) \sim C(2 t)^{-1 / 2} e^{-x^{2} / 4 t}$ as $t \rightarrow 0$. We choose $C=(2 \pi)^{-1 / 2}$, so that $u(x, t)$ approaches the Euclidean heat kernel as $t \rightarrow 0$. Hence:

$$
u(x, t)=(2 \pi \sinh 2 t)^{-1 / 2} \exp \left(-\frac{\operatorname{coth} 2 t x^{2}}{2}\right)
$$

and we certainly have $\left(H+\partial_{t}\right) u(x, t)=0$. Also, as $t \rightarrow 0, u(x, t) \rightarrow(4 \pi t)^{-1 / 2} \exp \left(-x^{2} / 4 t\right)$, the Euclidean heat kernel on $\mathbb{R}$, and we know by the Proposition 11.1.1 that the Euclidean heat kernel tends to $\delta_{x}$ as $t \rightarrow 0$. (Actually, more precise estimates are needed to justify these limits in the space of tempered distributions $\mathcal{S}^{\prime}(\mathbb{R})$, but we leave these details to the reader). The proposition follows.

Corollary 16.1.5 (Mehler's Formula II). Let us define the 2 nd order differential operator on $C^{\infty}(\mathbb{R})$ given by:

$$
H:=-\partial_{x}^{2}+\frac{a^{2} x^{2}}{16}+b
$$

where $a, b \in \mathbb{R}$. Then a fundamental solution to $\left(H+\partial_{t}\right) v(x, t)=0$ is given by:

$$
v(x, t)=(4 \pi t)^{-1 / 2}\left(\frac{a t / 2}{\sinh (a t / 2)}\right)^{\frac{1}{2}} \exp \left(-(a t / 2) \operatorname{coth}(a t / 2)\left(x^{2} / 4 t\right)-b t\right)
$$

Proof: First we try to find a solution to $\left(H+\partial_{t}\right) v(x, t)=0$ by tinkering with the fundamental solution of the foregoing proposition. So let $u(y, s)$ be the fundamental solution satisfying:

$$
\begin{equation*}
\left(-\partial_{y}^{2}+y^{2}+\partial_{s}\right) u(y, s)=0 \tag{48}
\end{equation*}
$$

where $u(y, s)$ is as in the statement of Proposition 16.1.4 above. Define:

$$
v(x, t)=e^{-b t} u\left(\lambda^{1 / 2} x, \lambda t\right)=e^{-b t} u(y, s) \quad \text { where } \quad y:=\lambda^{1 / 2} x, \quad s:=\lambda t
$$

Then:

$$
\partial_{t} v(x, t)=-b e^{-b t} u(y, s)+e^{-b t} \lambda \partial_{s} u(y, s)
$$

which implies

$$
\begin{equation*}
\left(\partial_{t}+b\right) v(x, t)=\lambda e^{-b t} \partial_{s} u(y, s) \tag{49}
\end{equation*}
$$

Now for the space derivatives:

$$
\begin{align*}
\left(-\partial_{x}^{2}+\lambda^{2} x^{2}\right) v(x, t) & =\lambda\left(-\frac{1}{\lambda} \partial_{x}^{2}+\lambda x^{2}\right) v(x, t) \\
& =\lambda\left(-\partial_{y}^{2}+y^{2}\right) e^{-b t} u(y, s)=\lambda e^{-b t}\left(-\partial_{y}^{2}+y^{2}\right) u(y, s) \tag{50}
\end{align*}
$$

Adding the equations (49) and (50), we find:

$$
\left(-\partial_{x}^{2}+\lambda^{2} x^{2}+b+\partial_{t}\right) v(x, t)=\lambda e^{-b t}\left(-\partial_{y}^{2}+y^{2}+\partial_{s}\right) u(y, s)=0
$$

from equation (48). Thus $v(x, t)$ is a solution to the equation in the statement by setting $\lambda=a / 4$. Thus by using the explicit formula for $u(y, s)$ derived in Proposition 16.1.4, the fundamental solution we seek is given by:

$$
\begin{aligned}
v(x, t) & =C e^{-b t} u\left(\frac{a^{1 / 2} x}{2}, a t / 4\right)=C e^{-b t}(2 \pi \sinh (a t / 2))^{-1 / 2} \exp \left[-\operatorname{coth}(a t / 2)\left(a x^{2} / 8\right)\right] \\
& =\widetilde{C}(4 \pi t)^{-1 / 2}(a t / 2)^{1 / 2}(\sinh (a t / 2))^{-1 / 2} \exp \left[-(a t / 2) \operatorname{coth}(a t / 2)\left(x^{2} / 4 t\right)-b t\right]
\end{aligned}
$$

Note that as $\lim _{t \rightarrow 0} \frac{\sinh (a t / 2)}{(a t / 2)}=1, \lim _{t \rightarrow 0} b t=0$ and $\lim _{t \rightarrow 0} \cosh (a t / 2)=1$, which implies that $\lim _{t \rightarrow 0}(a t / 2) \operatorname{coth}(a t / 2)=1$. Thus $v(x, t)$ above approaches the Euclidean heat kernel $(4 \pi t)^{-1 / 2} \exp \left(-x^{2} / 4 t\right)$ if we set $\widetilde{C}=1$ (Again pointwise limits are not good enough, one needs to use Lebesgue's dominated convergence theorem to make these assertions in $\mathcal{S}^{\prime}(\mathbb{R})$. We leave these matters to the reader.) This proves the corollary.

We would like to write a multivariate and matrix formulation of the above Mehler formula. First we make a definition.

Definition 16.1.6. Denote by $\mathcal{A}$ the commutative algebra $\Lambda_{\mathbb{C}}^{e v}\left(\mathbb{R}^{2 m}\right)$, with $\wedge$ being the multiplication. Note that any word of $a_{1} a_{2} \ldots a_{i}$ of length $i>m$ and with $a_{j} \in \oplus_{k \geq 1} \Lambda_{\mathbb{C}}^{2 k}$ (i.e. no $a_{j}$ has a constant term) vanishes.

Let $R$ be a skew-symmetric $2 m \times 2 m$ matrix whose entries are in $\Lambda_{\mathbb{C}}^{2}\left(\mathbb{R}^{2 m}\right)$. Note that $R$ is automatically a nilpotent matrix, by the remark above. Hence all power series in $t R$ for $t \in(0, \infty)$ are actually polynomials in $t$. For such an $\mathbb{R}$, define the $\mathcal{A}$-valued function:

$$
j(R):=\operatorname{det}\left(\frac{e^{R / 2}-e^{-R / 2}}{R}\right)=\operatorname{det}\left(\frac{\sinh (R / 2)}{(R / 2)}\right)
$$

Note that $e^{R / 2}-e^{-R / 2}$ is the series (=polynomial) (of $\mathcal{A}$-valued matrices) given by $2 \sinh (R / 2)$, and involving only odd powers of $R$, whence $\frac{1}{2}\left(e^{R / 2}-e^{-R / 2}\right)=\sinh (R / 2)=(R / 2) \alpha(R)$, for another polynomial $\alpha(R)$ with leading coefficient 1 . Then $j(R)$ is the determinant of $\alpha(R)$.

Indeed, if $t$ is small enough, the series:

$$
\alpha(t R)=I+\frac{t^{2} R^{2}}{2^{2} 3!}+\frac{t^{4} R^{4}}{2^{4} 5!}+\ldots
$$

is a polynomial, and an invertible element, since every term except the first is nilpotent. Its determinant $j(t R):=\operatorname{det}(\alpha(t R))$ is a unit in $\mathcal{A}$, and again a polynomial in $t$. Then one can define $j(t R)^{-1 / 2}=$ $[\operatorname{det}(\alpha(t R))]^{-1 / 2}$ as a formal power series

$$
(j(t R))^{-1 / 2}=1+\sum_{i=1}^{\infty} t^{i} f_{i}(R)
$$

where $f_{i}$ are polynomials in the entries of $R$. This formal power series is again a polynomial in $t$ since $j(t R)-I$ is a nilpotent element.

Now we consider the symmetric matrix

$$
\beta(t R):=(t R / 2) \operatorname{coth}(t R / 2)
$$

Then it defines an $\mathcal{A}$-valued symmetric bilinear form (or quadratic form) on $\mathbb{R}^{2 m}$ by the formula:

$$
\langle x|(t R / 2) \operatorname{coth}(t R / 2)|y\rangle:=\sum_{i, j=1}^{2 m} x_{i}(t R / 2) \operatorname{coth}(t R / 2)_{i j} y_{j}
$$

Again, we have a power series expansion for $(t R / 2) \operatorname{coth}(t R / 2)$ in terms of even powers of $t$, which starts with $I$ (since $\cosh (t R / 2)$, and $\frac{t R / 2}{\sinh (t R / 2)}=\alpha(t R)^{-1}$ both have even power series starting with $I$ ). So the quadratic form above has a power series expansion:

$$
\langle x|(t R / 2) \operatorname{coth}(t R / 2)|x\rangle=\|x\|^{2}+\sum_{k=1}^{\infty} t^{2 k} c_{k}\langle x| R^{2 k}|x\rangle
$$

Since $R$ is nilpotent, this power series is again a polynomial.

Proposition 16.1.7 (Mehler's Formula III). Let $R$ be a skew-symmetric $2 m \times 2 m$ matrix, and let $F$ be any $N \times N$ matrix, both matrices having coefficients in $\Lambda_{\mathbb{C}}^{2}\left(\mathbb{R}^{2 m}\right)$. Set $\mathcal{A}:=\Lambda_{\mathbb{C}}^{e v}\left(\mathbb{R}^{2 m}\right)$. Note both matrices are constant with respect to $x \in \mathbb{R}^{2 m}$.

Define the generalised harmonic oscillator to be the operator defined on $C^{\infty}\left(\mathbb{R}^{2 m}, \mathcal{A} \otimes \operatorname{End}_{\mathbb{C}}\left(\mathbb{C}^{N}\right)\right)$ by:

$$
H(f \otimes T)=-\left[\sum_{i=1}^{2 m}\left(\partial_{i}+\frac{1}{4} \sum_{j} R_{i j} x_{j}\right)^{2} f\right] \otimes T+f \otimes F . T \quad \text { for } f \in C^{\infty}\left(\mathbb{R}^{n}, \mathcal{A}\right), T \in \operatorname{End}_{\mathbb{C}}\left(\mathbb{C}^{N}\right)
$$

Then the associated heat operator $\left(H+\partial_{t}\right)$ has a fundamental solution $p_{t}(x, R, F)$ defined by:

$$
p_{t}(x, R, F)=(4 \pi t)^{-m} j(t R)^{-1 / 2} \exp \left(\frac{-1}{4 t}\langle x|(t R / 2) \operatorname{coth}(t R / 2)|x\rangle\right) \exp (-t F)
$$

with $\lim _{t \rightarrow 0} p_{t}(x, R, F)=\delta_{x}(1 \otimes I d)$. (Note that by the discussion in Definition 16.1.6 above, $j(t R)$, $\langle x|(t R / 2) \operatorname{coth}(t R / 2)|x\rangle$ and $\exp (-t F)$ are all polynomials in $t$, by the nilpotency of the matrices $R$ and $F)$.

Proof: As remarked in the discussion following Definition 16.1.6, the power $(j(t R))^{-1 / 2},\langle x|(t R / 2) \operatorname{coth}(t R / 2)|s\rangle$, and $\exp (-t F)$ are all polynomials in $t$, whose coefficients are polynomials in the coefficients of $R$ and $F$, because the entries of $R$ and $F$ are in $\Lambda_{\mathbb{C}}^{2}\left(\mathbb{R}^{2 m}\right)$, and any word $a_{1} a_{2} . . a_{i}$ of length $i>m$ with $a_{i} \in \oplus_{i \geq 1} \Lambda_{\mathbb{C}}^{2 i}\left(\mathbb{R}^{2 m}\right)$ vanishes. Suppose we verify the formula:

$$
\left(H+\partial_{t}\right) p_{t}(x, R, F)=0
$$

for $R$ and $F$ matrices with real entries. Since $\left(H+\partial_{t}\right) p_{t}(x, R, F)$ is an analytic function of $F_{i j}$ and $R_{i j}$, it will follow that the equation $\left(H+\partial_{t}\right) p_{t}(x, R, F)=0$ for all $F$ with complex entries $F_{i j}$ and all skew-symmetric matrices $R$ with complex entries $R_{i j}$. That is, we will have an identity of power series in $R_{i j}$ and $F_{i j}$. Hence this identity will hold when we substitute $F$ a nilpotent matrix with entries in $\Lambda_{\mathbb{C}}^{2}\left(\mathbb{R}^{2 m}\right)$, and $R$ an antisymmetric matrix with entries in $\Lambda_{\mathbb{C}}^{2}\left(\mathbb{R}^{2 m}\right)$. So we may assume without loss of generality that $R$ is a real antisymmetric matrix, and $F$ is a matrix with real entries. Note that all the power series (in $t$ ) occurring above in the expression for $p_{t}(x, R, F)$ converge for small values of $t$ at least.

Note that $R \in \Lambda^{2}\left(\mathbb{R}^{2 m}\right)=\mathfrak{s o}(2 m)$, and there is a matrix $P \in S O(2 m)$ which conjugates $R$ into the Cartan subalagebra of $\mathfrak{s o}(2 m)$. That is, there is a change of orthonormal basis (given by $P$ ) for $\mathbb{R}^{2 m}$ so that $P R P^{t}=S$ is in block diagonal form, where the $i$-th block of $S$ is:

$$
S^{i}:=\left(\begin{array}{cc}
0 & -a_{i} \\
a_{i} & 0
\end{array}\right)
$$

Since $j(t R)$ is the determinant of $\alpha(t R)$, we will have $j(t R)=j(t S)$. Setting $y=P x$, we find that the quadratic form:

$$
\begin{aligned}
\langle x|(t R / 2) \operatorname{coth}(t R / 2)|x\rangle & =\langle(t R / 2) \operatorname{coth}(t R / 2) x, x\rangle=\langle P(t R / 2) \operatorname{coth}(t R / 2) x, P x\rangle \\
& =\langle(t S / 2) \operatorname{coth}(t S / 2) P x, P x\rangle=\langle y|(t S / 2) \operatorname{coth}(t S / 2)|y\rangle
\end{aligned}
$$

Finally, note that

$$
\begin{aligned}
\frac{\partial}{\partial x_{i}}+\frac{1}{4} \sum_{j} R_{i j} x_{j} & =\sum_{j} \frac{\partial y_{j}}{\partial x_{i}} \frac{\partial}{\partial y_{j}}+\frac{1}{4}(R x)_{i}=\sum_{j} P_{j i} \partial_{y, j}+\frac{1}{4}\left(P^{t} S P x\right)_{i} \\
& =\left[P^{t}\left(\partial_{y}+\frac{1}{4} S y\right)\right]_{i}
\end{aligned}
$$

which implies that the "norm" of the "vector $\left(\partial_{x}+\frac{1}{4} R x\right)$ is the same as that of $\left(\partial_{y}+\frac{1}{4} S y\right)$, that is:

$$
\sum_{i}\left(\frac{\partial}{\partial x_{i}}+\frac{1}{4} \sum_{j} R_{i j} x_{j}\right)^{2}=\sum_{i}\left(\frac{\partial}{\partial y_{i}}+\frac{1}{4} \sum_{j} S_{i j} y_{j}\right)^{2}
$$

Of course $F$ will not change, so under the change of variables $x \mapsto y=P x$, the form of the operator $H$ will remain the same, with $R$ replaced by $S$ and $x$ replaced by $y$. Hence proving that $\left(H+\partial_{t}\right) p_{t}(x, R, F)=0$ is equivalent to proving that $\left(H+\partial_{t}\right) p_{t}(y, S, F)$. Hence we may assume without loss of generality that $R$ is in block diagonal form.

But once $R$ is in block diagonal form, we are reduced to showing the identity for $m=1$. Indeed, defining the $2 \times 2$ block operator:

$$
H^{i}=-\left(\partial_{2 i-1}-a_{i} x_{2 i}\right)^{2}-\left(\partial_{2 i}+a_{i} x_{2 i-1}\right)^{2}+\frac{1}{m} F \quad \text { for } i=1,2, . ., m
$$

and denoting $x^{i}:=\left(x_{2 i-1}, x_{2 i}\right)$, and its fundamental solution by $p_{t}^{i}\left(x^{i}, S^{i}, \frac{1}{m} F\right)$, we note that $p_{t}(x, S, F)=$ $\prod_{i=1}^{m} p_{t}^{i}\left(x^{i}, S^{i}, \frac{1}{m} F\right)$ obeys the equation

$$
H p_{t}=\sum_{i=1}^{m}\left(H^{i} p_{t}\right)=\sum_{i=1}^{m}\left(p_{t}^{1} \ldots \widehat{p_{t}^{i}} . . p_{t}^{m}\right) H^{i} p_{t}^{i}=-\sum_{i=1}^{m}\left(p_{t}^{1} \ldots \widehat{p^{i}}{ }_{t} . . p_{t}^{m}\right) \partial_{i} p_{t}^{i}=-\partial_{t} p_{t}
$$

Also, as $t \rightarrow 0$, each $p^{i} \rightarrow \delta_{x_{2 i-1}, x_{2 i}}$, and so $p \rightarrow \delta_{x}$, since the Dirac distribution in several variables is the product of the Dirac distributions in each variable. Thus we need to only find the two variable solution $p_{t}^{i}$.

Also note that the expression on the right, viz.

$$
(4 \pi t)^{-m} j(t S)^{-1 / 2} \exp \left(\frac{-1}{4 t}\langle x|(t S / 2) \operatorname{coth}(t S / 2)|x\rangle-t F\right)
$$

is exactly the expression:

$$
=\prod_{i=1}^{m}\left((4 \pi t)^{-1} j\left(t S^{i}\right)^{-1 / 2} \exp \left(\frac{-1}{4 t}\left\langle x^{i}\right|\left(t S^{i} / 2\right) \operatorname{coth}\left(t S^{i} / 2\right)\left|x^{i}\right\rangle-t F / m\right)\right)
$$

since determinants (like $j(t S)$ ) are mulplicative with respect to direct sum of $(2 \times 2)$-blocks, and quadratic forms are additive.

Thus we may as well assume that we are in $\mathbb{R}^{2}$. That is, $m=1$, and

$$
R=\left(\begin{array}{cc}
0 & -a \\
a & 0
\end{array}\right)
$$

In this event:

$$
\begin{align*}
H & =-\left(\partial_{1}-\frac{1}{4} a x_{2}\right)^{2}-\left(\partial_{2}+\frac{1}{4} a x_{1}\right)^{2}+F \\
& =-\left(\partial_{1}^{2}+\partial_{2}^{2}\right)+\frac{a}{2}\left(x_{2} \partial_{1}-x_{1} \partial_{2}\right)-\frac{a^{2}}{16}\|x\|^{2}+F \\
& =\left(-\partial_{1}^{2}+\left(\frac{i a}{4}\right)^{2} x_{1}^{2}+\frac{F}{2}\right)+\left(-\partial_{2}^{2}+\left(\frac{i a}{4}\right)^{2} x_{2}^{2}+\frac{F}{2}\right)+\frac{a}{2}\left(x_{2} \partial_{1}-x_{1} \partial_{2}\right) \tag{51}
\end{align*}
$$

On the other hand, by diagonalising $R$ over $\mathbb{C}$, we have:

$$
j(t R)=\operatorname{det}\left(\frac{e^{t R / 2}-e^{-t R / 2}}{t R}\right)=\left(\frac{\sinh (i a t / 2)}{(i a t / 2)}\right)\left(\frac{\sinh (-i a t / 2)}{(-i a t / 2)}\right)=\left(\frac{\sinh (i a t / 2)}{(i a t / 2)}\right)^{2}
$$

Similarly, the quadratic form:

$$
\langle x|(t R / 2) \operatorname{coth}(t R / 2)|x\rangle=(i a t / 2) \operatorname{coth}(i a t / 2) x_{1}^{2}+(-i a t / 2) \operatorname{coth}(-i a t / 2) x_{2}^{2}=(i a t / 2) \operatorname{coth}(i a t / 2)\left(x_{1}^{2}+x_{2}^{2}\right)
$$

So the function:

$$
\begin{aligned}
p_{t}\left(x_{1}, x_{2}\right): & =(4 \pi t)^{-1} j(t R)^{-1 / 2} \exp \left(-\frac{1}{4 t}\langle x|(t R / 2) \operatorname{coth}(t R / 2)|x\rangle-t F\right) \\
& =\prod_{j=1}^{2}\left[(4 \pi t)^{-1 / 2}\left(\frac{i a t / 2}{\sinh (i a t / 2)}\right)^{1 / 2} \exp \left(-(i a t / 2) \operatorname{coth}(i a t / 2) \frac{x_{j}^{2}}{4 t}-t F / 2\right)\right]
\end{aligned}
$$

is a fundamental solution for the operator

$$
\left(-\partial_{1}^{2}+(i a / 4)^{2} x_{1}^{2}+F / 2\right)+\left(-\partial_{2}^{2}+(i a / 4)^{2} x_{1}^{2}+F / 2\right)
$$

by the Corollary 16.1.5. (We have to soup up that Corollary to include all complex $a$, but that is straightforward). Also, since the function $p_{t}\left(x_{1}, x_{2}\right)$ is a function only of $\left(x_{1}^{2}+x_{2}^{2}\right)$ in the space variables, it is annihilated by the operator $\left(x_{2} \partial_{1}-x_{1} \partial_{2}\right)$. Hence it is a fundamental solution of $H$ in equation (51). This proves the proposition.

### 16.2. The Heat Kernel and Index Density.

Proposition 16.2.1. Let $\mathcal{E} \rightarrow M$ be a Dirac bundle on a compact Riemannian oriented manifold of dimension $2 m$. Then for the two term elliptic complex:

$$
\operatorname{Str} D:=\operatorname{ind}\left(D^{+}\right)=\operatorname{dim} \operatorname{ker} D^{+}-\operatorname{dim} \operatorname{ker} D^{-}=\int_{M} \operatorname{str} k_{t}(x, x) d V(x)=-\operatorname{ind} D^{-}
$$

where $k_{t}(x, x)$ is a self-adjoint endomorphism of $\mathcal{E}_{x}$ that maps $\mathcal{E}_{x}^{ \pm}$to $\mathcal{E}_{x}^{ \pm}$. Indeed, $k_{t}(x, y)$ is the integral kernel which represents the heat-operator $e^{-t D^{2}}$ for the Dirac laplacian $D^{2}$. (Note that for a endomorphism $T \in \operatorname{hom}_{\mathbb{C}}\left(\mathcal{E}_{x}, \mathcal{E}_{x}\right)$, which preserves the grading, we define the supertrace as in Definition 14.5.1, i.e. $\operatorname{str} T=$ $\left.\operatorname{tr} T^{+}-\operatorname{tr} T^{-}\right)$.

Proof: By the Corollary 15.4.8, the two term complex:

$$
D^{+}: C^{\infty}\left(M, \mathcal{E}^{+}\right) \rightarrow C^{\infty}\left(M, \mathcal{E}^{-}\right)
$$

is an elliptic complex. By (iii) of the Proposition 10.1.3, the infinitely smoothing heat operators $e^{-t \Delta^{ \pm}}=$ $e^{-t D^{\mp} D^{ \pm}}$on $C^{\infty}\left(M, \mathcal{E}^{ \pm}\right)$have integral heat kernels:

$$
k_{t}^{ \pm}(x, y) \in C^{\infty}\left(M \times M, \operatorname{hom}_{\mathbb{C}}\left(\pi_{2}^{*} \mathcal{E}^{ \pm}, \pi_{1}^{*} \mathcal{E}^{ \pm}\right)\right)
$$

By the McKean-Singer Theorem 10.2.1, we have:

$$
\operatorname{ind}\left(D^{+}\right)=\int_{M}\left(\operatorname{tr}\left(k_{t}^{+}(x, x)\right)-\operatorname{tr}\left(k_{t}^{-}(x, x)\right) d V(x)\right.
$$

Now note that $k_{t}(x, x): \mathcal{E}_{x} \rightarrow \mathcal{E}_{x}$ can be defined as the operator which is $k_{t}^{ \pm}(x, x)$ on $\mathcal{E}_{x}^{ \pm}$, in which case, its supertrace:

$$
\operatorname{str} k_{t}(x, x)=\operatorname{tr} k_{t}^{+}(x, x)-\operatorname{tr} k_{t}^{-}(x, x)
$$

by the definition of supertrace above. The proposition follows.

Now let us consider the case of a spin-manifold $M$. We have:
Lemma 16.2.2. Let $M$ be a spin manifold of dimension $2 m$, and $\mathcal{E}$ be a Dirac bundle on it. By the Corollary 15.2.11 $\mathcal{E}$ is isomorphic as a Dirac bundle to $\mathcal{S}(M) \otimes_{\mathbb{C}} \mathcal{V}$, where $\mathcal{S}(M) \rightarrow M$ is the spin bundle on $M$ with its spin connection $\nabla^{\mathcal{S}}$, and $\mathcal{V}$ is a twisting bundle with some unitary connection $\nabla^{\mathcal{V}}$, and $\nabla^{\mathcal{E}}$ is the tensor product connection of these two connections. Then:
(i): There is an isomorphism of complex vector bundles:

$$
\operatorname{hom}_{\mathbb{C}}(\mathcal{E}, \mathcal{E}) \simeq \Lambda_{\mathbb{C}}^{*} T^{*} M \otimes \operatorname{hom}_{\mathbb{C}}(\mathcal{V}, \mathcal{V})
$$

(ii): For an endomorphism $K:=\alpha \otimes F \in \operatorname{hom}_{\mathbb{C}}\left(\mathcal{E}_{x}, \mathcal{E}_{x}\right)$ where $\alpha \in \Lambda_{\mathbb{C}}^{*} T_{x}^{*} M$ and $F \in \operatorname{hom}_{\mathbb{C}}\left(\mathcal{V}_{x}, \mathcal{V}_{x}\right)$, the supertrace:

$$
\operatorname{str}_{\mathcal{E}} K=(-2 i)^{m} T(\alpha) \operatorname{tr}_{\mathcal{V}} F
$$

where $T: \Lambda_{\mathbb{C}}^{*}\left(T^{*} M\right) \rightarrow \Lambda_{\mathbb{C}}^{2 m}\left(T^{*} M\right)$ is the projection into the top-degree forms, introduced in Definition 14.5.1.

Proof: First note that the map:

$$
\begin{aligned}
\mathbb{C} l_{2 m} & \rightarrow \operatorname{hom}_{\mathbb{C}}\left(S_{2 m}, S_{2 m}\right) \\
c & \mapsto c .(-)
\end{aligned}
$$

is an isomorphism by the last assertion in Proposition 14.1.19. Since this isomorphism is canonical, we have a bundle isomorphism:

$$
\mathbb{C} l(M) \simeq \operatorname{hom}_{\mathbb{C}}(\mathcal{S}(M), \mathcal{S}(M))
$$

By the symbol map $\mathbb{C l}(M) \simeq \Lambda_{\mathbb{C}}^{*}\left(T^{*} M\right)$. Finally note that:

$$
\begin{aligned}
\operatorname{hom}_{\mathbb{C}}(\mathcal{E}, \mathcal{E}) & \simeq \operatorname{hom}_{\mathbb{C}}\left(\mathcal{S}(M) \otimes_{\mathbb{C}} \mathcal{V}, \mathcal{S}(M) \otimes_{\mathbb{C}} \mathcal{V}\right) \simeq \operatorname{hom}_{\mathbb{C}}(\mathcal{S}(M), \mathcal{S}(M)) \otimes_{\mathbb{C}} \operatorname{hom}_{\mathbb{C}}(\mathcal{V}, \mathcal{V}) \\
& \simeq \Lambda_{\mathbb{C}}^{*}\left(T^{*} M\right) \otimes_{\mathbb{C}} \operatorname{hom}_{\mathbb{C}}(\mathcal{V}, \mathcal{V})
\end{aligned}
$$

This proves (i).
To see (ii), note that we may view

$$
K \in \mathbb{C l}(M)_{x} \otimes_{\mathbb{C}} \operatorname{End}_{\mathbb{C}}\left(\mathcal{V}_{x}\right)
$$

as the element $c(\alpha) \otimes F$, where $c$ is the quantisation map identifying $\Lambda_{\mathbb{C}}^{*} T_{x}^{*} M$ with $\mathbb{C l}(M)_{x}$, and $c(\alpha)$ means the element of defined by Clifford multiplication by $c(\alpha) \in \operatorname{End}_{\mathbb{C}}\left(\mathcal{S}(M)_{x}\right)$. Then by definition:

$$
\operatorname{str}_{\mathcal{E}} K=\operatorname{str}_{\mathcal{E}}(c(\alpha) \otimes F)=\operatorname{tr}_{\mathcal{E}}\left(\tau_{2 m} \circ(c(\alpha) \otimes F)\right)=\operatorname{tr}_{\mathcal{E}}\left(\tau_{2 m} c(\alpha) \otimes F\right)=\operatorname{tr}_{\mathcal{S}}\left(\tau_{2 m} c(\alpha)\right) \cdot \operatorname{tr}_{\mathcal{V}} F=\operatorname{str}_{\mathcal{S}} c(\alpha) \cdot \operatorname{tr}_{\mathcal{V}} F
$$

where $\tau_{2 m}$ is the chirality element in $\mathbb{C l}(M)_{x}$. Now by the Lemma 14.5.2, we have

$$
\operatorname{str}_{\mathcal{S}} c(\alpha)=(-i)^{m}\left(\operatorname{dim}_{\mathbb{C}} \mathcal{S}\right)\left(T \circ \sigma(c(\alpha))=(-2 i)^{m}(T(\alpha))\right.
$$

Plugging this into the foregoing equation, we have (ii), and the lemma follows.

Proposition 16.2.3 (The index density). Let $M$ be a compact spin manifold of dimension $2 m$. Let $\mathcal{E} \rightarrow M$ be a Dirac bundle on $M$, and $D$ the corresponding Dirac operator. Note that $D^{2}$ is a generalised Laplacian by Corollary 15.4.8. For $a \in M$, there is an asymptotic expansion: 12.3.5:

$$
k_{t}(x, a) \sim(4 \pi t)^{-m} \exp \left(-\frac{\delta(x, a)^{2}}{4 t}\right)\left(\sum_{i=0}^{\infty} t^{i} k_{i}(x)\right)
$$

where $k_{0}(a)=I d_{\mathcal{E}}$. The the index of the Dirac operator is given by:

$$
\text { ind } D=\int_{M} \operatorname{str}_{\mathcal{E}} k_{t}(a, a) d V(a)=(4 \pi)^{-m} \int_{M} \operatorname{str}_{\mathcal{E}} k_{m}(a) d V(a)=\int_{M} \nu(a, \mathcal{E}) d V(a)
$$

The quantity $\nu(a, \mathcal{E})$ is called the index density of the Dirac operator of $\mathcal{E}$, and is is a polynomial in the jets of the metric and the unitary connection $\nabla^{\mathcal{V}}$ on the twisting bundle $\mathcal{V}$ at the point $a$.

Proof: As remarked earlier, we know from (iii) of the Proposition 10.1.3, that the smooth integral kernel $k_{t}(x, y)$ exists, and by the Proposition 16.2.1,

$$
\operatorname{ind} D=\int_{M} \operatorname{str}_{\mathcal{E}} k_{t}(a, a) d V(a)
$$

Choose a local framing $\left\{e_{j, y}\right\}$ for $\mathcal{E}_{\mid U}$ and $y \in U$ a neighbourhood of $a$, and note that for a basis element $e_{j, a} \in \mathcal{E}_{a}, u_{t}^{j}(y):=k_{t}(y, a) e_{j, a}$ is the fundamental solution of $e^{-t D^{2}}$ with pole at $\left(a, e_{j, a}\right)$, by Proposition 11.2.2. Thus

$$
k_{t}(y, a)=\sum_{j=1}^{\operatorname{dim} \mathcal{E}} u_{t}^{j}(y) \otimes e_{j, a}^{*}
$$

By the Theorem 12.3.5, there is an asymptotic expansion:

$$
u_{t}^{j}(y) \sim(4 \pi t)^{-m} \exp \left(-\frac{\delta(y, a)^{2}}{4 t}\right)\left(\sum_{i=0}^{\infty} t^{i} u_{i}^{j}(y)\right)
$$

such that $u_{0}^{j}(a)=e_{j} \in \mathcal{E}_{a}$, and for all $i$, the vector $u_{j}(a) \in \mathcal{E}_{a}$ is given as a polynomial in the jets of the coefficients of the Clifford connection $\Delta^{\mathcal{E}}$ at $a$. (See the last statement of Theorem 12.3.5, and note that the Dirac Laplacian is a generalised laplacian whose 0 -th and 1st order terms involve the connection coefficients of $\nabla^{\mathcal{E}}$, by 15.4.8). By considering $\sum_{j} u_{t}^{j}(y) \otimes e_{j, y}^{*}$, we have a corresponding asymptotic expansion for $k_{t}(y, a)$ given by:

$$
k_{t}(y, a)=(4 \pi t)^{-m} \exp \left(-\frac{\delta(y, a)^{2}}{4 t}\right)\left(\sum_{i=0}^{\infty} t^{i} k_{i}(y)\right)
$$

with $k_{0}(a)=\sum_{j} u_{0}^{j}(a) \otimes e_{j, a}^{*}=\sum_{j} e_{j, a} \otimes e_{j, a}^{*}=I_{\mathcal{E}_{a}}$, and $k_{i}(a)$ all depending polynomially on the jets of the connection coefficients of $\nabla^{\mathcal{E}}$ at $a$.

By (iii) of the Theorem 12.3.5

$$
\left\|u_{t}^{j}(y)-S_{k}(y)\right\|_{l, \infty} \leq C t^{N+1}
$$

for a sufficiently long partial sum $S_{k}$ of the asymptotic series for $u_{t}^{j}(y)$, and so a corresponding statement holds for $k_{t}(y, a)$. Since $\delta(a, a)=0$, it follows that the difference

$$
\left|\int_{M} \operatorname{str}_{\mathcal{E}} k_{t}(a, a) d V(a)-(4 \pi t)^{-m} \sum_{i=0}^{k} t^{i} \int_{M} \operatorname{str}_{\mathcal{E}} k_{i}(a)\right|<C t^{N+1}<\epsilon
$$

for $t<\delta$ and $k$ large enough depending on $m$ and $N$. The first integral inside the modulus sign is the index of $D$, and constant in $t$, so it follows that

$$
\operatorname{ind} D=\int_{M} \operatorname{str}_{\mathcal{E}} k_{t}(a, a) d V(a)=(4 \pi)^{-m} \int_{M} \operatorname{str}_{\mathcal{E}} k_{m}(a) d V(a)
$$

where $k_{m}(a) \in \operatorname{End}_{\mathbb{C}}\left(\mathcal{E}_{a}\right)$ depends polynomially on the jets of the connection coefficients of $\nabla^{\mathcal{E}}$ at $a$.
By (i) of the previous Lemma 16.2.2, we can write the endomorphism $k_{m}(a) \in \operatorname{End}_{\mathbb{C}}\left(\mathcal{E}_{a}\right)$ as:

$$
(4 \pi)^{-m} k_{m}(a)=\sum_{i=1}^{r} \alpha_{i}(a) \otimes F_{i}(a) \quad \alpha_{i} \in \Lambda_{\mathbb{C}}^{*}\left(T_{a}^{*} M\right), \quad F_{i}(a) \in \operatorname{End}_{\mathbb{C}}\left(\mathcal{V}_{a}\right)
$$

and by (ii) of the same Lemma,

$$
\operatorname{str}_{\mathcal{E}} k_{m}(a)=(-2 i)^{m} \sum_{i} T\left(\alpha_{i}(a)\right) \operatorname{tr}_{\mathcal{V}} F_{i}(a)=: \nu(a, \mathcal{E})
$$

Since the Clifford connection $\nabla^{\mathcal{E}}$ is the tensor product of the spin connection on $\mathcal{S}(M)$, and the unitary connection $\nabla^{\mathcal{V}}$, the polynomial dependence of $\nu(a, \mathcal{E})$ on the jets of the metric $g$ on $M$ and connection coefficients of $\nabla^{\mathcal{V}}$ at $a$ is clear from the corresponding fact about $k_{m}(a)$ stated above. This proves the proposition.
16.3. Local expression for $\nabla^{\mathcal{E}}$. In the light of the Proposition 16.2.3, all we need to do now is to find out $\nu(a, \mathcal{E})$ where

$$
\operatorname{str}_{\mathcal{E}} k_{m}(a)=(-2 i)^{m}(T \otimes \operatorname{tr})\left(k_{m}(a)\right)=\nu(a, \mathcal{E}) d V(a)
$$

This is a purely a point-wise problem at each $a \in M$. So we introduce the geodesic normal coordinates on a geodesically convex neighbourhood $U$ of $a$ (via the exponential map), with $\exp _{a}: T_{a} M \rightarrow M$ a diffeomorphism of some neighbourhood $W$ of $0 \in T_{a} M$ with $U$, and $\exp _{a}(0)=a$. This will give a local formula for the Dirac operator $D^{\mathcal{E}}$ on the neighbourhood $W$. If we can write the asymptotic expansion coefficient $k_{m}(0)$ for the heat kernel $k_{t}$ of $D^{\mathcal{E}}$ from this expression, then one can compute its supertrace.

We first need a lemma about synchronous frames (which we have been using in for $T^{*} M$ in the past).

Lemma 16.3.1 (Synchronous framings). Let $V \rightarrow M$ be a complex vector bundle with a connection $\nabla, M$ a Riemannian manifold. Then for $a \in M$, there exists a neighbourhood $U$ of $a$, a trivialisation of $V_{\mid U}$ by sections $\left\{s_{\alpha}\right\}$, and a coordinate system $\left\{x_{i}\right\}$ on $U$ with $a=(0, . ., 0)$ such that:
(i): The Cartan connection coefficients of $\nabla$ are given on $U$ by:

$$
\nabla s_{\alpha}=\omega \cdot s_{\alpha}=\sum_{\beta} \omega_{\beta \alpha} s_{\beta}
$$

where $\omega$ is a 1 -form with values in $\operatorname{End}_{\mathbb{C}}(V)$.
(ii): Denoting the curvature 2-form of $\nabla$ by $F$, and denoting the curvature coefficients by:

$$
F=\sum_{i<j} F_{i j} d x_{i} \wedge d x_{j}
$$

with $F_{i j}(x):=F\left(\partial_{i}, \partial_{j}\right)(x) \in \operatorname{End}\left(V_{x}\right)$, we have:

$$
\omega(x)=-\frac{1}{2} \sum_{i, j} F_{i j}(0) x_{j} d x_{i}+O\left(|x|^{2}\right)
$$

where $O\left(|x|^{2}\right)$ is a 1 -form with values in $\operatorname{End}_{\mathbb{C}}\left(V_{x}\right)$ (i.e. a section in $C^{\infty}\left(U, \Lambda^{1} T^{*} M \otimes E\right)$ ).
(iii): If $\nabla$ is a unitary connection with respect to a hermitian metric $(-,-)$ on $V$, we can choose $\left\{s_{\alpha}\right\}$ to be a frame that is orthonormal at each point of $U$.

Proof: We define the neighbourhood $U$ to be a geodesically convex neighbourhood of $a$, and the coordinate system via the exponential map. That is $\left(x_{1}, . ., x_{n}\right)$ are the coordinates of $x=\exp _{a}\left(x_{1}, . ., x_{n}\right)$, so $a=(0, . ., 0)$. Choose a frame $\left\{s_{\alpha}(0)\right\}$ of $V_{a}$, and for $x=\exp _{a}(x)$, define the framing $\left\{s_{\alpha}(x)\right\}$ of $V_{x}$ by parallel transport of $s_{\alpha}(0)$ to $x$ along the radial geodesic $\exp _{a}(t x)$. In case the connection is unitary, choose $\left\{s_{\alpha}(0)\right\}$ to be an orthonormal frame of $V_{a}$. In this event, since parallel transport preserves inner products, $\left\{s_{\alpha}(x)\right\}$ will be an orthonormal frame at $x$ for all $x \in U$. Hence (i) and (iii) are automatic by this definition. We need to verify (ii)

Clearly for every $v \in T_{a}(M)$, denoting parallel translation along $\exp _{a}(t v)$ by $P_{t}^{v}$, we have:

$$
\left(\nabla_{v} s_{\alpha}\right)(0)=\lim _{t \rightarrow 0} \frac{P_{-t}^{v} s_{\alpha}\left(\exp _{a}(t v)\right)-s_{\alpha}(0)}{t}=\lim _{t \rightarrow 0} \frac{s_{\alpha}(0)-s_{\alpha}(0)}{t}=0
$$

it follows that $\omega_{\beta \alpha}(0)=0$ for all $\alpha, \beta$, that is:

$$
\begin{equation*}
\omega(0)=0 \tag{52}
\end{equation*}
$$

Define the radial vector field $u:=\sum_{j} x_{j} \partial_{j}$, and by $i_{u}$ the operator $\left.u\right\lrcorner(-)$. Since $u(x)$ is tangent to the radial geodesic at $\exp _{a}(t x)$ through $x$, it follows by the definition of $s_{\alpha}$ that $\nabla_{u} s_{\alpha} \equiv 0$ on $U$. Thus, for each connection coefficient $\omega_{\beta \alpha}$, we have $i_{u} \omega_{\beta \alpha}=0$. That is, $i_{u} \omega \equiv 0$ on $U$. Writing $\omega=\sum_{i} \omega_{i} d x_{i}$, we have:

$$
0=i_{u} \omega(x)=\sum_{i} x_{i} \omega_{i}(x) \text { for all } x \in U
$$

Let us take the derivative of this last equation with respect to $x_{j}$. Then:

$$
\sum_{i}\left(\delta_{i j} \omega_{i}(x)+x_{i} \partial_{j} \omega_{i}(x)\right)=\omega_{j}(x)+\sum_{i} x_{i} \partial_{j} \omega_{i}(x)=0
$$

From which it follows by again applying $\partial_{i}$ that:

$$
\partial_{i} \omega_{j}(x)+\partial_{j} \omega_{i}(x)+\sum_{k} x_{k} \partial_{i} \partial_{j} \omega_{k}(x)=0 \text { for all } x \in U
$$

This shows that

$$
\partial_{i} \omega_{j}(0)=-\partial_{j} \omega_{i}(0)
$$

That is, the matrix $\partial_{i} \omega_{j}(0)$ is skew-symmetric.
Since $F=d \omega+\omega \wedge \omega$, and $\omega(0)=0$, it follows that

$$
F(0)=\sum_{i<j} F_{i j}(0) d x_{i} \wedge d x_{j}=\sum_{i<j} d \omega(0)=\sum_{i<j}\left(\partial_{i} \omega_{j}-\partial_{j} \omega_{i}\right)(0)\left(d x_{i} \wedge d x_{j}\right)=-2 \sum_{i<j} \partial_{j} \omega_{i}(0) d x_{i} \wedge d x_{j}
$$

That is,

$$
\partial_{j} \omega_{i}(0)=-\frac{1}{2} F_{i j}(0)
$$

Now we Taylor expand $\omega$ about 0 , noting that by (52) that $\omega_{i}(0)=0$ for all $i=1, . ., n$. Hence:

$$
\omega(x)=\sum_{i} \omega_{i}(x) d x_{i}=\sum_{i, j} x_{j} \partial_{j} \omega_{i}(0) d x_{i}+O\left(|x|^{2}\right)=-\frac{1}{2} \sum_{i, j} F_{i j}(0) x_{j} d x_{i}+O\left(|x|^{2}\right)
$$

where $O\left(|x|^{2}\right)$ is a 1-form with values in $\operatorname{End}\left(V_{x}\right)$. This proves the lemma.

Lemma 16.3.2 (Local expression for $\left.\nabla^{\mathcal{E}}\right)$. Let $M$ be a spin manifold of dimension $2 m$, and $\mathcal{E} \rightarrow M$ be Dirac bundle on $M$, with $\mathcal{E}=\mathcal{S}(M) \otimes_{\mathbb{C}} \mathcal{V}$, where $\mathcal{S}$ is the spin bundle on $M$, with its spin connection $\nabla^{\mathcal{S}}, \mathcal{V}$ is the twisting bundle with the twisting connection $\nabla^{\mathcal{V}}$, and $\nabla^{\mathcal{E}}$ the tensor product connection. and let ( $x_{1}, . ., x_{2 m}$ ) denote the geodesic coordinate system given by $\exp _{a}$ in a geodesically convex neighbourhood $U$ of $a=(0, . ., 0)$. Also, for $x \in U$, let the bundle $\mathcal{S}$ be trivialised over $U$ by parallel transport of an orthonormal frame $s_{\alpha}$ of $\mathcal{S}_{a}$ (with respect to $\nabla^{\mathcal{S}}$ ). Similarly, let $\mathcal{V}$ be trivialised over $U$ by parallel transport of an orthonormal frame $\left\{v_{\beta}\right\}$ of $\mathcal{V}_{a}$ (with respect to $\nabla^{\mathcal{V}}$ ). We will let $\left\{e_{i}(x)\right\}$ be local orthonormal frame for $T_{x}^{*} M$ obtained by parallel transport of a fixed orthonormal frame $e_{i}(a)=\partial_{i, a}$ along radial geodesics, with respect to the Levi-Civita connection on $T^{*} M$. Let $c_{i}=e_{i}(a) .(-)$ be Clifford multiplication on $\mathcal{E}_{a}$ by $e_{i}(a)$.

Then the covariant derivative $\nabla^{\mathcal{E}}$ is given by the formula:

$$
\nabla_{i}^{\mathcal{E}}=\nabla_{\partial_{i}}^{\mathcal{E}}=\frac{\partial}{\partial x_{i}}+\frac{1}{4} \sum_{j ; k<l} x_{j} R_{i j k l}(0) c_{k} c_{l}+\sum_{k<l} f_{i k l}(x) c_{k} c_{l}+g_{i}(x)
$$

where:

$$
\begin{aligned}
R_{i j k l} & =\left\langle R\left(\partial_{i}, \partial_{j}\right) e_{k}, e_{l}\right\rangle=\text { Riemann curvature tensor of } M \\
f_{i k l}(x) & \in C^{\infty}(U), \text { with } f_{i k l}(x)=O\left(|x|^{2}\right) \\
g_{i}(x) & \in C^{\infty}\left(U, \operatorname{End}_{\mathbb{C}}(\mathcal{V})\right)=C^{\infty}\left(U, \operatorname{End}_{\mathbb{C} l(M)}(\mathcal{E})\right) \text { with } g_{i}=O(|x|)
\end{aligned}
$$

(Here $|x|^{2}:=\sum_{i=1}^{2 m} x_{i}^{2}$ is the Euclidean norm of $x$.)

Proof: Using the geodesic (exponential) coordinate system above, we have $a=(0, . ., 0)$, so we will write 0 for $a$.

Define orthonormal framings $\left\{s_{\alpha}\right\}$ of $\mathcal{S}$ and $\left\{v_{\beta}\right\}$ of $\mathcal{V}$ on a geodesically convex neighbourhood $U$ of $a$ as stated above (and in the Lemma 16.3.1). By the fact that $\nabla^{\mathcal{E}}$ is the tensor product connection of $\nabla^{\mathcal{S}}$ and $\nabla^{\mathcal{V}}$, it follows that the framing $\left\{s_{\alpha} \otimes v_{\beta}\right\}$ is a orthonormal framing of $\mathcal{E}$ on $U$, which is parallel along radial geodesics. Likewise, for $T^{*} M_{\mid U}$, by the orthonormal frame field $\left\{e_{i}(x)\right\}$, with the further provision that $e_{i}(0)=\partial_{i, 0}$ (The derivative of the exponential map $\exp _{a}: T_{a} M \rightarrow M$ is the identity map, so $\left\{\partial_{i}\right\}$ can be taken as the image of an orthonormal basis in $\left.T_{a}(M)\right)$.

We first claim that with the above trivialisations of $\mathcal{E}$ and $T^{*} M$ on $U$, the operation of Clifford multiplication $c\left(e_{i}(x)\right)$ by $e_{i}(x) \in T_{x}^{*} M$ on $\mathcal{E}_{x}$ is the same as $c_{i}=c\left(e_{i}(0)\right)$. That is, the Clifford multiplication $c\left(e_{i}(x)\right)$ is a constant endomorphism of $\operatorname{End}\left(V_{x}\right) \simeq \operatorname{End}\left(V_{0}\right)$. This follows from the fact that $\nabla^{\mathcal{E}}$ is a Clifford connection, and is seen as follows.

That is, let $\partial_{r, x}:=\exp _{a *}\left(\partial_{r}\right)$ denote the radial vector field on $U$, and $e_{i, r}$ denote $e_{i}\left(\exp _{a}(r x)\right)$. Let $s$ be a vector in $\mathcal{E}_{0}$, and let $s_{r}:=s\left(\exp _{a}(r x)\right)=P_{r} s$, where $P_{r}$ denotes parallel transport from 0 to $\exp _{a}(r x)$ along the radial ray $r \mapsto r x$, with respect to $\nabla^{\mathcal{E}}$. Since $\nabla^{\mathcal{E}}$ is a Clifford connection, we have:

$$
\nabla_{\partial_{r}}^{\mathcal{E}}\left(c\left(e_{i, r}\right) s_{r}\right)=c\left(\nabla_{\partial_{r}}\left(e_{i, r}\right)\right) s_{r}+c\left(e_{i, r}\right) \nabla_{\partial_{r}}^{\mathcal{E}}\left(s_{r}\right)=0
$$

because $e_{i, r}$ is parallel along $\exp _{a}(r x)$ with respect to the Levi-Civita connection $\nabla$, and $s_{r}$ are parallel along $\exp _{a}(r x)$ with respect to the Clifford connection $\nabla^{\mathcal{E}}$ by definition of $s_{r}$. Now if we write:

$$
c\left(e_{i, r}\right) s_{\alpha, r}=\sum_{\beta} A_{\beta, \alpha}(r) s_{\beta, r}
$$

with respect to the parallel frame $\left\{s_{\alpha, r}\right\}$ of $\mathcal{E}_{\exp (r x)}$, then

$$
0=\nabla_{\partial_{r}}\left(c\left(e_{i, r}\right) s_{\alpha, r}\right)=\sum_{\beta}\left(\partial_{r} A_{\beta, \alpha}(r)\right) s_{\beta, r}
$$

by Leibnitz rule, and since $s_{\beta, r}$ is parallel. Hence $\partial_{r} A_{\beta, \alpha}(r)=0$, and hence $A_{\beta \alpha}(r)=A_{\beta \alpha}(0)$. Thus Clifford multiplication $c\left(e_{i, r}\right)$ is a constant operator along the geodesic rays, with the trivialisation above. Hence the claim.

Let us denote the Cartan connection 1-form on $U$ for $\nabla^{\mathcal{S}}$ by $\omega^{\mathcal{S}}$. Similarly denote the Cartan connection 1 -form for $\nabla^{\mathcal{V}}$ as $\omega^{\mathcal{V}}$. By definition,

$$
\begin{equation*}
\omega^{\mathcal{E}}=\omega^{\mathcal{S}}+\omega^{\mathcal{V}} \tag{53}
\end{equation*}
$$

Now, because of the trivialisations we have chosen, we may appeal to the Lemma 16.3.1, we may write:

$$
\begin{equation*}
\omega^{\mathcal{S}}(x)=-\frac{1}{2} \sum_{i, j} F_{i j}^{\mathcal{S}}(0) x_{j} d x_{i}+f(x) \tag{54}
\end{equation*}
$$

where $f(x) \in C^{\infty}\left(U, \Lambda^{1} T^{*} M \otimes \operatorname{End}(\mathcal{S})\right.$ is $O(|x|)^{2}$. We have already seen in the proof of Weitzenbock's formula in 15.4.3 for the spin bundle $\mathcal{S}$ that:

$$
F_{i j}^{\mathcal{S}}(x)=\Omega_{i j}^{\mathcal{S}}=-\frac{1}{2} \sum_{k, l} R_{i j k l}(x) c\left(e_{k}(x)\right) c\left(e_{l}(x)\right)
$$

so that:

$$
F_{i j}^{\mathcal{S}}(0)=-\frac{1}{2} \sum_{k<l} R_{i j k l}(0) c_{k} c_{l}
$$

where $R$ is the Riemann curvature tensor of $M$. Finally, since by the constancy of Clifford multiplication on $U$ proved above, we have

$$
\Lambda^{1}(U) \otimes \operatorname{End}\left(S_{0}\right) \simeq C^{\infty}\left(U, \Lambda^{1} T^{*} M \otimes \operatorname{End}\left(S_{x}\right)\right)
$$

via the isomorphism $\omega(x) \otimes c_{k} c_{l} \mapsto \omega(x) \otimes c\left(e_{k}(x)\right) c\left(e_{l}(x)\right)$. Thus we may write

$$
f(x)=\sum_{k<l} f_{k l}(x) c_{k} c_{l}=\sum_{i ; k<l} f_{i k l}(x) c_{k} c_{l} d x_{i}
$$

where $f_{i k l} \in C^{\infty}(U)$ and $f_{i k l}=O\left(|x|^{2}\right)$. Substituting in equation (54) we find that:

$$
\begin{equation*}
\omega^{\mathcal{S}}(x)=\sum_{i}\left[\frac{1}{4} \sum_{j ; k<l} R_{i j k l}(0) x_{j} c_{k} c_{l}+\sum_{k<l} f_{i k l}(x) c_{k} c_{l}\right] d x_{i} \quad \text { where } f_{i k l} \in C^{\infty}(U) \text { is } O\left(|x|^{2}\right) \tag{55}
\end{equation*}
$$

Now, applying the Lemma 16.3 .1 to the twisting connection $\nabla^{\mathcal{V}}$, we again find that with the framing and coordinate system we have used:

$$
\omega^{\mathcal{V}}(x)=-\frac{1}{2} \sum_{i, j} F_{i j}^{\mathcal{V}}(0) x_{j} d x_{i}+h
$$

where $h \in C^{\infty}\left(U, \Lambda^{1} T^{*} M \otimes \operatorname{End}(\mathcal{V})\right)$ is $O\left(|x|^{2}\right)$. Thus

$$
\begin{equation*}
\omega^{\mathcal{V}}(x)=\sum_{i} g_{i}(x) d x_{i} \tag{56}
\end{equation*}
$$

where $g_{i}(x):=-\frac{1}{2} \sum_{j} F_{i j}^{\mathcal{V}}(0) x_{j} d x_{i}+h_{i} \in C^{\infty}(U, \operatorname{End}(\mathcal{V}))$ and also $g_{i}=O(|x|)$.
Plugging the equations (55) and (56) into (53), we find that:

$$
\omega^{\mathcal{E}}=\sum_{i}\left[\frac{1}{4} \sum_{j ; k<l} R_{i j k l}(0) x_{j} c_{k} c_{l}+\sum_{k<l} f_{i k l}(x) c_{k} c_{l}+g_{i}(x)\right] d x_{i}
$$

where $f_{i k l}(x) \in C^{\infty}(U)$ is $O\left(|x|^{2}\right)$ and $g_{i}(x) \in C^{\infty}(U, \operatorname{End}(\mathcal{V}))$ is $O(|x|)$. Since $\nabla_{i}^{\mathcal{E}} s=\partial_{i} s+\omega^{\mathcal{E}}\left(\partial_{i}\right) s$, the lemma follows.
16.4. $u$-scaling. We will let $U$ be a neighbourhood of 0 in $\mathbb{R}^{2 m}$ which maps diffeomorphically onto a geodesically convex neighbourhood of a fixed point $a \in M$, where $M$ is a spin manifold of dimension $2 m$. This $U$ is to be thought of as the same $U$ encountered in all the Lemmas of the last subsection. Then, for $x \in U$, we will have $u^{1 / 2} x \in U$ for all $u \in(0,1]$.

Definition 16.4.1 (The scaling operators). Let $u \in(0,1]$. For a smooth section

$$
\alpha \in C^{\infty}\left((0, \infty) \times U, \Lambda_{\mathbb{C}}^{i}\left(\mathbb{R}^{2 m *}\right) \otimes \operatorname{End}(\mathcal{V})\right)
$$

where $\mathcal{V}$ is a fixed complex vector space, define the operator:

$$
\delta_{u}(\alpha)=u^{-i / 2} \alpha\left(u t, u^{1 / 2} x\right)
$$

For an operator

$$
T: C^{\infty}\left((0, \infty) \times U, \Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m *}\right) \otimes \operatorname{End}(\mathcal{V})\right) \rightarrow C^{\infty}\left((0, \infty) \times U, \Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m *}\right) \otimes \operatorname{End}(\mathcal{V})\right)
$$

define the operator $\delta_{u} T \delta_{u}^{-1}$, which may often be denoted as $T^{u}$, by

$$
\delta_{u} T \delta_{u}^{-1} \alpha=\delta_{u}\left(T\left(\delta_{u}^{-1} \alpha\right)\right)
$$

Lemma 16.4.2. We have the following identities:

$$
\begin{aligned}
\delta_{u} \phi(x) \delta_{u}^{-1} & =\phi\left(u^{1 / 2} x\right) \text { for } \phi \in C^{\infty}(U) \\
\delta_{u} \partial_{t} \delta_{u}^{-1} & =u^{-1} \partial_{t} \\
\delta_{u} \partial_{i} \delta_{u}^{-1} & =u^{-1 / 2} \partial_{i} \\
\delta_{u} e(\omega) \delta_{u}^{-1}:=\delta_{u}(\omega \wedge(-)) \delta_{u}^{-1} & =u^{-1 / 2} e(\omega) \text { for } \omega \in \mathbb{R}^{2 m *} \\
\left.\delta_{u} i(\omega) \delta_{u}^{-1}:=\delta_{u}(\omega\lrcorner(-)\right) \delta_{u}^{-1} & =u^{1 / 2} i(\omega) \text { for } \omega \in \mathbb{R}^{2 m *}
\end{aligned}
$$

Proof: If $\phi(x) \in C^{\infty}(U)$ is regarded as the operator of multiplication, then for $\alpha \in C^{\infty}\left((0, \infty) \times U, \Lambda^{i}\left(\mathbb{R}^{2 m *}\right) \otimes\right.$ $\operatorname{End}(\mathcal{V}))$ we have:

$$
\begin{aligned}
{\left[\delta_{u} \phi(x) \delta_{u}^{-1}(\alpha)\right](t, x) } & =\delta_{u}\left(\phi ( x ) \left(u^{i / 2} \alpha\left(u^{-1} t, u^{-1 / 2} x\right)\right.\right. \\
& =u^{-i / 2} \phi\left(u^{1 / 2}(x)\right) u^{i / 2} \alpha\left(u u^{-1} t, u^{1 / 2} u^{-1 / 2} t\right) \\
& =\phi\left(u^{1 / 2}(x)\right) \alpha(t, x)
\end{aligned}
$$

which proves the first identity. The next two are similar. For the fourth one, note:

$$
\left[\delta_{u} e(\omega) \delta_{u}^{-1}(\alpha)\right](t, x)=\delta_{u}\left(\omega \wedge u^{i / 2} \alpha\left(u^{-1} t, u^{-1 / 2} x\right)=u^{\frac{-i-1}{2}} u^{i / 2} \omega \wedge \alpha(t, x)=u^{-1 / 2}(e(\omega) \alpha)(t, x)\right.
$$

Similarly the last identity, since $i(\omega)$ reduces degree in $\Lambda^{*}$.

Remark 16.4.3. We have defined the scaling $\delta_{u}(-) \delta_{u}^{-1}$ on 1-forms $\omega$ on $U$, viewed as endomorphisms $e(\omega)$ or $i(\omega)$ of $\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m}\right) \otimes \mathcal{V}$. Our aim is to deform the Dirac Laplacian over $U$ into the generalised harmonic oscillator by letting the scaling factor $u \rightarrow 0$.

We note here that in the local formula for $\nabla_{i}^{\mathcal{E}}$ derived in Lemma 16.3.2, the terms involving $c_{k}, c_{l}$ signify Clifford multiplication by $e_{k}$ and $e_{l}$, regarded as endomorphisms of $\mathcal{S}(M)_{a}$. That is, as elements of $\mathbb{C l}(M)_{a}=$ End $_{\mathbb{C}}\left(\mathcal{S}(M)_{a}\right)$. Thus $c_{k} c_{l}$ in not a nilpotent endomorphism, and the hope is that after scaling, it will become a nilpotent endomorphism, indeed the element $e_{k} \wedge e_{l}$ in $\mathcal{A}=\Lambda_{\mathbb{C}}^{*}\left(T^{*} M_{a}\right)$. That this is indeed the case is the content of the next lemma.

Lemma 16.4.4 (The $u$-scaling on $\mathbb{C l}_{2 m}$ ). Note that the constant section $e_{i}$ on $U$ corresponds to the endomorphism $e_{i} .(-)$ in $\operatorname{End}_{\mathbb{C}}\left(S_{2 m}\right)=\mathbb{C} l_{2 m}$. If one identifies $\mathbb{C} l_{2 m}$ with the full exterior algebra $\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m}\right)$, then $e_{i}$ maps to $c_{i}=e\left(e_{i}\right)-i\left(e_{i}\right)$. The scaled Clifford section $c_{i}$, by definition in Lemma 16.4.2, is:

$$
\delta_{u}\left(c_{i}\right) \delta_{u}^{-1}=\delta_{u}\left(e\left(e_{i}\right)-i\left(e_{i}\right)\right) \delta_{u}^{-1}=u^{-1 / 2} e\left(e_{i}\right)-u^{1 / 2} i\left(e_{i}\right)
$$

Now we may extend this definition all over $\mathbb{C} l_{2 m}$ by setting:

$$
\delta_{u}\left(c_{1} \cdot c_{2}\right) \delta_{u}^{-1}=\left(\delta_{u} c_{1} \delta_{u}^{-1}\right)\left(\delta_{u} c_{2} \delta_{u}^{-1}\right)
$$

since Clifford multiplication in $\mathbb{C} l_{2 m}$ corresponds to composition of maps in $\operatorname{End}_{\mathbb{C}}\left(S_{2 m}\right)$. Then we have:
(i): If $c$ is a homogeneous element in $\mathbb{C} l_{2 m}$, that is $c=\sum_{|I|=k} a_{I} c_{I}$, where $c_{I}$ denotes the Clifford product $c_{i_{1}} \cdot c_{i_{2}} \ldots c_{i_{k}}$ in $\mathbb{C}_{2 m}$, we have:

$$
\delta_{u} c \delta_{u}^{-1}=u^{-k / 2}[e(\sigma(c))+O(u)]=u^{-k / 2}[\sigma(c) \wedge(-)+O(u)]
$$

where $\sigma$ is the symbol map.
(ii): For any $c=\sum_{|I| \leq k} f_{I}(x) c_{I} \in \mathbb{C l}(U)$, where $f_{I} \in C^{\infty}(U)$, and with leading homogeneous term of degree $k$. Then

$$
\lim _{u \rightarrow 0} u^{k} \delta_{u} c \delta_{u}^{-1}=\sum_{|I|=k} f_{I}(0)\left(e_{I} \wedge(-)\right)
$$

Proof: We first prove (i). Let $\left\{e_{i}\right\}$ denote an orthonormal frame for $T^{*}\left(\mathbb{R}^{2 m}\right)$, and let $c(x)=\sum_{|I|=k} a_{I}(x) c_{I}$ be homogeneous of degree $k$. Then, by definition, Then:

$$
\begin{aligned}
\left(\delta_{u} c \delta_{u}^{-1}\right)(x) & =\sum_{|I|=k} a_{I}\left(u^{-1 / 2} e\left(e_{i_{1}}\right)-u^{1 / 2} i\left(e_{i_{1}}\right)\right) \cdot\left(u^{-1 / 2} e\left(e_{i_{2}}\right)-u^{1 / 2} i\left(e_{i_{2}}\right)\right) \ldots\left(u^{-1 / 2} e\left(e_{i_{k}}\right)-u^{1 / 2} i\left(e_{i_{k}}\right)\right) \\
& =u^{-k / 2}\left(\sum_{I} a_{I} e\left(e_{i_{1}} \wedge e_{i_{2}} \wedge \ldots e_{i_{k}}\right)+\left(\text { terms with } u^{j} \text { with } j \geq-k / 2+1\right)\right. \\
& =u^{-k / 2}(e(\sigma(c))+O(u))
\end{aligned}
$$

Now (ii) follows immediately from (i).

Proposition 16.4.5. Let $M$ be a spin manifold of dimension $2 m$, and $D: C^{\infty}(M, \mathcal{E}) \rightarrow C^{\infty}(M, \mathcal{E})$ be the Dirac operator on the Dirac bundle $\mathcal{E}=\mathcal{S}(M) \otimes \mathcal{V}$. For each $a \in M$, there exists a coordinate chart $U$ around $a$, and framings of $\mathcal{S}, \mathcal{V}$ and $T^{*} M$ such that:
(i): The rescaled covariant derivative $\nabla^{\mathcal{E}, u}:=\delta_{u} \nabla_{i}^{\mathcal{E}} \delta_{u}^{-1}$ is given by:

$$
\nabla_{i}^{\mathcal{E}, u}=u^{-1 / 2}\left(\partial_{i}+\frac{1}{4} \sum_{j} R_{i j} x_{j}+\rho(u)\right)
$$

where $R_{i j}=\sum_{k<l} R_{i j k l}(a) e_{k} \wedge e_{l}$ is the curvature 2-form (as an endomorphism of $\Lambda_{\mathbb{C}}^{*}\left(T_{a}^{*} M\right)$ ), $R_{i j k l}$ being the Riemann curvature tensor of $M$ at $a$, and $\rho(u) \in \operatorname{End}\left(\mathcal{V}_{a}\right)$ is $O\left(u^{1 / 2}\right)$.
(ii): The rescaled Dirac Laplacian is given by:

$$
u\left(D^{u}\right)^{2}:=u\left(\delta_{u} D \delta_{u}^{-1}\right)^{2}=-\sum_{i=1}^{2 m}\left(\partial_{i}+\frac{1}{4} \sum_{j=1}^{2 m} R_{i j} x_{j}\right)^{2}+F+g(u)
$$

where $F=\Omega^{\mathcal{V}}(a)$ is the curvature 2-form of $\mathcal{V}$ at $a$, and $g(u) \in \operatorname{End}\left(\mathcal{V}_{a}\right)=O\left(u^{1 / 2}\right)$.
(iii): The limit as $u \rightarrow 0$ of $\left.u\left(D^{u}\right)^{2}\right)$ is given by:

$$
\lim _{u \rightarrow 0} u\left(D^{u}\right)^{2}=-\sum_{i=1}^{2 m}\left(\partial_{i}+\frac{1}{4} \sum_{j=1}^{2 m} R_{i j} x_{j}\right)^{2}+F
$$

the right hand side being exactly the generalised harmonic oscillator introduced in Proposition 16.1.7

Proof: We note that by the Lemma 16.3.2, and with the geodesically convex neighbourhood $U$ of $a$ and synchronous framings of $T^{*} M$ and $\mathcal{V}$ constructed there, we have $a=(0, . ., 0)$ and:

$$
\nabla_{i}^{\mathcal{E}}=\nabla_{\partial_{i}}^{\mathcal{E}}=\frac{\partial}{\partial x_{i}}+\frac{1}{4} \sum_{j ; k<l} x_{j} R_{i j k l}(0) c_{k} c_{l}+\sum_{k<l} f_{i k l}(x) c_{k} c_{l}+g_{i}(x)
$$

where:

$$
\begin{aligned}
R_{i j k l} & =\left\langle R\left(\partial_{i}, \partial_{j}\right) e_{k}, e_{l}\right\rangle=\text { Riemann curvature tensor of } M \\
f_{i k l}(x) & \in C^{\infty}(U), \text { with } f_{i k l}(x)=O\left(|x|^{2}\right) \\
g_{i}(x) & \in C^{\infty}\left(U, \operatorname{End}_{\mathbb{C}}(\mathcal{V})\right)=C^{\infty}\left(U, \operatorname{End}_{\mathbb{C l}(M)}(\mathcal{E})\right) \text { with } g_{i}=O(|x|)
\end{aligned}
$$

Now, by the previous Lemma 16.4.2, $\delta_{u} \partial_{i} \delta_{u}^{-1}=u^{-1 / 2} \partial_{i}$.
Next, since $\left.R_{i j}:=\sum_{k<l} R_{i j k l}(0) c_{k} c_{l} \in \operatorname{End}\left(\mathcal{E}_{a}\right)=\operatorname{End}\left(S_{2 m}\right) \otimes \operatorname{End}\left(\mathcal{V}_{a}\right) \simeq \mathbb{C l}(M)_{a} \otimes \operatorname{End}\left(\mathcal{V}_{a}\right)=\Lambda^{*} T_{a}^{*} M\right) \otimes$ $\operatorname{End}\left(\mathcal{V}_{a}\right)$, so we have

$$
\delta_{u} c_{k} c_{l} \delta_{u}^{-1}=u^{-1}\left(e_{k} \wedge e_{l}+O(u)\right)
$$

by (i) of the Lemma 16.4.4. On the other hand $\delta_{u} x_{j} \delta_{u}^{-1}=u^{1 / 2} x_{j}$ by the same Lemma. Thus :

$$
\delta_{u}\left(\sum_{j} R_{i j} x_{j}\right) \delta_{u}^{-1}=\delta_{u}\left(\sum_{j ; k<l} R_{i j k l}(0) x_{j} c_{k} c_{l}\right) \delta_{u}^{-1}=u^{-1 / 2}\left(\sum_{j} R_{i j} x_{j}+h(u)\right)
$$

where $h(u)=O(u)$ and $R_{i j}:=\sum_{k<l} R_{i j k l}(0) e_{k} \wedge e_{l}$.
Since $f_{i k l}(x) \in C^{\infty}(U)$ and $O\left(|x|^{2}\right)$, we have $\delta_{u} f_{i k l} \delta_{u}^{-1}=f_{i k l}\left(u^{1 / 2} x\right)=O(u)$. On the other hand we observed above that $\delta_{u}\left(c_{k} c_{l}\right) \delta_{u}^{-1}=u^{-1}\left(e_{k} \wedge e_{l}+O(u)\right)$. Thus

$$
\delta_{u} f_{i k l}(x) c_{k} c_{l} \delta_{u}^{-1}=O(1)
$$

Finally, since $g_{i}(x) \in \operatorname{End}(\mathcal{V})_{x}$ is $O(|x|)$, we have:

$$
\delta_{u} g_{i} \delta_{u}^{-1}=g_{i}\left(u^{1 / 2} x\right)=O\left(u^{1 / 2}\right)
$$

Thus we write

$$
\rho(u):=u^{1 / 2} \delta_{u}\left(\sum_{k<l} f_{i k l} c_{k} c_{l}+g_{i}\right) \delta_{u}^{-1}+h(u)
$$

and by the foregoing, we have $\rho(u)=O\left(u^{1 / 2}\right)$, and

$$
\left.\delta_{u} \nabla_{i}^{\mathcal{E}, u} \delta_{u}^{-1}=u^{-1 / 2}\left(\partial_{i}+\frac{1}{4} \sum_{j} R_{i j} x_{j}\right)+\rho(u)\right)
$$

with $\rho(u)=O\left(u^{1 / 2}\right)$. This proves (i).

To see (ii), we use the formula for the Dirac Laplacian derived in Corollary 15.4.6, viz.

$$
D^{2}=\nabla^{\mathcal{E} *} \nabla^{\mathcal{E}}+\frac{1}{2} \Omega^{\mathcal{E}}=\nabla^{\mathcal{E} *} \nabla^{\mathcal{E}}+\frac{1}{4} k+\Omega^{\mathcal{V}}
$$

where $\Omega^{\mathcal{V}}$ is the curvature operator of $\mathcal{V}$. Also, in the proof of the Corollary 15.4.8 and Lemma 12.2.4, we have seen that the first term is:

$$
\nabla^{\mathcal{E} *} \nabla^{\mathcal{E}}=\Delta^{\mathcal{E}}=-\sum_{i, j} g^{i j}\left(\nabla_{i}^{\mathcal{E}} \nabla_{j}^{\mathcal{E}}-\nabla_{\nabla_{\partial_{i}} \partial_{j}}^{\mathcal{E}}\right)=-\sum_{i, j} g^{i j}\left(\nabla_{i}^{\mathcal{E}} \nabla_{j}^{\mathcal{E}}-\sum_{k} \Gamma_{i j}^{k} \nabla_{k}^{\mathcal{E}}\right)
$$

where $\Gamma_{i j}^{k}$ are the Christoffel symbols of the Riemannian metric on $M$ with respect to the basis $\partial_{i}$. Since $g^{i j}(a)=\delta^{i j}$, we have $g^{i j}(x)=\delta_{i j}+h_{i j}(x)$, where $h_{i j}(x)=O(|x|)$.

Similarly, since $\nabla_{\partial_{i}} \partial_{j}(a)=\nabla_{e_{i}} e_{j}=0$ for all $i, j$, we have $\Gamma_{i j}^{k}(x)=O(|x|)$ by choice of synchronous framing of $T^{*} M$ on $U$. Thus we may write:

$$
\nabla^{\mathcal{E} *} \nabla^{\mathcal{E}}=-\sum_{i} \nabla_{i}^{\mathcal{E}} \nabla_{i}^{\mathcal{E}}+\left(\sum_{i, j} h_{i j}(x) \nabla_{i}^{\mathcal{E}} \nabla_{j}^{\mathcal{E}}+g_{i j} \Gamma_{i j}^{k} \nabla_{k}^{\mathcal{E}}\right)
$$

Thus, using (i) above, we have:

$$
\begin{aligned}
u \delta_{u}\left(\nabla^{\mathcal{E} *} \nabla^{\mathcal{E}}\right) \delta_{u}^{-1} & =-\sum_{i}\left(u^{1 / 2} \delta_{u} \nabla_{i}^{\mathcal{E}} \delta_{u}^{-1}\right)^{2}+\sum_{i, j} h_{i j}\left(u^{1 / 2} x\right)\left(u^{1 / 2} \delta_{u} \nabla_{i}^{\mathcal{E}} \delta_{u}^{-1}\right)\left(u^{1 / 2} \delta_{u} \nabla_{j}^{\mathcal{E}} \delta_{u}^{-1}\right) \\
& +u^{1 / 2} g_{i j}\left(u^{1 / 2} x\right) \Gamma_{i j}^{k}\left(u^{1 / 2} x\right)\left(u^{1 / 2} \delta_{u} \nabla_{k}^{\mathcal{E}} \delta_{u}^{-1}\right) \\
& =-\sum_{i}\left(\partial_{i}+\frac{1}{4} \sum_{j} R_{i j} x_{j}+\rho_{i}(u)\right)^{2} \\
& +\sum_{i, j} h_{i j}\left(u^{1 / 2} x\right)\left(\partial_{i}+\frac{1}{4} \sum_{l} R_{i l} x_{l}+\rho_{i}(u)\right)\left(\partial_{j}+\frac{1}{4} \sum_{l} R_{j l} x_{l}+\rho_{j}(u)\right) \\
& +g_{i j}\left(u^{1 / 2} x\right) \Gamma_{i j}^{k}\left(u^{1 / 2} x\right)\left(\partial_{k}+\frac{1}{4} \sum_{l} R_{k l} x_{l}+\rho_{k}(u)\right)
\end{aligned}
$$

Since $\rho_{i}(u)=O\left(u^{1 / 2}\right), h_{i j}\left(u^{1 / 2} x\right)=O\left(u^{1 / 2}\right)$ and $\Gamma_{i j}^{k}=O\left(u^{1 / 2}\right)$, we have finally:

$$
\begin{equation*}
u \delta_{u}\left(\nabla^{\mathcal{E} *} \nabla^{\mathcal{E}}\right) \delta_{u}^{-1}=-\sum_{i}\left(\partial_{i}+\frac{1}{4} \sum_{j} R_{i j} x_{j}\right)^{2}+\delta(u) \tag{57}
\end{equation*}
$$

where $\delta(u)=O\left(u^{1 / 2}\right)$.
Note that

$$
\begin{equation*}
u \delta_{u} k_{M} \delta_{u}^{-1}=u k\left(u^{1 / 2} x\right)=\epsilon(u)=O\left(u^{1 / 2}\right) \tag{58}
\end{equation*}
$$

Finally, reverting to the Corollary 15.4.6, we have that

$$
\Omega^{\mathcal{V}}(s \otimes \sigma)=R^{\mathcal{V}}(s \otimes \sigma)=\sum_{i<j} c_{i} c_{j} s \otimes \Omega_{i j}^{\mathcal{V}} \sigma
$$

where $\Omega_{i j}^{\mathcal{V}}(x)=\Omega^{\mathcal{V}}\left(e_{i}, e_{j}\right)(x)$ is the curvature endomorphism of $\mathcal{V}_{x}$. Thus again appealing to (i) of Lemma 16.4.4:

$$
\begin{align*}
u \delta_{u} \Omega^{\mathcal{V}} \delta_{u}^{-1} & =u \sum_{i<j}\left(\delta_{u} c_{i} c_{j} \delta_{u}^{-1} \Omega_{i j}^{\mathcal{V}}\left(u^{1 / 2} x\right)\right. \\
& =u \sum_{i<j}\left(u^{-1}\left(e_{i} \wedge e_{j}+O(u)\right)\left(\Omega_{i j}^{\mathcal{V}}(0)+O(u)\right)\right. \\
& =\sum_{i<j} \Omega^{\mathcal{V}}(0)\left(e_{i} \wedge e_{j}\right)+\gamma(u)=F+\gamma(u) \tag{59}
\end{align*}
$$

where $\gamma(u)=O\left(u^{1 / 2}\right)$, and $F=\sum_{i<j} \Omega_{i j}^{\mathcal{V}}(0) e_{i} \wedge e_{j}$ is the curvature endomorphism of $\mathcal{V}$ at $a=0$.
Adding together the equations (57), (58) and (59), we arrive at (ii), with $g(u):=\delta(u)+\epsilon(u)+\gamma(u)$ being $O\left(u^{1 / 2}\right)$.
(iii) is immediate from (ii). The proposition follows.

Now we need to construct the $u$-scaled heat kernel for the $u$-scaled heat operator $e^{-t u D^{u 2}}$. To this end, we have the following:

Proposition 16.4.6. Let $\mathcal{E} \rightarrow M$ be a Dirac bundle on the spin manifold $M$ of dimension $2 m$, and $D$ the associated Dirac operator. Let $k(t, x)$ be the fundamental solution for the heat operator of the Dirac laplacian:

$$
e^{-t D^{2}}: C^{\infty}(M, \mathcal{E}) \rightarrow C^{\infty}(M, \mathcal{E})
$$

with pole at $\left(a, I_{\mathcal{E}}\right)$ (as discussed in Proposition 16.2.3). Note that for each $t, k(t, x)$ is a smooth section of $\operatorname{End}(\mathcal{E})=\Lambda_{\mathbb{C}}^{*} T^{*} M \otimes \operatorname{End}_{\mathbb{C}} \mathcal{V}$. For $x \in U$, where $U$ is the neighbourhood of $a$ defined in Lemma 16.3.2 and its sequel, set

$$
r(u, t, x):=u^{m}\left(\delta_{u} k\right)(t, x)
$$

Then
(i):

$$
\left(\partial_{t}+u\left(D^{u}\right)^{2}\right) r(u, t, x)=0 \text { for } t \in(0, \infty) x \in U
$$

(ii): Denoting the identity map $I d: \mathcal{E}_{a} \rightarrow \mathcal{E}_{a}$ by $I_{a}$, we have:

$$
\lim _{t \rightarrow 0} r(u, t, x)=\delta_{a} I_{a}
$$

That is, $r(u, t, x)$ is the fundamental solution for the heat equation of the (scaled) elliptic operator $u D^{u 2}$ on the neighbourhood $U$, with pole at $\left(a, I_{a}\right)$.

Proof: We first check (i). Note that by Lemma 16.4.2 and that $\partial_{t} k=-D^{2} k$, we have:

$$
\begin{aligned}
\partial_{t}(r(u, t, x)) & =\partial_{t}\left(u^{m}\left(\delta_{u} k\right)(t, x)\right)=u^{m+1}\left[\left(u^{-1} \partial_{t}\right) \delta_{u} k\right](t, x) \\
& =u^{m+1}\left[\left(\delta_{u} \partial_{t} \delta_{u}^{-1}\right)\left(\delta_{u} k\right)\right](x, t) \\
& =u^{m+1}\left[\delta_{u} \partial_{t} k\right](t, x)=u^{m+1}\left[\delta_{u}\left(-D^{2} k\right)\right](t, x) \\
& =-u^{m+1}\left[\left(\delta_{u} D^{2} \delta_{u}^{-1}\right)\left(\delta_{u} k\right)\right](t, x)=-\left[u\left(D^{u}\right)^{2}\left(u^{m} \delta_{u} k\right)\right](t, x)=-u\left(D^{u}\right)^{2} r(u, t, x)
\end{aligned}
$$

This proves (i).
To see (ii), we note that by our asymptotic expansion for $k(t, x)$, we have, denoting $\delta(x, a)=|x|$ :

$$
k(t, x) \sim(4 \pi t)^{-m} \exp \left(-|x|^{2} / 4 t\right)\left[k_{0}(x)+\sum_{i \geq 1} k_{i}(x) t^{i}\right]
$$

where $k_{0}(a)=I_{a}$. Then note that:

$$
\begin{aligned}
r(u, t, x) & =u^{m}\left[\delta_{u} k\right](t, x)=u^{m}(4 \pi t u)^{-m} \exp \left(-\left|u^{1 / 2} x\right|^{2} / 4 u t\right)\left[k_{0}\left(u^{1 / 2} x\right)+\sum_{i \geq 1}(u t)^{i} \delta_{u} k_{i}\right] \\
& =(4 \pi t)^{-m} \exp \left(-|x|^{2} / 4 t\right)\left[k_{0}\left(u^{1 / 2} x\right)+\sum_{i \geq 1}(u t)^{i} \delta_{u} k_{i}\right]
\end{aligned}
$$

Now as $t \rightarrow 0$, the euclidean heat kernel $(4 \pi t)^{-m} \exp \left(-|x|^{2} / 4 t\right) \rightarrow \delta_{a}$. So the first term of the series on the right has the limit we want, viz.

$$
\lim _{t \rightarrow 0}(4 \pi t)^{-m} \exp \left(-|x|^{2} / 4 t\right) k_{0}\left(u^{1 / 2} x\right)=\delta_{a} k_{0}(a)=\delta_{0} I_{a}
$$

The other terms of course tend to 0 because they involve strictly positive powers of $t$. This proves (ii), and the proposition.

The relation of the fundamental solution $r(u, t, x)$ to $k(t, x)$ naturally implies a relation between their asymptotic expansions. More precisely:

Proposition 16.4.7. Let $U$ be a neighbourhood of $a=0$ as in the last proposition. Let us denote the fibres $\mathcal{E}_{0}=: E, \mathcal{V}_{0}=: V$, and the smooth function in $C^{\infty}((0, \infty) \times U)$

$$
q_{t}(x):=(4 \pi t)^{-m} \exp \left(-\delta(x, a)^{2} / 4 t\right)
$$

There exist $\Lambda^{*}\left(\mathbb{R}^{2 m *}\right) \otimes \operatorname{End}(V)$-valued polynomials $\gamma_{i}(t, x)$ on $(0, \infty) \times U$ such that:

$$
r(u, t, x) \sim q_{t}(x) \sum_{i=-2 m}^{\infty} u^{i / 2} \gamma_{i}(t, x)
$$

satisfying:

$$
\| \partial_{t}^{j} \partial_{x}^{\alpha}\left(r(u, t, x)-\sum_{i=-2 m}^{N} u^{i / 2} \gamma_{i}(t, x) \|_{\infty} \leq C(N, j, \alpha) u^{N}\right.
$$

where $\|-\|_{\infty}$ is the supremum norm on $U$. Furthermore, $\gamma_{i}(0,0)=0$ for $i \neq 0$, and $\gamma_{0}(0,0)=I_{E}$.

Proof: By (iii) of the Proposition 12.3.5, we have the asymptotic expansion:

$$
k(t, x) \sim q_{t}(x) \sum_{i=0}^{\infty} t^{i} k_{i}(x)
$$

where each $k_{i}(x) \in C^{\infty}\left(U, \operatorname{End}_{\mathbb{C}}(E)=C^{\infty}\left(U, \Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m *}\right) \otimes \operatorname{End}_{\mathbb{C}}(V)\right)\right.$, and $k_{0}(0)=I_{V}$. The symbol " $\sim$ " means that for the partial sum of the series on the right upto $i=l$ we have a sup-norm estimate:

$$
\left\|k(t, x)-q_{t}(x) \sum_{i=0}^{l} k_{i}(x) t^{i}\right\|_{\infty} \leq C_{l} t^{N} \quad \text { for all } l>N+2 m, t \in(0, T]
$$

where the sup-norm is over $U$.
First we want to replace the smooth $k_{i}(x)$ by polynomials $\psi_{i}(x)$ with coefficients in $\Lambda^{*}\left(\mathbb{R}^{2 m *}\right) \otimes \operatorname{End}_{\mathbb{C}}(V)$. Note that :

$$
r^{k} e^{-r^{2} / 4 t} \leq C_{k} t^{k / 2} \quad \text { for all } r \in[0, \infty)
$$

which implies that:

$$
\|x\|^{k} e^{-\|x\|^{2} / 4 t} \leq C_{k} t^{k / 2} \quad \text { for all } x \in \mathbb{R}^{2 m}
$$

Thus, if we define the polynomials $\psi_{i}(x)$ to be the Taylor polynomial of $k_{i}(x)$ of order $2(N+m-i)$, then by Taylor's theorem

$$
\left|k_{i}(x)-\psi_{i}(x)\right| \leq A_{i}\|x\|^{2 N+2 m-2 i} \quad \text { for } x \in U
$$

so that for some constants $B_{i}$ independent of $x$ and $t$ :

$$
\left|q_{t}(x) \sum_{i=0}^{l} t^{i}\left(k_{i}(x)-\psi_{i}(x)\right)\right| \leq \sum_{i=0}^{l} A_{i} q_{t}(x)\|x\|^{2 N+2 m-2 i}=(4 \pi)^{-m} \sum_{i=0}^{l} B_{i} t^{-m} t^{i} e^{-\|x\|^{2} / 4 t}\|x\|^{2 N+2 m-2 i} \leq C t^{N}
$$

Thus

$$
\left\|k(t, x)-q_{t}(x) \sum_{i=0}^{l} t^{i} \psi_{i}(x)\right\|_{\infty} \leq C_{l} t^{N} \quad \text { for all } l>N+2 m, t \in(0, T]
$$

and now $\psi_{i}(x)$ are polynomials in $x$ with coefficients in $\operatorname{End}_{\mathbb{C}}(E)=\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m *}\right) \otimes \operatorname{End}_{\mathbb{C}}(V)$.
For an element of $\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m *}\right) \otimes \operatorname{End}_{\mathbb{C}}(V)$, let us denote the by $\alpha_{[p]}$ the component of $\alpha$ in the summand $\Lambda^{p}\left(\mathbb{R}^{2 m *}\right) \otimes \operatorname{End}_{\mathbb{C}}(V)$.

Then the last sup-norm inequality above implies sup-norm inequalities for each $p$-component, viz.

$$
\left\|k(t, x)_{[p]}-q_{t}(x) \sum_{i=0}^{l} t^{i}\left(\psi_{i}(x)\right)_{[p]}\right\|_{\infty} \leq C_{l} t^{N} \quad \text { for all } l>N+2 m, t \in(0, T]
$$

which implies, on multiplying both sides by $u^{m-p / 2}$, and resetting $x \mapsto u^{1 / 2} x$ (which maps $U$ to itself) and $t \mapsto u t$ that:

$$
\left\|u^{m} u^{-p / 2} k\left(u t, u^{1 / 2} x\right)_{[p]}-u^{m} u^{-p / 2} q_{u t}\left(u^{1 / 2} x\right) \sum_{i=0}^{l}(u t)^{i}\left(\psi_{i}\left(u^{1 / 2} x\right)\right)_{[p]}\right\|_{\infty} \leq u^{m-p / 2} C_{l}(u t)^{N}
$$

$$
\text { for all } l>N+2 m, \quad t \in(0, T]
$$

Since $u^{m} u^{-p / 2} k\left(u t, u^{1 / 2} x\right)=u^{m} \delta_{u} k(t, x)_{[p]}=r(u, t, x)_{[p]}$, and $u^{m} q_{u t}\left(u^{1 / 2} x\right)=q_{t}(x)$, the last inequality can be rewritten as:

$$
\left\|r(u, t, x)_{[p]}-u^{-p / 2} q_{t}(x) \sum_{i=0}^{l}(u t)^{i}\left(\psi_{i}\left(u^{1 / 2} x\right)\right)_{[p]}\right\|_{\infty} \leq u^{m-p / 2+N} C_{l} t^{N} \quad \text { for all } l>N+2 m, t \in(0, T]
$$

We need to let $l \rightarrow \infty$, and arrange the sum in the norm signs on the left hand side in powers of $u$. Note that since $\psi_{i}$ are polynomials, they will contribute non-negative powers of $u^{1 / 2}$, so that $u^{j / 2}$ will be contributed only by terms from $i=0$ to $i=j / 2+p / 2$, and $j / 2$ will run from $-p / 2$ onwards.

So define:

$$
\left(\gamma_{j}\right)_{[p]}(x, t):=\text { coefficient of } u^{j / 2} \text { in } u^{-p / 2} \sum_{i=0}^{j / 2+p / 2}(u t)^{i} \psi_{i}\left(u^{1 / 2} x\right)_{[p]}
$$

and rewrite the last inequality above for the particular value $l=(j+p) / 2$ as:
$\left\|r(u, t, x)_{[p]}-q_{t}(x) \sum_{i=-p}^{j+p} u^{i / 2} \gamma_{i}(t, x)_{[p]}\right\|_{\infty} \leq u^{m-p / 2+N} C_{j} t^{N} \quad$ for all $(j+p) / 2>N+2 m, t \in(0, T], u \in(0,1]$
Set $\gamma_{i}(t, x):=\sum_{p=0}^{2 m} \gamma_{i}(t, x)_{[p]}$, and note that $p / 2 \leq m$ for all $p$, we have $(j+p) / 2>N+2 m \Leftrightarrow j / 2-m>$ $N+m-p / 2=N^{\prime}$ will be satisfied if we choose $j>2 N^{\prime}+2 m$. Replacing $N^{\prime}=N+m-p / 2$ by $N$, and noting that $t^{N}$ is bounded on $(0, T]$, we then have the inequality:

$$
\begin{equation*}
\left\|r(u, t, x)-q_{t}(x) \sum_{i=-m}^{2 N+2 m} u^{i / 2} \gamma_{i}(t, x)\right\|_{\infty} \leq C u^{N} \text { for all } N, t \in(0, T] \tag{60}
\end{equation*}
$$

A similar argument maybe given for the derivatives $\partial_{i}^{\alpha} \partial_{t}^{\beta} r(u, t, x)$, which is omitted.
Now for the final statement about $\gamma_{i}(0,0)$. Since by definition

$$
\begin{aligned}
\sum_{j=-m}^{\infty} u^{j / 2} \gamma_{j}(x, t) & =\sum_{j=-m}^{\infty} \sum_{p=0}^{2 m} u^{j / 2}\left(\gamma_{j}\right)_{[p]}(x, t) \\
& =\sum_{p=0}^{2 m} \sum_{i=0}^{\infty} u^{-p / 2}(u t)^{i} \psi_{i}\left(u^{1 / 2} x\right)_{[p]} \\
& =\sum_{p=0}^{2 m}\left(\delta_{u} \psi_{[p]}\right)(t, x)=\delta_{u} \psi(t, x)
\end{aligned}
$$

where we define:

$$
\psi(t, x):=\sum_{i=0}^{\infty} t^{i} \psi_{i}(x)
$$

Now note that by the above definition,

$$
\left(\delta_{u} \psi\right)(0,0)=\sum_{i=0}^{\infty}(u .0)^{i} \delta_{u} \psi_{i}(0)=\psi_{0}(0)=I_{0}
$$

Thus

$$
\sum_{j=-m}^{\infty} u^{j / 2} \gamma_{j}(0,0)=\delta_{u} \psi(0,0)=I_{0}
$$

which shows that $\gamma_{j}(0,0)=0$ for $j \neq 0$, and $\gamma_{0}(0,0)=I_{0}$. This proves the proposition.
Now we can combine the Propositions 16.1.7, 16.4.5, 16.4.6 16.4.7 to deduce the following corollary.

Corollary 16.4.8. In the $u$-expansion

$$
r(u, t, x) \sim q_{t}(x) \sum_{j=-m}^{\infty} u^{j / 2} \gamma_{j}(t, x)
$$

deduced in the last proposition, we have $\gamma_{j}(t, x) \equiv 0$ for $j<0$. That is, the Laurent expansion in $u^{1 / 2}$ of $r(u, t, x)$ about 0 has no poles. Secondly, $q_{t}(x) \gamma_{0}(t, x)$ is a formal fundamental solution to the heat equation for the generalised harmonic oscillator $H$ in Mehler's formula of 16.1 .7 with pole at $\left(0, I_{0}\right)$. That is,

$$
q_{t}(x) \gamma_{0}(t, x)=(4 \pi t)^{-m} j(t R)^{-1 / 2} \exp \left(\frac{-1}{4 t}\langle x|(t R / 2) \operatorname{coth}(t R / 2)|x\rangle\right) \exp (-t F)
$$

where $R$ is the nilpotent matrix $\sum_{i<j} R_{i j} e_{i} \wedge e_{j} \in \Lambda^{2}\left(\mathbb{R}^{2 m *}\right)$.

Proof: By the Proposition 16.4.6 we have:

$$
\left(\partial_{t}+u\left(D^{u}\right)^{2}\right) r(u, t, x)=0 \quad \text { for } t>0, \quad x \in U
$$

Let $\gamma_{-s}(t, x)$ be the first term in the series $r(u, t, x)=q_{t}(x) \sum_{j=-m} u^{j / 2} \gamma_{j}(t, x)$ which is not identically zero. Since all space and time derivatives of $r(u, t, x)$ are uniformly approximated on $U$ upto an arbitrarily large power of $u$ by the space and time derivatives of some partial sum of the asymptotic series above (by Proposition 16.4.7 above), and the $\gamma_{i}(t, x)$ are polynomials in $t$ and $x$, it follows that the asymptotic series is a formal power series solution to the scaled heat equation $\left(\partial_{t}+u\left(D^{u}\right)^{2}\right.$, that is:

$$
\left(\partial_{t}+u\left(D^{u}\right)^{2}\right)\left(q_{t}(x) \sum_{j=-s}^{\infty}\left(u^{j / 2} \gamma_{j}(t, x)\right)=0\right.
$$

Denoting by $H:=-\sum_{i=1}^{2 m}\left(\partial_{i}+1 / 4 \sum_{j} R_{i j} x_{j}\right)^{2}+F$ the generalised harmonic oscillator introduced earlier, and noting that by (ii) of Proposition 16.4.5 we have:

$$
u\left(D^{u}\right)^{2}=H+O\left(u^{1 / 2}\right)
$$

we have:

$$
\left(\partial_{t}+H+O\left(u^{1 / 2}\right)\right)\left(q_{t}(x) \sum_{j=-s}^{\infty}\left(u^{j / 2} \gamma_{j}(t, x)\right)=0\right.
$$

as an identity or formal power series in $u$. Since the lowest power of $u$ occurring on the right is from the first non-vanishing term of the formal series, we have

$$
\left(\partial_{t}+H\right)\left(u^{-s / 2} q_{t}(x) \gamma_{-s}(t, x)\right)=0
$$

as an identity in $(t, x)$. It follows that

$$
\left(\partial_{t}+H\right)\left(q_{t}(x) \gamma_{-s}(t, x)\right)=0
$$

is a solution to the heat equation. That is, $q_{t}(x) \gamma_{-s}(t, x)$ is a formal solution to the heat equation for the generalised harmonic oscillator. From the Proposition 16.1.7, it follows that this solution is determined by its initial value at $(t, x)=(0,0)$. But, from the Proposition 16.4.7, we have seen that $\gamma_{s}(0,0)=0$ for $s \neq 0$. It follows that $\gamma_{-s}(t, x) \equiv 0$ for all $s>0$. The first assertion follows.

For the second assertion, the above reasoning shows that we have

$$
\left(\partial_{t}+H\right)\left(q_{t}(x) \gamma_{0}(t, x)\right)=0
$$

with $\gamma_{0}(0,0)=I_{0}$ by the last statement of Proposition 16.4.7. Since the fundamental solution for this harmonic oscillator is unique, and

$$
p_{t}(x)=(4 \pi t)^{-m} j(t R)^{-1 / 2} \exp \left(\frac{-1}{4 t}\langle x|(t R / 2) \operatorname{coth}(t R / 2)|x\rangle\right) \exp (-t F)
$$

satisfies the same equation, with $p_{0}(0)=I_{0}$, we have the second assertion. The proposition follows.

### 16.5. The Index Theorem.

Theorem 16.5.1 (Atiyah-Singer). Let $M$ be a compact spin manifold of dimension $2 m$, and let $\mathcal{E}=\mathcal{S}(M) \otimes \mathcal{V}$ be a Dirac bundle on it, where $\mathcal{S}(M) \rightarrow M$ is the spin bundle on $M$, with its unitary spin connection $\left.\nabla^{\mathcal{S}}\right)$, $\mathcal{V}$ a twisting complex vector bundle with a unitary connection $\nabla^{\mathcal{V}}$ on it, and the Clifford connection $\nabla^{\mathcal{E}}$ the tensor product connection of $\nabla^{\mathcal{S}}$ and $\nabla^{\mathcal{V}}$. Then the index of the Dirac operator $D^{+}: C^{\infty}\left(M, \mathcal{E}^{+}\right) \rightarrow C^{\infty}\left(M, \mathcal{E}^{-}\right)$is given by the formula:

$$
\operatorname{ind}\left(D^{+}\right)=\int_{M} \widehat{A}(M) \wedge \operatorname{ch}(\mathcal{V})
$$

Proof: By the Proposition 16.2.3 we have:

$$
\operatorname{ind} D^{+}=\int_{M} \operatorname{str}_{\mathcal{E}} k_{t}(a, a) d V(a)=(4 \pi)^{-m} \int_{M} \operatorname{str}_{\mathcal{E}} k_{m}(a) d V(a)
$$

where we have expanded asymptotically:

$$
\begin{equation*}
k_{t}(x, a) \sim(4 \pi t)^{-m} \exp \left(\delta(x, a)^{2} / 4 t\right) \sum_{i=0}^{\infty} t^{i} k_{i}(x) \tag{61}
\end{equation*}
$$

By the Proposition 16.4.7 we have a neighbourhood $U$ of $a=0$ such that on $U$

$$
\begin{equation*}
r(u, t, x)=u^{m} \delta_{u} k_{t}(x, 0)=u^{m} \sum_{p=0}^{2 m} u^{-p / 2} k\left(u t, u^{1 / 2} x\right)_{[p]}=\sum_{p=0}^{2 m} u^{m-p / 2} k\left(u t, u^{1 / 2} x\right)_{[p]} \tag{62}
\end{equation*}
$$

Denote $\mathcal{E}_{a}=\mathcal{E}_{0}=E, \mathcal{S}_{a}=S$, and $\mathcal{V}_{a}=V$.
By the Lemma 14.5.2, 16.2.3 we have

$$
\operatorname{str}_{E}(\alpha \otimes F) d V(a)=(-2 i)^{m} T(\alpha) \operatorname{tr}_{V} d V(a)=(-2 i)^{m} \operatorname{tr}_{V}\left((\alpha \otimes F)_{[2 m]}\right)
$$

where $(\alpha \otimes F)$ is to be regarded as a differential form with coefficients in $\operatorname{End}(E)=\Lambda^{*}\left(\mathbb{R}^{2 m *}\right) \otimes \operatorname{End}(V)$, and $\operatorname{tr}_{V}$ is applied to these coefficients, and the $2 m$-component applies to $\alpha$. In particular, for any element $r$ in $\operatorname{End}(E)=\Lambda^{*}\left(\mathbb{R}^{2 m *}\right) \otimes \operatorname{End}(V)$, we have:

$$
\operatorname{str}_{E}(k)=\operatorname{str}_{E} k_{[2 m]}
$$

Applying this to the equation (62) above, and using the Corollary 16.4.8 we find that:

$$
\begin{equation*}
\operatorname{str}_{E} k(u t, 0)=\operatorname{str}_{E} k(u t, 0)_{[2 m]}=\operatorname{str}_{E} r(u, t, 0)=q_{t}(0) \operatorname{str}_{E}\left(\sum_{j=0}^{\infty} u^{j / 2} \gamma_{j}(t, 0)\right) \tag{63}
\end{equation*}
$$

On the other hand we have from (61) and Proposition 16.2 .3 that scaling time by $u$ does not affect the integral over $M$, since only the time independent term $q_{t}(a) k_{m}(a)$ contributes to the integral. That is,

$$
\begin{aligned}
\int_{M} \operatorname{str}_{E} k_{t}(a, a) d V(a) & \left.=\int_{M} \operatorname{str}_{E}\left(q_{t}(a) t^{m} k_{m}(a)\right) d V(a)\right)=\int_{M}\left(q_{u t}(a)(u t)^{m} k_{m}(a)\right) d V(a) \\
& =\int_{M} \operatorname{str}_{E} k_{u t}(a, a) d V(a) \text { for all } u \in(0,1], t \in(0, T]
\end{aligned}
$$

In particular, by substituting (63) into this relation, and noting that $k(t, 0)=k_{t}(a, a)$ by definitions, we have:

$$
\begin{aligned}
\int_{M} \operatorname{str}_{E} k_{t}(a, a) d V(a) & =\lim _{u \rightarrow 0} \int_{M} \operatorname{str}_{E} k_{u t}(a, a) d V(a)=\lim _{u \rightarrow 0} \int_{M} \operatorname{str}_{E} k(u t, 0) d V(a) \\
& =\lim _{u \rightarrow 0} \int_{M} \operatorname{str}_{E} r(u, t, 0) d V(a)=\int_{M} \operatorname{str}_{E} q_{t}(0) \gamma_{0}(t, 0) d V(a)
\end{aligned}
$$

Since the left hand side is independent of $t$, we can evaluate the right hand side at $t=1$. From the Corollary 16.4.8,

$$
q_{1}(0) \gamma(1,0)=(4 \pi)^{-m} j(R)^{-1 / 2} \exp (-F)
$$

where $R=\sum_{i j} R_{i j} e_{i} \wedge e_{j}$ is the nilpotent curvature form with $R_{i j}=\frac{1}{2} R_{i j k l}(a) c_{k} c_{l}$, and $F=\sum_{i<j} \Omega_{i<j}^{\mathcal{V}} e_{i} \wedge e_{j}$ is the curvature form, (being regarded as a ( $\operatorname{dim} V \times \operatorname{dim} V$ )-matrix whose entries are 2-forms, i.e. in the nilpotent algebra $\mathcal{A}=\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m *}\right)$ ). By the Lemma 14.5.2, we have therefore:

$$
\begin{aligned}
\operatorname{str}_{E}\left(q_{1}(0) \gamma(1,0)\right) d V & =(4 \pi)^{-m}\left(\operatorname{str}_{E}\left(j(R)^{-1 / 2} \exp (-F)\right)=(-2 i)^{m}(4 \pi)^{-m} \operatorname{tr}_{V}\left(j(R)^{-1 / 2} \exp (-F)\right)_{[2 m]}\right. \\
& =(2 \pi i)^{-m}\left(\operatorname{tr}_{V}\left(j(R)^{-1 / 2} \exp (-F)\right)_{[2 m]}\right.
\end{aligned}
$$

where $R=R_{0}=R_{a}$, and $F=F_{0}=F_{a}$.
Now $j\left(R_{a}\right)^{-1 / 2}=\left[\operatorname{det}\left(\frac{R_{a} / 2}{\sinh R_{a} / 2}\right)\right]^{1 / 2}$ is by definition the element $\widehat{A}(M)(a)=\sum_{p=0}^{2 m} \widehat{A}(M)_{[p]}(a) \in \Lambda_{\mathbb{C}}^{*}(M)$. Similarly, the Chern character of $\mathcal{V}$ is defined by

$$
\operatorname{ch}(\mathcal{V})(a)=\sum_{p=0}^{2 m} c h_{[p]}(\mathcal{V})=(2 \pi i)^{-m} \operatorname{tr}_{V_{a}}(\exp (-F(a)))
$$

so that

$$
(2 \pi i)^{-m}\left(\operatorname{tr}_{V}\left(j(R)^{-1 / 2} \exp (-F)\right)_{[2 m]}=\widehat{A}(M) \wedge \operatorname{ch}(\mathcal{V})_{[2 m]}\right.
$$

and so

$$
\operatorname{ind}\left(D^{+}\right)=\int_{M} \widehat{A}(M) \operatorname{ch}(\mathcal{V})_{[2 m]}=: \int_{M} \widehat{A}(M) \operatorname{ch}(\mathcal{V})
$$

and the theorem follows.

Corollary 16.5.2 (Atiyah-Singer). Let $M$ be a compact spin manifold of dimension $2 m$. Then, for the Dirac operator $D^{\mathcal{S}}$ of the spin bundle $\mathcal{S}$ (called the Atiyah-Singer operator), we have:

$$
\operatorname{ind}\left(D^{\mathcal{S}}\right)=\widehat{A} \text {-genus of } M:=\int_{M} \widehat{A}(M)
$$

Proof: Set $\mathcal{E}=\mathcal{S}$, and $\mathcal{V}=M \times \mathbb{C}$, the trivial bundle of rank 1 , whose chern character is 1 , and apply the Theorem 16.5.1 above.

Corollary 16.5.3 (Lichnerowicz). Let $M$ be a compact spin manifold of everywhere strictly positive scalar curvature. Then the $\widehat{A}$ - genus of $M$ is zero.

Proof: By the Corollary 15.4.5, there are no harmonic spinors on $M$, i.e. dim ker $D^{\mathcal{S}}=0$. In particular both $D^{+}$and $D^{-}$have vanishing kernels, so ind $D^{+}=0$. This implies $\widehat{A}(M)=0$ by the Corollary 16.5.2 above.

## 17. Some Consequences of the Index Theorem

Definition 17.0.4 (L-class). Let $R=\sum_{i<j} R_{i j} e_{i} \wedge e_{j}$ denote the curvature form of an oriented Riemannian manifold $M$ of dimension $4 m$. Define the L-class of $M$ to be

$$
L(M)=(-2 \pi)^{-m}\left(\operatorname{det}\left(\frac{R / 2}{\tanh R / 2}\right)\right)^{1 / 2}
$$

We note that the justification for taking the square root of the determinant is identical to the one we had for the $\widehat{A}(M)$ class, see Definition 16.1.6. Its top degree component, viz. $L(M)_{[4 m]}$ turns out to be a polynomial in the Pontragin forms of $M$, called the Hirzebruch L-polynomial.

Theorem 17.0.5 (Hirzebruch Signature). Let $M$ be a compact oriented manifold of dimension $4 m$. Then the cup product pairing:

$$
\cup: H^{2 m}(M, \mathbb{R}) \otimes H^{2 m}(M, \mathbb{R}) \rightarrow \mathbb{R}
$$

is symmetric, and its signature is a homotopy invariant of $M$ called the signature of $M$, and denoted $\sigma(M)$. There is the following integral formula:

$$
\sigma(M)=\int_{M} L(M)_{[4 m]}
$$

Proof: We first note that there is the chirality operator $\tau_{4 m}$ which acts on $\mathcal{E}:=\mathbb{C l}(M)=\Lambda_{\mathbb{C}}^{*}\left(T^{*} M\right)$, and decomposes it into the $\pm 1$-eigenbundles $\mathcal{E}^{ \pm}$. With its Levi-Civita connection, we know by 15.1.8 and 15.2.8 that this gives it the structure of a Dirac bundle, and indeed, in the proof of Bochner's theorem 15.4.7, we saw that the Dirac operator $D=d+\delta$. We just need to (a) show that the index of $D$ is the signature $\sigma(M)$, and (b) identify the integrand which is the supertrace $\operatorname{str}_{\mathcal{E}}\left(k_{t}(a, a)\right) d V(a)$.

In (v) of Lemma 14.1.7, we showed that

$$
\tau_{4 m} \phi=i^{p+k(8 m+k-1)}(* \phi) \text { for } \phi \in \Lambda^{k}
$$

where $p=\left[\frac{4 m+1}{2}\right]=2 m$. Thus

$$
\tau_{4 m}=\epsilon(k) * \quad \text { on } \Lambda^{k} \text { where } \epsilon(k):=(-1)^{m+k(k-1) / 2}
$$

In particular $\tau_{4 m}$ is a real operator on $\Lambda_{\mathbb{C}}^{*}\left(T^{*} M\right)$, and for the middle dimension $m+2 m(2 m-1) / 2=m+$ $m(2 m-1)=2 m^{2}$, so $\tau_{4 m}$ agrees with $*$ on $\Lambda^{2 m}$. Hence we have:

$$
\begin{aligned}
& \Lambda_{\mathbb{C}}^{+}\left(T^{*} M\right)=\oplus_{0 \leq k \leq 2 m}(1+\epsilon(k) *) \Lambda_{\mathbb{C}}^{k}\left(T^{*} M\right)=\oplus_{2 m \leq k \leq 4 m}(1+\epsilon(k) *) \Lambda_{\mathbb{C}}^{4 m-k}\left(T^{*} M\right) \\
& \Lambda_{\mathbb{C}}^{-}\left(T^{*} M\right)=\oplus_{0 \leq k \leq 2 m}(1-\epsilon(k) *) \Lambda_{\mathbb{C}}^{k}\left(T^{*} M\right)=\oplus_{2 m \leq k \leq 4 m}(1-\epsilon(k) *) \Lambda_{\mathbb{C}}^{4 m-k}\left(T^{*} M\right)
\end{aligned}
$$

We know that $D\left(\tau_{4 m} \omega\right)=-\tau_{4 m} D \omega$, since $\nabla_{X} \tau_{4 m}=i^{2 m} \nabla_{X} \omega_{4 m}=0$ and $\tau_{2 m}$ anticommutes with $e_{i}$ in the Clifford algebra. Hence

$$
D \circ *=( \pm 1) * \circ D
$$

From this it follows that $\omega \in \Lambda^{k}(M, \mathbb{C})$ is a form in the kernel of $D^{2}=d \delta+\delta d=\Delta$, iff $* \omega \in \Lambda^{4 m-k}(M, \mathbb{C})$ is in the kernel of $D^{2}=\Delta$ as well. Denoting the harmonic forms in $\Lambda^{k}(M, \mathbb{C})$ by $\mathcal{H}^{k}$, the above decompositions imply that for $\Delta^{+}=D^{-} D^{+}$and $\Delta^{-}=D^{+} D^{-}$we have:

$$
\begin{aligned}
\operatorname{ker}\left(\Delta^{+}\right) & =\oplus_{0 \leq k \leq 2 m}(1+\epsilon(k) *) \mathcal{H}^{k} \\
\operatorname{ker}\left(\Delta^{-}\right) & =\oplus_{0 \leq k \leq 2 m}(1-\epsilon(k) *) \mathcal{H}^{k}
\end{aligned}
$$

Now, for $0 \leq k<2 m$, since $*$ maps $\mathcal{H}^{k}$ isomorphically to the space $\mathcal{H}^{4 m-k}$ with $\mathcal{H}^{k} \cap \mathcal{H}^{4 m-k}=\{0\}$, we see that $(1+\epsilon(k) *) \mathcal{H}^{k}$ and $(1-\epsilon(k) *) \mathcal{H}^{k}$ are isomorphic for $0 \leq k<2 m$.

For $k=2 m$, we have $\epsilon(2 m)=1$, and $(1 \pm \epsilon(2 m) *) \mathcal{H}^{2 m}$ are precisely the $( \pm 1)$-eigenspaces of $*: \mathcal{H}^{2 m} \rightarrow \mathcal{H}^{2 m}$. By the Hodge theorem, these are precisely the $( \pm 1)$-eigenspaces of the star operator $*$ on $H^{2 m}(M, \mathbb{C})$. Call them $H_{ \pm}^{2 m}$. Since:

$$
\langle\alpha \cup \beta,[M]\rangle=\int_{M} \alpha \wedge \beta
$$

it follows that the cup product pairing is positive definite (resp. negative definite) on the space $H^{2 m}(M, \mathbb{R})^{+}$ which is the real form of $H^{2 m+}$ (since $*$ is a real operator) (resp. $H^{2 m}(M, \mathbb{R})^{-}$, the real form of $H^{2 m-}$ ). Thus

$$
\begin{aligned}
\operatorname{ind}\left(D^{+}\right) & =\operatorname{dim}_{\mathbb{C}}\left(\operatorname{ker} \Delta^{+}\right)-\operatorname{dim}_{\mathbb{C}}\left(\operatorname{ker} \Delta^{-}\right)=\operatorname{dim}_{\mathbb{C}} H^{2 m+}-\operatorname{dim}_{\mathbb{C}} H^{2 m-} \\
& =\operatorname{dim}_{\mathbb{R}} H^{2 m}(M, \mathbb{R})^{+}-\operatorname{dim}_{\mathbb{R}} H^{2 m}(M, \mathbb{R})^{-}=\sigma(M)
\end{aligned}
$$

Now it remains to identify the integrand. Since every manifold is locally spin, say on some coordinate chart $U$, and so we have the identification:

$$
\mathbb{C l}(M)_{\mid U}=\mathcal{S}(M)_{\mid U} \otimes \mathcal{S}(M)_{\mid U}
$$

by (i) in Example 15.1.10. We need to apply the Atiyah-Singer theorem to get the local integrand, with $\mathcal{V}=\mathcal{S}$.

We already know $\widehat{A}(M)=\left[\operatorname{det}\left(\frac{R / 2}{\sinh (R / 2)}\right)\right]^{1 / 2}$ where $R$ is the curvature operator. We need to compute

$$
\operatorname{tr}_{\mathcal{V}}(\exp (-F))=\operatorname{tr}_{\mathcal{S}}(\exp (-F))
$$

where $F$ is the curvature form of $\mathcal{S}$ with respect to the connection $\nabla^{\mathcal{V}}=\nabla^{\mathcal{S}}$, i.e. the spin connection on $\mathcal{S}$. We have already seen that $F=-R$ as elements of $\Lambda^{2} \otimes s o(2 m) \simeq \Lambda^{2} \otimes C_{2}$. So we need a formula for $\operatorname{tr}_{\mathcal{S}}(\exp (R))$. Note that $R$ is a skew-symmetric $2 m \times 2 m$-matrix of 2 -forms since the spin connection is unitary.

Since we are at a point $a \in M$, we replace $\mathcal{S}_{a}$ by $S_{2 m}$. First assume $R \in \operatorname{End}_{\mathbb{C}}\left(S_{2 m}\right)$ is a skew-symmetric matrix with real scalar entries, instead of 2-form entries. If we find a power series representation for $\operatorname{tr}_{S_{2 m}}(\exp (R))$ in this case, then we can use the same power series representation when $R$ has entries in $\Lambda_{\mathbb{C}}^{2}$, since $R$ would then be nilpotent. (The same principle we applied in the proof of the Proposition 16.1.7).

First note that as a $\mathbb{C} l_{2 m}$-module by left multiplication, $\mathbb{C} l_{2 m}$ breaks up into $2^{m}$ identical copies of $S_{2 m}$, by (i) of Proposition 14.4.1. Thus for an endomorphism $R \in \operatorname{End}_{\mathbb{C}}\left(S_{2 m}\right)=\mathbb{C} l_{2 m}$, we have:

$$
\operatorname{tr}_{S}(\exp (R))=2^{-m} \operatorname{tr}_{\mathbb{C} l_{2 m}}(\exp (R))
$$

Suppose $R \in \mathbb{C} l_{2 m}$ is of the special block-diagonal form:

$$
R=t_{1} e_{1} e_{2}+t_{2} e_{3} e_{4}+\ldots+t_{m} e_{2 m-1} e_{2 m}
$$

Then, since $e_{2 j-1} e_{2 j}$ commutes with $e_{2 k-1} e_{2 k}$ for $j \neq k$, we have:

$$
\exp (R)=\exp \left(t_{1} e_{1} e_{2}\right) \exp \left(t_{2} e_{3} e_{4}\right) \ldots \exp \left(t_{m} e_{2 m-1} e_{2 m}\right)
$$

We have already seen in the proof of (v) in Proposition 13.2.2 that

$$
\exp \left(t_{j} e_{2 j-1} e_{2 j}\right)=\cos t_{j} . I+\sin t_{j} e_{2 j-1} e_{2 j}
$$

Note that $e_{2 j-1} e_{2 j}$ acts as a skew symmetric matrix on the plane spanned by $e_{2 j-1}$ and $e_{2 j}$, and a skewsymmetric matrix on the span of 1 and $e_{2 j-1} e_{2 j}$, and off-diagonal on all the rest of $\mathbb{C} l_{2 m}$. Thus it contributes nothing to the trace of $R$. Similar reasoning applies to any product of distinct doublets $e_{2 j-1} e_{2 j}$. Thus one sees that:

$$
\operatorname{tr}_{\mathbb{C} l_{2 m}} \exp \left(t_{1} e_{1} e_{2}\right) \exp \left(t_{2} e_{3} e_{4}\right) \ldots \exp \left(t_{m} e_{2 m-1} e_{2 m}\right)=\operatorname{tr}_{\mathbb{C}_{2 m}} \cos t_{1} \cos t_{2} \ldots \cos t_{m} I=2^{2 m} \prod_{j=1}^{m} \cos t_{j}
$$

Now the endomorphism $\left.R=t_{1} e_{1} e_{2}+t_{2} e_{3} e_{4}+\ldots t_{m} e_{2 m-1} e_{2 m}\right)$ is in $C_{2}(V)=\operatorname{Lie} \operatorname{Spin}(2 m) \simeq s o(2 m)$, and is identified with the matrix with $2 \times 2$-blocks of the form:

$$
\left(\begin{array}{cc}
0 & -2 t_{j} \\
2 t_{j} & 0
\end{array}\right)
$$

whose eigenvalues are $\pm \sqrt{-1}\left(2 t_{j}\right)$. Thus cosh $R / 2$ has eigenvalues $\cosh \left( \pm \sqrt{-1} t_{j}\right)=\cos t_{j}$. Hence det $\cosh R / 2=$ $\prod_{j=1}^{m} \cos ^{2} t_{j}$. As a consequence, we find that for $R$ of the block diagonal form above:

$$
\left.\operatorname{tr}_{S_{2 m}}(\exp (R))=2^{-m} \operatorname{tr}_{\mathbb{C} l_{2 m}}(\exp (R))=2^{-m} 2^{2 m}(\operatorname{det} \cosh R / 2)\right)^{1 / 2}=2^{m}(\operatorname{det} \cosh R / 2)^{1 / 2}
$$

Now we can assert the same formula for any skew-symmetric $2 m \times 2 m$-matrix by choice of suitable orthonormal basis $e_{1}, \ldots, e_{2 m}$, since both quantities of the equation above are unaffected by such a change.

Thus
$\widehat{A}(M) \wedge \operatorname{ch}(\mathcal{V})=(2 \pi i)^{-2 m} 2^{m}\left[\operatorname{det}\left(\frac{R / 2}{\sinh R / 2}\right) \operatorname{det}(\cosh R / 2)\right]^{1 / 2}=(-2 \pi)^{-m}\left[\operatorname{det}\left(\frac{R / 2}{\tanh R / 2}\right)\right]^{1 / 2}=L(M)$
and we have the signature theorem

$$
\sigma(M)=\int_{M} L(M)
$$

from the Atiyah-Singer Theorem 16.5.1.

We recall the definition of the Pfaffian of a $2 m \times 2 m$ skew-symmetric real matrix $A$. We note that there is a polynomial of degree $m$ in the entries of $A$ called the Pfaffian of $A$, and which satisfies:

$$
(P f(A))^{2}=\operatorname{det} A
$$

The easiest way to write an explicit formula for the Pfaffian in the entries of $A$ is to note that by an orthogonal transfomation we can bring $A$ into a normal block diagonal form with $m 2 \times 2$-blocks each of the kind:

$$
\left(\begin{array}{cc}
0 & -a_{i} \\
a_{i} & 0
\end{array}\right)
$$

Then the Pfaffian is just $\prod_{i=1}^{m} a_{i}=(\operatorname{det} A)^{1 / 2}$. If we define the 2 -form associated with a skew-symmetric matrix $A$, viz. $\omega_{A}:=\sum_{i<j} A_{i j} e_{i} \wedge e_{j}$ then, at least for $A$ of the form above, we see that

$$
\operatorname{Pf}(A) \omega_{2 m}=\frac{1}{m!}\left(\omega_{A} \wedge \omega_{A} \wedge \ldots \wedge \omega_{A}\right)
$$

An easy computation shows that an orthonormal change of basis $e_{i} \mapsto P e_{i}:=f_{i}$ results in transforming $\sum_{i<j} A_{i j} e_{i} \wedge e_{j}$ into $\sum_{k<l}\left(P A P^{t}\right)_{k, l} f_{k} \wedge f_{l}$, and so the above formula holds good for all skew- symmetric $A$.

Expanding the right hand side, we find that

$$
P f(A)=\frac{1}{m!} \sum_{\sigma \in S_{2 m}} A_{\sigma(1) \sigma(2)} A_{\sigma(3) \sigma(4)} \ldots A_{\sigma(2 m-1) \sigma(2 m)}
$$

Definition 17.0.6 (Euler form). Let $M$ be an oriented Riemannian manifold of dimension $2 m$. Let $R=$ $\sum_{i<j} R_{i j} e_{i} \wedge e_{j}$ be its curvature 2-form, where each $R_{i j}$ is the skew-symmetric matrix $\frac{1}{2} \sum_{k<l} R_{i j k l} c_{k} c_{l}$. We can then regard $R$ as a $2 m \times 2 m$-skew-symmetric matrix whose $(k, l)$-entry is the 2 -form $R^{k l}=\frac{1}{2} \sum_{i<j} R_{i j k l} e_{i} \wedge e_{j}$. Then define the Euler form of $M$ by the formula:

$$
e(M)=\frac{1}{(2 \pi)^{m}} P f(R)=\frac{1}{(2 \pi)^{m} m!} \sum_{\sigma \in S_{2 m}} R^{\sigma(1) \sigma(2)} \wedge R^{\sigma(3) \sigma(4)} \ldots \wedge R^{\sigma(2 m-1) \sigma(2 m)}
$$

which is a $2 m$-form.

Theorem 17.0.7 (Gauss-Bonnet-Chern-Allendoerfer). For $M$ an oriented compact Riemannian manifold of dimension $2 m$, we have:

$$
\chi(M)(=\text { Euler characteristic of } M):=\int_{M} e(M)=\sum_{i=0}^{2 m}(-1)^{i} \operatorname{dim}_{\mathbb{C}} H^{i}(M, \mathbb{C})
$$

Proof: In this case the Dirac bundle is $\mathcal{E}=\mathbb{C} l(M)=\Lambda_{\mathbb{C}}^{*}\left(T^{*} M\right)$ ), and the grading is not the chirality grading, but the parity grading (which comes from conjugation by $\omega_{2 m} \in \operatorname{Spin}(2 m)$ when $M$ is a spin manifold). That is, $\mathcal{E}^{+}=\mathbb{C} l^{0}(M)=\Lambda_{\mathbb{C}}^{e v}\left(T^{*} M\right), \quad \mathcal{E}^{-}=\mathbb{C} l^{1}(M)=\Lambda_{\mathbb{C}}^{o} T^{*} M$ ) (see (ii) of Remark 15.1.11). The Dirac operator is of course $d+\delta$, as we saw in the proof of the Bochner theorem 15.4.7. Thus $D^{2}=\Delta$, the Laplace-Beltrami operator on $M$, and by the Hodge-deRham Theorem (Corollary 9.5.3), we have $\operatorname{ker}\left(D^{-} D^{+}\right)=\oplus_{i=0}^{m} H^{2 i}(M, \mathbb{C})$, and $\operatorname{ker}\left(D^{+} D^{-}\right)=\oplus_{i=0}^{m} H^{2 i+1}(M, \mathbb{C})$. Thus

$$
\text { ind } D^{+}=\operatorname{dim} \operatorname{ker}\left(D^{-} D^{+}\right)-\operatorname{dim} \operatorname{ker}\left(D^{+} D^{-}\right)=\sum_{i=0}^{2 m}(-1)^{i} \operatorname{dim}_{\mathbb{C}} H^{i}(M, \mathbb{C})
$$

Again, to identify the integrand, we may use the fact that a coordinate chart $U$ is spin, and decompose $\mathbb{C l}(U)=\mathcal{S}(U) \otimes \mathcal{S}(U)$. However, to compute the supertrace, we have to compute with respect to this parity grading.

In the decomposition $\mathcal{E}=\mathcal{S} \otimes \mathcal{V}$ of a Dirac bundle on a spin manifold, we have assumed $\mathcal{S}$ is given the chirality grading and $\mathcal{V}$ is ungraded. The integrand of the Atiyah-Singer index theorem (i.e. the index density) has been calculated for this situation by using the fact that if $\alpha \otimes F$ is an endomorphism of a Clifford module $E=S_{2 m} \otimes V$, with $\alpha \in \operatorname{End}_{\mathbb{C}}\left(S_{2 m}\right)=\Lambda^{*}\left(\mathbb{R}^{2 m *}\right)$ and $F \in \operatorname{End}_{\mathbb{C}}(V)$, then

$$
\begin{equation*}
\operatorname{str}_{E}(\alpha \otimes F)=\operatorname{tr}_{E}\left(\tau_{2 m} \circ(\alpha \otimes V)\right)=(-2 i)^{m}\left(\alpha_{[2 m]} \otimes \operatorname{tr}_{V} F\right) \tag{64}
\end{equation*}
$$

(See Lemma 14.5.2). In the present situation, $E=S_{2 m} \otimes S_{2 m}$, which we are regarding as a graded module with the grading operator $\omega_{2 m} \otimes \omega_{2 m}$ instead of the earlier grading operator $\tau_{2 m} \otimes 1$.

But then, if $\alpha \otimes F \in \operatorname{End}_{\mathbb{C}}\left(S_{2 m}\right) \otimes \operatorname{End}_{\mathbb{C}}\left(S_{2 m}\right)=\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m *}\right) \otimes \mathbb{C} l_{2 m}$, we have:

$$
\begin{aligned}
\operatorname{str}_{E}(\alpha \otimes F) & :=\operatorname{tr}_{E}\left[\left(\omega_{2 m} \otimes \omega_{2 m}\right) \circ(\alpha \otimes F)\right] \\
& =\operatorname{tr}_{E}\left[\left((i)^{-m} \tau_{2 m} \circ \alpha\right) \otimes\left(\omega_{2 m} \otimes F\right)\right]=(-i)^{m} \operatorname{tr}_{S_{2 m}}\left(\tau_{2 m} \alpha\right) \operatorname{tr}_{S_{2 m}}\left(\omega_{2 m} \circ F\right) \\
& =(-i)^{m}(-2 i)^{m}(\alpha)_{[2 m]} \operatorname{tr}_{S_{2 m}}\left(\omega_{2 m} \circ F\right) \\
& =(-2)^{m} \alpha_{[2 m]} \operatorname{tr}_{S_{2 m}}\left(\omega_{2 m} F\right)
\end{aligned}
$$

(Note incidentally that $\left.\operatorname{tr}_{S_{2 m}}\left(\omega_{2 m} \circ F\right)=(-i)^{m} \operatorname{tr}_{S_{2 m}}\left(\tau_{2 m} \circ F\right)=(-i)^{m} \operatorname{str}_{S_{2 m}} F\right)$
So for a general endomorphism $k \in \operatorname{End}_{\mathbb{C}}(E)=\Lambda^{*} \otimes \mathbb{C} l$, we must modify the formula (64) by the formula:

$$
\begin{equation*}
\operatorname{str}_{E} k=(-2)^{m}\left[\operatorname{tr}_{S_{2 m}}\left(\omega_{2 m} k\right)\right]_{[2 m]} \tag{65}
\end{equation*}
$$

where $k$ is to be regarded as an element of $\Lambda_{\mathbb{C}}^{*}\left(\mathbb{R}^{2 m}\right)$ with coefficients in $\mathbb{C} l_{2 m}=\operatorname{End}\left(S_{2 m}\right)$, and the trace is to be applied to the coefficients after composing with $\omega_{2 m}$.

So in the Atiyah-singer theorem, we will have to make the corresponding modification of the integrand to read:

$$
\operatorname{ind}(d+\delta)=(-2 \pi)^{-m} \int_{M}\left[\widehat{A}(M) \operatorname{tr}_{\mathcal{S}}\left(\omega_{2 m} \exp (R)\right)\right]_{[2 m]}
$$

Again, by the same reasoning as in the proof of Hirzebruch signature theorem 17.0.5,

$$
\operatorname{tr}_{\mathcal{S}}\left(\omega_{2 m} \exp (R)\right)=2^{-m} \operatorname{tr}_{\mathbb{C} l}\left(\omega_{2 m} \exp (R)\right)
$$

Taking $R$ of the special form $R=\sum_{j=1}^{m} t_{j} e_{2 j-1} e_{2 j}$ we had computed in the proof of 17.0.5 that

$$
\exp (R)=\prod_{j=1}^{m}\left(\left(\cos t_{j}\right) I+\left(\sin t_{j}\right) e_{2 j-1} e_{2 j}\right)
$$

which implies, since distinct doublets $e_{2 j-1} e_{2 j}$ and $e_{2 k-1} e_{2 k}$ commute, that:

$$
\omega_{2 m} \exp (R)=\prod_{j=1}^{m} e_{2 j-1} e_{2 j}\left(\left(\cos t_{j}\right) I+\left(\sin t_{j}\right) e_{2 j-1} e_{2 j}\right)=\prod_{j=1}^{m}\left(\left(-\sin t_{j}\right) I+\left(\cos t_{j}\right) e_{2 j-1} e_{2 j}\right)
$$

As in the proof of the Hirzebruch theorem again, only the scalar term contributes to the trace, and this trace is

$$
\operatorname{tr}_{\mathbb{C} l}\left(\omega_{2 m} \exp (R)\right)=(-1)^{m}\left(2^{2 m}\right) \prod_{j=1}^{m} \sin t_{j}
$$

Hence

$$
\operatorname{tr}_{\mathcal{S}}\left(\omega_{2 m} \exp (R)\right)=(-2)^{m} \prod_{j=1}^{m} \sin t_{j}
$$

On the other hand, $R$ corresponds to the block $(2 m \times 2 m)$-matrix whose $2 \times 2$ blocks are

$$
\left(\begin{array}{cc}
0 & -2 t_{j} \\
2 t_{j} & 0
\end{array}\right)
$$

so that

$$
\operatorname{det} \sinh (R / 2)=\prod_{j=1}^{m}\left(\sinh \left(i t_{j}\right)\right)\left(\sinh \left(-i t_{j}\right)=\prod_{j=1}^{2 m}\left(i \sin t_{j}\right)\left(-i \sin t_{j}\right)=\prod_{j=1}^{m} \sin ^{2} t_{j}\right.
$$

So we conclude that for $R$ of the special form above,

$$
(-2)^{m}\left(\operatorname{det}(\sinh (R / 2))^{1 / 2}=\operatorname{tr}_{\mathcal{S}}\left(\omega_{2 m} \exp (R)\right)\right.
$$

Now once concludes the above formula for all skew-symmetric $R$ as before, by change of orthonormal basis. Hence the index theorem now reads:

$$
\begin{aligned}
\operatorname{ind}(d+\delta) & =(-2 \pi)^{-m} \int_{M}\left[\widehat{A}(M) \operatorname{tr}_{\mathcal{S}}\left(\omega_{2 m} \exp (R)\right)\right]_{[2 m]} \\
& =(-2 \pi)^{-m} \int_{M}\left[\operatorname{det}\left(\frac{R / 2}{\sinh (R / 2)}\right)^{1 / 2} \cdot(-2)^{m}\left(\operatorname{det}(\sinh (R / 2))^{1 / 2}\right]_{[2 m]}\right. \\
& =(\pi)^{-m} \int_{M}(\operatorname{det}(R / 2))^{1 / 2}=(2 \pi)^{-m} \int_{M}(\operatorname{det}(R))^{1 / 2} \\
& =\int_{M}(2 \pi)^{-m} \operatorname{Pf}(R)=\int_{M} e(M)
\end{aligned}
$$

This proves the proposition.

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