# LECTURES ON SEMICLASSICAL ANALYSIS 

## VERSION 0.2

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## Preface

This book originates with a course MZ taught at UC Berkeley during the spring semester of 2003, notes for which LCE took in class. In this presentation we have tried hard to work out the full details for many proofs only sketched in the original lectures. We have reworked the order of presentation, added many additional topics, and included more heuristic commentary. We have as well introduced consistent notation, recounted in Appendix A. Relevant functional analysis and other background mathematics have been consolidated into Appendices B-D.

This version 0.2 represents our latest draft, the clarity of which we (still) hope greatly to improve in later editions, to be posted on our websites. We are quite aware that many errors remain in our exposition, and so we ask our readers to please send any comments or corrections to us at evans@math.berkeley.edu or zworski@math.berkeley.edu.

We should mention that two excellent treatments of mathematical semiclassical analysis have appeared recently. The book [D-S] by M. Dimassi and J. Sjöstrand starts with the WKB-method, develops the general semiclassical calculus, and then provides high tech spectral asymptotics. The presentation of Martinez $[\mathrm{M}]$ is based on a systematic development of FBI (Fourier-Bros-Iagolnitzer) transform techniques, with applications to microlocal exponential estimates and propagation estimates. These notes are intended as a more elementary and broader introduction. Except for the general symbol calculus, where we followed Chapter 7 of [D-S], there is little overlap with these other two texts, or with the early and influential book by Robert $[\mathrm{R}]$.

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## 1. Introduction

1.1 Basic themes
1.2 Classical and quantum mechanics
1.3 Overview

### 1.1 BASIC THEMES

Our primary goal is understanding the relationships between dynamical systems and the behavior of solutions to various linear PDE and pseudodifferential equations containing a small positive parameter $h$.

PDE with small parameters. The principal realm of motivation is quantum mechanics, in which case we understand $h$ as denoting Planck's constant. With this interpretation in mind, we break down our basic task into these two subquestions:
(i) How and to what extent do classical dynamics determine the behavior as $h \rightarrow 0$ of solutions to Schrödinger's equation

$$
i h \partial_{t} u=-h^{2} \Delta u+V u
$$

for the potential $V=V(x)$, and the related eigenvalue equation

$$
-h^{2} \Delta u+V u=E u ?
$$

The name "semiclassical" comes from this interpretation.
(ii) Conversely, given various mathematical objects associated with classical mechanics, for instance symplectic transformations, how can we profitably "quantize" them?

In fact the techniques of semiclassical analysis apply in many other settings and for many other sorts of PDE. For example we will later study the damped wave equation

$$
\begin{equation*}
\partial_{t}^{2} u+a \partial_{t} u-\Delta u=0 \tag{1.1}
\end{equation*}
$$

for large times. A rescaling in time will introduce the requisite small parameter $h$.

Basic techniques. We will construct in Chapters 2-4 and 9 a wide variety of mathematical tools to address these issues, among them:

- the apparatus of symplectic geometry (to record succintly the behavior of classical dynamical systems);
- the Fourier transform (to display dependence upon both the position variables $x$ and the momentum variables $\xi$ );
- stationary phase (to describe asymptotics as $h \rightarrow 0$ of various expressions involving rescaled Fourier transforms);
- pseudodifferential operators (to localize or, as is said in the trade, to microlocalize functional behavior in phase space).


### 1.2 CLASSICAL AND QUANTUM MECHANICS

In this section we introduce and foreshadow a bit about quantum and classical correspondences.

Observables. We can think of a given function $a: \mathbb{R}^{n} \times \mathbb{R}^{n} \rightarrow \mathbb{C}$, $a=a(x, \xi)$, as a classical observable on phase space, where as above $x$ denotes position and $\xi$ momentum. We will also call $a$ a symbol.

Let $h>0$ be given. We will associate with the observable $a$, a corresponding quantum observable $a^{\mathrm{w}}(x, h D)$, an operator defined by the formula

$$
a^{\mathrm{w}}(x, h D) u(x):=\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}\langle x-y, \xi\rangle} a\left(\frac{x+y}{2}, \xi\right) u(y) d \xi d y
$$

for appropriate smooth functions $u$. This is Weyl's quantization formula.

We will later learn that if we change variables in a symbol, we preserve the principal symbol up to lower order terms (that is, terms involving high powers of the small parameter $h$.)

Equations of evolution. We are concerned as well with the evolution in time of classical particles and quantum states.
Classical evolution. Our most important example will concern the symbol

$$
p(x, \xi):=|\xi|^{2}+V(x)
$$

corresponding to the phase space flow

$$
\left\{\begin{array}{l}
\dot{x}=2 \xi \\
\dot{\xi}=-\partial V
\end{array}\right.
$$

We generalize by introducing the arbitrary Hamiltonian $p: \mathbb{R}^{n} \times \mathbb{R}^{n} \rightarrow$ $\mathbb{R}, p=p(x, \xi)$, and the corresponding Hamiltonian dynamics

$$
\left\{\begin{array}{l}
\dot{x}=\partial_{\xi} p(x, \xi)  \tag{1.2}\\
\dot{\xi}=-\partial_{x} p(x, \xi) .
\end{array}\right.
$$

It is instructive to change our viewpoint somewhat, by first introducing some more notation. Let us write

$$
\varphi_{t}=\exp \left(t H_{p}\right)
$$

for the solution of (1.2), where

$$
H_{p} q:=\{p, q\}=\left\langle\partial_{\xi} p, \partial_{x} q\right\rangle-\left\langle\partial_{x} p, \partial_{\xi} q\right\rangle
$$

is the Poisson bracket. Select a symbol $a$ and set

$$
a_{t}(x, \xi):=a\left(\varphi_{t}(x, \xi)\right) .
$$

Then

$$
\begin{equation*}
\frac{d}{d t} a_{t}=\left\{p, a_{t}\right\} . \tag{1.3}
\end{equation*}
$$

This equation tells us how the symbol evolves in time.
Quantum evolution. We next quantize the foregoing by putting $P=p^{\mathrm{w}}(x, h D), A=a^{\mathrm{w}}(x, h D)$ and defining

$$
A_{t}:=e^{\frac{i t P}{n}} A e^{-\frac{i t P}{n}} .
$$

Then we have the evolution equation

$$
\frac{d}{d t} A_{t}=\frac{i}{h}\left[P, A_{t}\right],
$$

an obvious analog of (1.3). Here is a basic principle we will later work out in some detail: an assertion about Hamiltonian dynamics, and so the Poisson bracket $\{\cdot, \cdot\}$, will involve at the quantum level the commutator $[\cdot, \cdot]$.

### 1.3 OVERVIEW

Chapters 2-4 and 9 develop the basic machinery, and the other chapters cover applications to PDE. Here is a quick overview, with some of the highpoints:
Chapter 2: We start with a quick introduction to symplectic analysis and geometry and their implications for classical Hamiltonian dynamical systems.
Chapter 3: This chapter provides the basics of the Fourier transform and derives also important stationary phase asymptotic estimates, of the sort

$$
I_{h}=(2 \pi h)^{n / 2}\left|\operatorname{det} \partial^{2} \varphi\left(x_{0}\right)\right|^{-1 / 2} e^{\frac{i \pi}{4} \operatorname{sgn} \partial^{2} \varphi\left(x_{0}\right)} e^{\frac{i \varphi\left(x_{0}\right)}{h}} a\left(x_{0}\right)+O\left(h^{\frac{n+2}{2}}\right)
$$

as $h \rightarrow 0$, for the oscillatory integral

$$
I_{h}:=\int_{\mathbb{R}^{n}} e^{\frac{i \varphi}{h}} a d x
$$

We assume here that the gradient $\partial \varphi$ vanishes only at the point $x_{0}$.
Chapter 4: Next we introduce the Weyl quantization $a^{\mathrm{w}}(x, h D)$ of the symbol $a(x, \xi)$ and work out various properties, chief among them the composition formula

$$
a^{\mathrm{w}}(x, h D) \circ b^{\mathrm{w}}(x, h D)=c^{\mathrm{w}}(x, h D)
$$

where the symbol $c:=a \# b$ is computed explictly in terms of $a$ and $b$. We will prove as well the sharp Gårding inequality, stating that if $a$ is a nonnegative symbol, then

$$
\left\langle a^{\mathrm{w}}(x, h D) u, u\right\rangle \geq-C h\|u\|_{L^{2}}^{2}
$$

for all $u$ and sufficiently small $h>0$.
Chapter 5: This section introduces semiclassical defect measures, and uses them to derive decay estimates for the damped wave equation (1.1), where $a \geq 0$, on the flat torus $\mathbb{T}^{n}$.

Chapter 6: In Chapter 6 we begin our study of the eigenvalue problem

$$
P(h) u(h)=E(h) u(h),
$$

for the operator

$$
P(h):=-h^{2} \Delta+V(x)
$$

We prove Weyl's Law for the asymptotic distributions of eigenvalues as $h \rightarrow 0$, stating for all $a<b$ that

$$
\begin{aligned}
\#\{E(h) \mid a \leq & E(h) \leq b\} \\
& =\frac{1}{(2 \pi h)^{n}}\left(\left|\left\{a \leq|\xi|^{2}+V(x) \leq b\right\}\right|+o(1)\right) .
\end{aligned}
$$

Chapter 7: Chapter 7 continues the study of eigenfunctions, first establishing an exponential vanishing theorem in the "classically forbidden" region. We derive as well a Carlemann-type inequality

$$
\|u(h)\|_{L^{2}(E)} \geq e^{-\frac{C}{h}}\|u(h)\|_{L^{2}\left(\mathbb{R}^{n}\right)}
$$

where $E \subset \mathbb{R}^{n}$. This provides a quantitative estimate for quantum mechanical tunneling.
Chapter 8: Chapter 8 concerns the quantum implications of ergodicity for our underlying dynamical systems. A key assertion is that if the underlying dynamical system satisfies an appropriate ergodic condition, then

$$
(2 \pi h)^{n} \sum_{a \leq E_{j} \leq b}\left|\left\langle A u_{j}, u_{j}\right\rangle-f_{\{a \leq p \leq b\}} \sigma(A) d x d \xi\right|^{2} \longrightarrow 0
$$

as $h \rightarrow 0$, for a wide class of pseudodifferential operators $A$. In this expression the classical observable $\sigma(A)$ denotes the symbol of $A$.
Chapter 9: We return in Chapter 9 to the symbol calculus. We introduce the useful formalism of half-densities and use them to illustrate how changing variables in a symbol affects the Weyl quantization. We introduce also the notion of the semiclassical wave front set and show
how a natural localization in phase space leads to pointwise bounds on approximate solutions. We prove as well a semiclassical version of Beals's Theorem, characterizing pseudodifferential operators. As an application we show how quantization commutes with exponentiation at the level of order functions.
Chapter 10: The concluding Chapter 10 explains how to quantize symplectic transformations, with applications including local constructions of propagators, $L^{p}$ bounds on eigenfuctions, and normal forms of differential operators.
Appendices: Appendix A records our notation in one convenient location, and Appendix B is a quick review of differential forms. Appendix C collects various useful functional analysis theorems (with selected proofs). Appendix D discusses Fredholm operators within the framework of Grushin problems, and Appendix E discusses general manifolds and modifications our the symbol calculus to cover pseudodifferential operators on manifolds.

## 2. Symplectic analysis

2.1 Flows
2.2 Symplectic structure on $\mathbb{R}^{2 n}$
2.3 Changing variables
2.4 Hamiltonian vector fields

Since our task in these notes is understanding some interrelationships between dynamics and PDE, we provide in this chapter a quick discussion of the symplectric geometric structure on $\mathbb{R}^{n} \times \mathbb{R}^{n}=\mathbb{R}^{2 n}$ and its interplay with Hamiltonian dynamics.

The reader may wish to first review our basic notation and also the theory of differential forms, set forth respectively in Appendices A and B.

### 2.1 FLOWS

Let $V: \mathbb{R}^{N} \rightarrow \mathbb{R}^{N}$ denote a smooth vector field. Fix a point $z \in \mathbb{R}^{N}$ and solve the ODE

$$
\left\{\begin{array}{l}
\dot{z}(t)=V(z(t)) \quad(t \in \mathbb{R})  \tag{2.1}\\
z(0)=z
\end{array}\right.
$$

We assume that the solution of the flow (2.1) exists and is unique for all times $t \in \mathbb{R}$.

NOTATION. We define

$$
\varphi_{t} z:=z(t)
$$

and sometimes also write

$$
\varphi_{t}=: \exp (t V)
$$

We call $\left\{\varphi_{t}\right\}_{t \in \mathbb{R}}$ the exponential map.

The following lemma records some standard assertions from theory of ordinary differential equations:

## LEMMA 2.1 (Properties of flow map).

(i) $\varphi_{0} z=z$.
(ii) $\varphi_{t+s}=\varphi_{t} \circ \varphi_{s}$ for all $s, t \in \mathbb{R}$.
(iii) For each time $t \in \mathbb{R}$, the mapping $\varphi_{t}: \mathbb{R}^{N} \rightarrow \mathbb{R}^{N}$ is a diffeomorphism, with

$$
\left(\varphi_{t}\right)^{-1}=\varphi_{-t} .
$$

### 2.2 SYMPLECTIC STRUCTURE ON $\mathbb{R}^{2 n}$

We henceforth specialize to the even-dimensional space $\mathbb{R}^{N}=\mathbb{R}^{2 n}=$ $\mathbb{R}^{n} \times \mathbb{R}^{n}$.

NOTATION. We refine our previous notation and henceforth denote an element of $\mathbb{R}^{2 n}$ as

$$
z=(x, \xi)
$$

and interpret $x \in \mathbb{R}^{n}$ as denoting position, $\xi \in \mathbb{R}^{n}$ as momentum. Alternatively, we can think of $\xi$ as belonging to $T_{x}^{*} \mathbb{R}^{n}$, the cotangent space of $\mathbb{R}^{n}$ at $x$. We will likewise write

$$
w=(y, \eta)
$$

for another typical point of $\mathbb{R}^{2 n}$.
We let $\langle\cdot, \cdot\rangle$ denote the usual inner product on $\mathbb{R}^{n}$, and then define this pairing on $\mathbb{R}^{2 n}$ :

DEFINITION. Given $z=(x, \xi), w=(y, \eta)$ on $\mathbb{R}^{2 n}=\mathbb{R}^{n} \times \mathbb{R}^{n}$, define their symplectic product

$$
\begin{equation*}
\sigma(z, w):=\langle\xi, y\rangle-\langle x, \eta\rangle . \tag{2.2}
\end{equation*}
$$

Note that

$$
\begin{equation*}
\sigma(z, w)=\langle J z, w\rangle \tag{2.3}
\end{equation*}
$$

for the $2 n \times 2 n$ matrix

$$
J:=\left(\begin{array}{cc}
O & I  \tag{2.4}\\
-I & O
\end{array}\right)
$$

Observe

$$
J^{2}=-I, \quad J^{T}=-J .
$$

We will later in Section 10.2 interpret the transformation $J$ as the classical analog of the Fourier transform.

LEMMA 2.2 (Properties of $\boldsymbol{\sigma}$ ). The bilinear form $\sigma$ is antisymmetric:

$$
\sigma(z, w)=-\sigma(w, z)
$$

and nondegenerate:

$$
\text { if } \sigma(z, w)=0 \text { for all } w \text {, then } z=0 \text {. }
$$

These assertions are straightforward to check.
NOTATION. We now bring in the terminology of differential forms, reviewed in Appendix B. Using the notation discussed above, we introduce for $x=\left(x_{1}, \ldots, x_{n}\right)$ and $\xi=\left(\xi_{1}, \ldots, \xi_{n}\right)$ the 1 -forms $d x_{j}$ and $d \xi_{j}$ for $j=1, \ldots, n$. We then can write

$$
\begin{equation*}
\sigma=d \xi \wedge d x=\sum_{j=1}^{n} d \xi_{j} \wedge d x_{j} \tag{2.5}
\end{equation*}
$$

Observe also

$$
\begin{equation*}
\sigma=d \omega \quad \text { for } \omega:=\xi d x=\sum_{j=1}^{n} \xi_{j} d x_{j} . \tag{2.6}
\end{equation*}
$$

It follows that

$$
\begin{equation*}
d \sigma=0 \tag{2.7}
\end{equation*}
$$

### 2.3 CHANGING VARIABLES.

Suppose next that $U, V \subseteq \mathbb{R}^{2 n}$ are open sets and

$$
\kappa: U \rightarrow V
$$

is a smooth mapping. We will write

$$
\gamma(x, \xi)=(y, \eta)=(y(x, \xi), \eta(x, \xi))
$$

DEFINITION. We call $\gamma$ a symplectic mapping, or a symplectomorphism, provided

$$
\begin{equation*}
\gamma^{*} \sigma=\sigma \tag{2.8}
\end{equation*}
$$

Here the pull-back $\gamma^{*} \sigma$ of the symplectic product $\sigma$ is defined by

$$
\left(\gamma^{*} \sigma\right)(z, w):=\sigma\left(\gamma_{*}(z), \gamma_{*}(w)\right),
$$

$\gamma_{*}$ denoting the push-forward of vectors: see Appendix B.
NOTATION. We will usually write (2.8) in the more suggestive notation

$$
\begin{equation*}
d \eta \wedge d y=d \xi \wedge d x \tag{2.9}
\end{equation*}
$$

EXAMPLE 1: Linear symplectic mappings. Suppose $\kappa: \mathbb{R}^{2 n} \rightarrow$ $\mathbb{R}^{2 n}$ is linear:

$$
\kappa(x, \xi)=\left(\begin{array}{ll}
A & B \\
C & D
\end{array}\right)\binom{x}{\xi}=(A x+B \xi, C x+D \xi)=(y, \eta)
$$

where $A, B, C, D$ are $n \times n$ matrices.

THEOREM 2.3 (Symplectic matrices). The linear mapping $\gamma$ is symplectic if and only if the matrix

$$
M:=\left(\begin{array}{ll}
A & B \\
C & D
\end{array}\right)
$$

satisfies

$$
\begin{equation*}
M^{T} J M=J \tag{2.10}
\end{equation*}
$$

DEFINITION. We call a $2 n \times 2 n$ matrix $M$ symplectic if (2.10) holds.
In particular the linear mapping $(x, \xi) \mapsto(\xi,-x)$ determined by $J$ is symplectic.

Proof. Let us compute

$$
\begin{aligned}
d \eta \wedge d y & =(C d x+D d \xi) \wedge(A d x+B d \xi) \\
& =A^{T} C d x \wedge d x+B^{T} D d \xi \wedge d \xi+\left(A^{T} D-C^{T} B\right) d \xi \wedge d x \\
& =d \xi \wedge d x
\end{aligned}
$$

if and only if

$$
\begin{equation*}
A^{T} C \text { and } B^{T} D \text { are symmetric, } A^{T} D-C^{T} B=I \tag{2.11}
\end{equation*}
$$

Then

$$
\begin{aligned}
M^{T} J M & =\left(\begin{array}{ll}
A^{T} & C^{T} \\
B^{T} & D^{T}
\end{array}\right)\left(\begin{array}{cc}
O & I \\
-I & O
\end{array}\right)\left(\begin{array}{ll}
A & B \\
C & D
\end{array}\right) \\
& =\left(\begin{array}{ll}
A^{T} C-C^{T} A & A^{T} D-C^{T} B \\
B^{T} C-D^{T} A & B^{T} D-D^{T} B
\end{array}\right) \\
& =J
\end{aligned}
$$

if and only if (2.11) holds.
EXAMPLE 2: Nonlinear symplectic mappings. Assume next that $\kappa: \mathbb{R}^{2 n} \rightarrow \mathbb{R}^{2 n}$ is nonlinear:

$$
\kappa(x, \xi)=(y, \eta)
$$

for smooth functions $y=y(x, \xi), \eta=\eta(x, \xi)$. Its linearization is the $2 n \times 2 n$ matrix

$$
\partial \kappa=\partial_{x, \xi} \kappa=\left(\begin{array}{ll}
\partial_{x} y & \partial_{\xi} y \\
\partial_{x} \eta & \partial_{\xi} \eta
\end{array}\right) .
$$

THEOREM 2.4 (Symplectic transformations). The mapping $\kappa$ is symplectic if and only if the matrix $\partial \kappa$ is symplectic at each point.

Proof. We have

$$
d \eta \wedge d y=(C d x+D d \xi) \wedge(A d x+B d \xi)
$$

for

$$
A:=\partial_{x} y, B:=\partial_{\xi} y, C:=\partial_{x} \eta, D:=\partial_{\xi} \eta .
$$

Consequently, as in the previous proof, we have $d \eta \wedge d y=d \xi \wedge d x$ if and only if (2.11) is valid, which in turn is so if and only if $\partial \kappa$ is a symplectic matrix.

EXAMPLE 3: Lifting diffeomorphisms. Let

$$
\gamma: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}
$$

be a diffeomorphism on $\mathbb{R}^{n}$, with nondegenerate Jacobian matrix $\partial_{x} \gamma$. We propose to extend $\gamma$ to a symplectomorphism

$$
\kappa: \mathbb{R}^{2 n} \rightarrow \mathbb{R}^{2 n}
$$

having the form

$$
\begin{equation*}
\kappa(x, \xi)=(\gamma(x), \eta(x, \xi))=(y, \eta) \tag{2.12}
\end{equation*}
$$

by "lifting" to the momentum variables.

THEOREM 2.5 (Extending to a symplectic mapping). The transformation (2.12) is symplectic for

$$
\begin{equation*}
\eta(x, \xi):=\left[\partial_{x} \gamma(x)^{-1}\right]^{T} \xi . \tag{2.13}
\end{equation*}
$$

Proof. It turns out to be easier to look for $\xi$ as a function of $x$ and $\eta$. We compute

$$
d y=A d x, d \xi=E d x+F d \eta
$$

for

$$
A:=\partial_{x} y, \quad E:=\partial_{x} \xi, \quad F:=\partial_{\eta} \xi
$$

Therefore

$$
d \eta \wedge d y=d \eta \wedge(A d x)
$$

and

$$
d \xi \wedge d x=(E d x \wedge F d \eta) \wedge d x=E d x \wedge d x+d \eta \wedge F^{T} d x
$$

We would like to construct $\xi=\xi(x, \eta)$ so that

$$
A=F^{T} \quad \text { and } \quad E \text { is symmetric, }
$$

the latter condition implying that $E d x \wedge d x=0$. To do so, let us define

$$
\xi(x, \eta):=\left(\partial_{x} \gamma\right)^{T} \eta
$$

Then clearly $F^{T}=A$, and $E=E^{T}=\left(\left(\gamma_{x_{i} x_{j}}\right)\right)$, as required.

INTERPRETATION: This example will prove useful later, when we quantize symbols in Chapter 4 and learn that the partial differential operator

$$
P(h)=-h^{2} \Delta
$$

is associated with the symbol $p(x, \xi)=|\xi|^{2}$. If we change variables $y=$ $\gamma(x)$, it is natural to ask how $P(h)$ transforms. Now $\partial_{x}=\left(\partial_{x} \gamma\right)^{T} \partial_{y}$, and so

$$
P(h)=-h^{2}\left(\partial_{x} \gamma\right)^{T} \partial_{y}\left(\left(\partial_{x} \gamma\right)^{T} \partial_{y}\right)
$$

We will see later that the operator on the right is associated with the symbol

$$
\left\langle\left(\partial_{x} \gamma\right)^{T} \eta,\left(\partial_{x} \gamma\right)^{T} \eta\right\rangle .
$$

All this is consistent with the transformation (2.13).
Here by the way is an instance of another general principle: if we change variables in a symbol, we preserve the principal symbol, up to higher order terms.

EXAMPLE 4: Generating functions. Our last example demonstrates that we can, locally at least, build a symplectic transformation from a real-valued generating function.

Suppose $\varphi: \mathbb{R}^{n} \times \mathbb{R}^{n} \rightarrow \mathbb{R}, \varphi=\varphi(x, y)$, is smooth. Assume also that

$$
\begin{equation*}
\operatorname{det}\left(\partial_{x y}^{2} \varphi\left(x_{0}, y_{0}\right)\right) \neq 0 \tag{2.14}
\end{equation*}
$$

Define

$$
\begin{equation*}
\xi=\partial_{x} \varphi, \eta=-\partial_{y} \varphi \tag{2.15}
\end{equation*}
$$

and observe that the Implicit Function Theorem implies $(y, \eta)$ is a smooth function of $(x, \xi)$ near $\left(x_{0}, \partial_{x} \varphi\left(x_{0}, y_{0}\right)\right)$.

## THEOREM 2.6 (Generating functions and symplectic maps).

 The mapping $\gamma$ defined by$$
\begin{equation*}
\left(x, \partial_{x} \varphi(x, y)\right) \mapsto\left(y,-\partial_{y} \varphi(x, y)\right) \tag{2.16}
\end{equation*}
$$

is a symplectomorphism near $\left(x_{0}, \xi_{0}\right)$.

Proof. We compute

$$
\begin{aligned}
d \eta \wedge d y & =d\left(-\partial_{y} \varphi\right) \wedge d y \\
& =\left[\left(-\partial_{y}^{2} \varphi d y\right) \wedge d y\right]+\left[\left(-\partial_{x y}^{2} \varphi d x\right) \wedge d y\right] \\
& =-\left(\partial_{x y}^{2} \varphi\right) d x \wedge d y
\end{aligned}
$$

since $\partial_{y}^{2} \varphi$ is symmetric. Likewise,

$$
\begin{aligned}
d \xi \wedge d x & =d\left(\partial_{x} \varphi\right) \wedge d x \\
& =\left[\left(\partial_{x}^{2} \varphi d x\right) \wedge d x\right]+\left[\left(\partial_{x y}^{2} \varphi d y\right) \wedge d x\right] \\
& =-\left(\partial_{x y}^{2} \varphi\right) d x \wedge d y=d \eta \wedge d y
\end{aligned}
$$

TERMINOLOGY. In Greek, the word "symplectic" means "intertwined", This is consistent with Example 4, since the generating function $\varphi=\varphi(x, y)$ is a function of a mixture of half of the original variables $(x, \xi)$ and half of the new variables $(y, \eta)$. "Symplectic" can also be interpreted as "complex", mathematical usage due to Hermann Weyl who renamed "line complex group" the "symplectic group": see Cannas da Silva [CdS].

APPLICATION: Lagrangian submanifolds. A Lagrangian submanifold $\Lambda$ is an $n$-dimensional submanifold of $\mathbb{R}^{2 n}$ for which

$$
\left.\sigma\right|_{\Lambda}=0
$$

Then

$$
\left.d \omega\right|_{\Lambda}=\left.\sigma\right|_{\Lambda}=0
$$

and so according to Poincaré's Theorem B.4, we locally have

$$
\omega=d \varphi,
$$

for some smooth function $\varphi$ on $\Lambda$. We will exploit this observation in Section 10.2.

### 2.4 HAMILTONIAN VECTOR FIELDS

DEFINITION. Given $f \in C^{\infty}\left(\mathbb{R}^{2 n}\right)$, we define the corresponding Hamiltonian vector field by requiring

$$
\begin{equation*}
\sigma\left(z, H_{f}\right)=d f(z) \quad \text { for all } z=(x, \xi) \tag{2.17}
\end{equation*}
$$

This is well defined, since $\sigma$ is nondegenerate. We can write explicitly that

$$
\begin{equation*}
H_{f}=\left\langle\partial_{\xi} f, \partial_{x}\right\rangle-\left\langle\partial_{x} f, \partial_{\xi}\right\rangle=\sum_{j=1}^{n} f_{\xi_{j}} \partial_{x_{j}}-f_{x_{j}} \partial_{\xi_{j}} \tag{2.18}
\end{equation*}
$$

LEMMA 2.7 (Differentials and Hamiltonian vector fields). We have the relation

$$
\begin{equation*}
\left.d f=-\left(H_{f}\right\lrcorner \sigma\right) \tag{2.19}
\end{equation*}
$$

for the contraction $\lrcorner$ defined in Appendix $B$.

Proof. We calculate for each $z$ that

$$
\left.\left(H_{f}\right\lrcorner \sigma\right)(z)=\sigma\left(H_{f}, z\right)=-\sigma\left(z, H_{f}\right)=-d f(z) .
$$

DEFINITION. If $f, g \in C^{\infty}\left(\mathbb{R}^{2 n}\right)$, we define their Poisson bracket

$$
\begin{equation*}
\{f, g\}:=H_{f} g=\sigma(\partial f, \partial g) \tag{2.20}
\end{equation*}
$$

That is,

$$
\begin{equation*}
\{f, g\}=\left\langle\partial_{\xi} f, \partial_{x} g\right\rangle-\left\langle\partial_{x} f, \partial_{\xi} g\right\rangle=\sum_{j=1}^{n} f_{\xi_{j}} g_{x_{j}}-f_{x_{j}} g_{\xi_{j}} \tag{2.21}
\end{equation*}
$$

## LEMMA 2.8 (Brackets, commutators).

(i) We have Jacobi's identity

$$
\begin{equation*}
\{f,\{g, h\}\}+\{g,\{h, f\}\}+\{h,\{f, g\}\}=0 \tag{2.22}
\end{equation*}
$$

for all functions $f, g, h \in C^{\infty}\left(\mathbb{R}^{2 n}\right)$.
(ii) Furthermore,

$$
\begin{equation*}
H_{\{f, g\}}=\left[H_{f}, H_{g}\right] . \tag{2.23}
\end{equation*}
$$

Proof. 1. A direct calculation verifies assertion (i). For an alternative proof, note that Lemma B. 1 provides the identity

$$
\begin{align*}
0= & d \sigma\left(H_{f}, H_{g}, H_{h}\right) \\
= & H_{f} \sigma\left(H_{g}, H_{h}\right)+H_{g} \sigma\left(H_{h}, H_{f}\right)+H_{h} \sigma\left(H_{f}, H_{g}\right)  \tag{2.24}\\
& -\sigma\left(\left[H_{f}, H_{g}\right], H_{h}\right)-\sigma\left(\left[H_{g}, H_{h}\right], H_{f}\right)-\sigma\left(\left[H_{h}, H_{f}\right], H_{g}\right),
\end{align*}
$$

since $d \sigma=0$. Now (2.20) implies

$$
H_{f} \sigma\left(H_{g}, H_{h}\right)=\{f,\{g, h\}\}
$$

and

$$
\begin{aligned}
\sigma\left(\left[H_{f}, H_{g}\right], H_{h}\right) & =\left[H_{f}, H_{g}\right] h=H_{f} H_{g} h-H_{g} H_{f} h \\
& =\{f,\{g, h\}\}-\{g,\{f, h\}\}
\end{aligned}
$$

Similar identities hold for other terms. Substituting into (2.24) gives Jacobi's identity.
2. We observe that

$$
H_{\{f, g\}} h=\left[H_{f}, H_{g}\right] h
$$

is a rewriting of (2.22).

THEOREM 2.9 (Jacobi's Theorem). If $\gamma$ is a symplectomorphism, then

$$
\begin{equation*}
H_{f}=\gamma_{*}\left(H_{\gamma^{*} f}\right) \tag{2.25}
\end{equation*}
$$

In other words, the pull-back of a Hamiltonian vector field generated by $f$,

$$
\begin{equation*}
\gamma^{*} H_{f}:=\left(\gamma^{-1}\right)_{*} H_{f}, \tag{2.26}
\end{equation*}
$$

is the Hamiltonian vector field generated by the pull-back of $f$.
Proof. Using the notation of (2.26),

$$
\begin{aligned}
\left.\gamma^{*}\left(H_{f}\right)\right\lrcorner \sigma & \left.\left.=\gamma^{*}\left(H_{f}\right)\right\lrcorner \gamma^{*} \sigma=\gamma^{*}\left(H_{f}\right\lrcorner \sigma\right) \\
& =-\gamma^{*}(d f)=-d\left(\gamma^{*} f\right) \\
& \left.=H_{\gamma^{*} f}\right\lrcorner \sigma .
\end{aligned}
$$

Since $\sigma$ is nondegenerate, (2.25) follows.
EXAMPLE. Define $\gamma=J$, so that $\gamma(x, \xi)=(\xi,-x)$; and recall $\gamma$ is a symplectomorphism. We have $\gamma^{*} f(x, \xi)=f(\xi,-x)$, and therefore

$$
H_{\gamma^{*} f}=\left\langle\partial_{x} f(\xi,-x), \partial_{x}\right\rangle+\left\langle\partial_{\xi} f(\xi,-x), \partial_{\xi}\right\rangle .
$$

Then

$$
\kappa^{*} H_{f}=\left\langle\partial_{\xi} f(\xi,-x), \partial_{\xi}\right\rangle-\left\langle\partial_{x} f(\xi,-x), \partial_{-x}\right\rangle=H_{\gamma^{*} f}
$$

THEOREM 2.10 (Hamiltonian flows as symplectomorphisms).
If $f$ is smooth, then for each time $t$, the mapping

$$
(x, \xi) \mapsto \varphi_{t}(x, \xi)=\exp \left(t H_{f}\right)
$$

is a symplectomorphism.

Proof. According to Cartan's formula (Theorem B.3), we have

$$
\left.\left.\frac{d}{d t}\left(\left(\varphi_{t}\right)^{*} \sigma\right)=\mathcal{L}_{H_{f}} \sigma=d\left(H_{f}\right\lrcorner \sigma\right)+\left(H_{f}\right\lrcorner d \sigma\right) .
$$

Since $d \sigma=0$, it follows that

$$
\frac{d}{d t}\left(\left(\varphi_{t}\right)^{*} \sigma\right)=d(-d f)=-d^{2} f=0
$$

Thus $\left(\varphi_{t}\right)^{*} \sigma=\sigma$ for all times $t$.

THEOREM 2.11 (Darboux's Theorem). Let $U$ be a neighborhood of $\left(x_{0}, \xi_{0}\right)$ and suppose $\eta$ is a nondegenerate 2 -form defined on $U$, satisfying

$$
d \eta=0
$$

Then near $\left(x_{0}, \xi_{0}\right)$ there exists a diffeomorphism $\gamma$ such that

$$
\begin{equation*}
\gamma^{*} \eta=\sigma \tag{2.27}
\end{equation*}
$$

This means that all symplectic structures are identical locally, in the sense that all are equivalent to that generated by $\sigma$.

Proof. 1. Let us assume $\left(x_{0}, \xi_{0}\right)=(0,0)$. We first find a linear mapping $L$ so that

$$
L^{*} \eta(0,0)=\sigma(0,0)
$$

This means that we find a basis $\left\{e_{k}, f_{k}\right\}_{k=1}^{n}$ of $\mathbb{R}^{2 n}$ such that

$$
\left\{\begin{array}{l}
\eta\left(f_{l}, e_{k}\right)=\delta_{k l} \\
\eta\left(e_{k}, e_{l}\right)=0 \\
\eta\left(f_{k}, f_{l}\right)=0
\end{array}\right.
$$

for all $1 \leq k, l \leq n$. Then if $u=\sum_{i=1}^{n} x_{i} e_{i}+\xi_{i} f_{i}, v=\sum_{j=1}^{n} y_{j} e_{j}+\eta_{j} f_{j}$, we have

$$
\begin{aligned}
\eta(u, v) & =\sum_{i, j=1}^{n} x_{i} y_{j} \eta\left(e_{i}, e_{j}\right)+\xi_{i} \eta_{j} \eta\left(f_{i}, f_{j}\right)+x_{i} \eta_{j} \sigma\left(e_{i}, f_{j}\right)+\xi_{i} y_{j} \sigma\left(f_{i}, e_{j}\right) \\
& =\langle\xi, y\rangle-\langle x, \eta\rangle=\sigma((x, \xi),(y, \eta))
\end{aligned}
$$

2. Next, define $\eta_{t}:=t \eta+(1-t) \sigma$ for $0 \leq t \leq 1$. Our intention is to find $\gamma_{t}$ so that $\gamma_{t}^{*} \eta_{t}=\sigma$ near $(0,0)$; then $\gamma:=\gamma_{1}$ solves our problem. We will construct $\gamma_{t}$ by solving the flow

$$
\left\{\begin{array}{l}
\dot{z}(t)=V_{t}(z(t)) \quad(0 \leq t \leq 1)  \tag{2.28}\\
z(0)=z
\end{array}\right.
$$

and setting $\gamma_{t}:=\varphi_{t}$.
For this to work, we must design the vector fields $V_{t}$ in (2.28) so that $\frac{d}{d t}\left(\gamma_{t}^{*} \eta_{t}\right)=0$. Let us therefore calculate

$$
\begin{aligned}
\frac{d}{d t}\left(\gamma_{t}^{*} \eta_{t}\right) & =\kappa_{t}^{*}\left(\frac{d}{d t} \eta_{t}\right)+\gamma_{t}^{*} \mathcal{L}_{V_{t}} \eta_{t} \\
& \left.\left.=\gamma_{t}^{*}\left[(\eta-\sigma)+d\left(V_{t}\right\lrcorner \eta_{t}\right)+V_{t}\right\lrcorner d \eta_{t}\right]
\end{aligned}
$$

where we used Cartan's formula, Theorem B.3. Note that $d \eta_{t}=t d \eta+$ $(1-t) d \sigma$. Hence $\frac{d}{d t}\left(\gamma_{t}^{*} \eta_{t}\right)=0$ provided

$$
\begin{equation*}
\left.(\eta-\sigma)+d\left(V_{t}\right\lrcorner \eta_{t}\right)=0 \tag{2.29}
\end{equation*}
$$

According to Poincaré's Theorem B.4, we can write

$$
\eta-\sigma=d \alpha \quad \text { near }(0,0)
$$

So (2.29) will hold, provided

$$
\begin{equation*}
\left.V_{t}\right\lrcorner \eta_{t}=-\alpha \quad(0 \leq t \leq 1) . \tag{2.30}
\end{equation*}
$$

Since $\eta=\sigma$ at $(0,0), \eta_{t}=\sigma$ at $(0,0)$. In particular, $\eta_{t}$ is nondegenerate for $0 \leq t \leq 1$ in a neighbourhood of ( 0,0 ), and hence we can solve (2.29) for the vector field $V_{t}$.

## 3. Fourier transform, stationary phase

3.1 Fourier transform on $\mathcal{S}$
3.2 Fourier transform on $\mathcal{S}^{\prime}$
3.3 Semiclassical Fourier transform
3.4 Stationary phase in one dimension
3.5 Stationary phase in higher dimensions
3.6 Important examples

We discuss in this chapter how to define the Fourier transform $\mathcal{F}$ and its inverse $\mathcal{F}^{-1}$ on various classes of smooth functions and nonsmooth distributions. We introduce also the rescaled semiclassical transforms $\mathcal{F}_{h}, \mathcal{F}_{h}^{-1}$ depending on the small parameter $h$, and develop stationary phase asymptotics to help us understand various formulas involving $\mathcal{F}_{h}$ in the limit as $h \rightarrow 0$.

### 3.1 FOURIER TRANSFORM ON $\mathcal{S}$

We begin by defining and investigating the Fourier transform of smooth functions that decay rapidly as $|x| \rightarrow \infty$.

DEFINITIONS (i) The Schwartz space is

$$
\begin{aligned}
\mathcal{S}=\mathcal{S}\left(\mathbb{R}^{n}\right) & := \\
& \left\{\varphi \in C^{\infty}\left(\mathbb{R}^{n}\right)\left|\sup _{\mathbb{R}^{n}}\right| x^{\alpha} \partial^{\beta} \varphi \mid<\infty \text { for all multiindices } \alpha, \beta\right\} .
\end{aligned}
$$

(ii) We say

$$
\varphi_{j} \rightarrow \varphi \quad \text { in } \mathcal{S}
$$

provided

$$
\sup _{\mathbb{R}^{n}}\left|x^{\alpha} \partial^{\beta}\left(\varphi_{j}-\varphi\right)\right| \rightarrow 0
$$

for all multiindices $\alpha, \beta$.

DEFINITION. If $\varphi \in \mathcal{S}$, define the Fourier transform

$$
\begin{equation*}
\mathcal{F} \varphi(\xi)=\hat{\varphi}(\xi):=\int_{\mathbb{R}^{n}} e^{-i\langle x, \xi\rangle} \varphi(x) d x \quad\left(\xi \in \mathbb{R}^{n}\right) \tag{3.1}
\end{equation*}
$$

The reader is warned that many other texts use slightly different definitions, entailing normalizing factors involving $\pi$.

## EXAMPLE: Exponential of a real quadratic form.

THEOREM 3.1 (Transform of a real exponential). Let $Q$ be a real, symmetric, positive definite $n \times n$ matrix. Then

$$
\begin{equation*}
\mathcal{F}\left(e^{-\frac{1}{2}\langle Q x, x\rangle}\right)=\frac{(2 \pi)^{n / 2}}{(\operatorname{det} Q)^{1 / 2}} e^{-\frac{1}{2}\left\langle Q^{-1} \xi, \xi\right\rangle} . \tag{3.2}
\end{equation*}
$$

Proof. Let us calculate

$$
\begin{aligned}
\mathcal{F}\left(e^{-\frac{1}{2}\langle Q x, x\rangle}\right) & =\int_{\mathbb{R}^{n}} e^{-\frac{1}{2}\langle Q x, x\rangle-i\langle x, \xi\rangle} d x \\
& =\int_{\mathbb{R}^{n}} e^{-\frac{1}{2}\left\langle Q\left(x+i Q^{-1} \xi\right), x+i Q^{-1} \xi\right\rangle} e^{-\frac{1}{2}\left\langle Q^{-1} \xi, \xi\right\rangle} d x \\
& =e^{-\frac{1}{2}\left\langle Q^{-1} \xi, \xi\right\rangle} \int_{\mathbb{R}^{n}} e^{-\frac{1}{2}\langle Q y, y\rangle} d y .
\end{aligned}
$$

We compute the last integral by making an orthogonal change of variables that converts $Q$ into diagonal form $\operatorname{diag}\left(\lambda_{1}, \ldots, \lambda_{n}\right)$. Then

$$
\begin{aligned}
\int_{\mathbb{R}^{n}} e^{-\frac{1}{2}\langle Q y, y\rangle} d y & =\int_{\mathbb{R}^{n}} e^{-\frac{1}{2} \sum_{k=1}^{n} \lambda_{k} w_{k}^{2}} d w=\prod_{k=1}^{n} \int_{-\infty}^{\infty} e^{-\frac{\lambda_{k}}{2} w^{2}} d w \\
& =\prod_{k=1}^{n} \frac{2^{1 / 2}}{\lambda_{k}^{1 / 2}} \int_{-\infty}^{\infty} e^{-y^{2}} d y \\
& =\frac{(2 \pi)^{n / 2}}{\left(\lambda_{1} \cdots \lambda_{n}\right)^{1 / 2}}=\frac{(2 \pi)^{n / 2}}{(\operatorname{det} Q)^{1 / 2}} .
\end{aligned}
$$

The Fourier transform $\mathcal{F}$ lets us move from position variables $x$ to momentum variables $\xi$, and we need to catalog how it converts various algebraic and analytic expressions in $x$ into related expressions in $\xi$ :

## THEOREM 3.2 (Properties of Fourier transform).

(i) The mapping $\mathcal{F}: \mathcal{S} \rightarrow \mathcal{S}$ is an isomorphism.
(ii) We have the Fourier inversion formula

$$
\begin{equation*}
\mathcal{F}^{-1}=\frac{1}{(2 \pi)^{n}} R \circ \mathcal{F}, \tag{3.3}
\end{equation*}
$$

where $R f(x):=f(-x)$. In other words,

$$
\begin{equation*}
\mathcal{F}^{-1} \psi(x)=\frac{1}{(2 \pi)^{n}} \int_{\mathbb{R}^{n}} e^{i\langle x, \xi\rangle} \psi(\xi) d \xi ; \tag{3.4}
\end{equation*}
$$

and therefore

$$
\begin{equation*}
\varphi(x)=\frac{1}{(2 \pi)^{n}} \int_{\mathbb{R}^{n}} e^{i\langle x, \xi\rangle} \hat{\varphi}(\xi) d \xi \tag{3.5}
\end{equation*}
$$

(iii) In addition,

$$
\begin{equation*}
D_{\xi}^{\alpha}(\mathcal{F} \varphi)=\mathcal{F}\left((-x)^{\alpha} \varphi\right) \tag{3.6}
\end{equation*}
$$

and

$$
\begin{equation*}
\mathcal{F}\left(D_{x}^{\alpha} \varphi\right)=\xi^{\alpha} \mathcal{F} \varphi \tag{3.7}
\end{equation*}
$$

(iv) Furthermore,

$$
\begin{equation*}
\mathcal{F}(\varphi \psi)=\frac{1}{(2 \pi)^{n}} \mathcal{F}(\varphi) * \mathcal{F}(\psi) \tag{3.8}
\end{equation*}
$$

REMARKS. (i) In these formulas we employ the notation from Appendix A that

$$
D^{\alpha}=\frac{1}{i^{|\alpha|}} \partial^{\alpha}
$$

(ii) We will later interpret the Fourier inversion formula (3.4) as saying that

$$
\begin{equation*}
\delta_{\{y=x\}}=\frac{1}{(2 \pi)^{n}} \int_{\mathbb{R}^{n}} e^{i\langle x-y, \xi\rangle} d \xi \quad \text { in the sense of distributions, } \tag{3.9}
\end{equation*}
$$

$\delta$ denoting the Dirac measure.

Proof. 1. Let us calculate for $\varphi \in \mathcal{S}$ that

$$
\begin{aligned}
D_{\xi}^{\alpha}(\mathcal{F} \varphi) & =D_{\xi}^{\alpha} \int_{\mathbb{R}^{n}} e^{-i\langle x, \xi\rangle} \varphi(x) d x=\frac{1}{i^{\alpha}} \int_{\mathbb{R}^{n}} e^{-i\langle x, \xi\rangle}(-i x)^{\alpha} \varphi(x) d x \\
& =\int_{\mathbb{R}^{n}} e^{-i\langle x, \xi\rangle}(-x)^{\alpha} \varphi(x) d x=\mathcal{F}\left((-x)^{\alpha} \varphi\right)
\end{aligned}
$$

Likewise,

$$
\begin{aligned}
\mathcal{F}\left(D_{x}^{\alpha} \varphi\right) & =\int_{\mathbb{R}^{n}} e^{-i\langle x, \xi\rangle} D_{x}^{\alpha} \varphi d x=(-1)^{|\alpha|} \int_{\mathbb{R}^{n}} D_{x}^{\alpha}\left(e^{-i\langle x, \xi\rangle}\right) \varphi d x \\
& =(-1)^{|\alpha|} \int_{\mathbb{R}^{n}} \frac{1}{i^{|\alpha|}}(-i \xi)^{\alpha} e^{-i\langle x, \xi\rangle} \varphi d x=\xi^{\alpha}(\mathcal{F} \varphi) .
\end{aligned}
$$

This proves (iii).
2. Recall from Appendix A the notation $\langle x\rangle=\left(1+|x|^{2}\right)^{\frac{1}{2}}$. Then for all multiindices $\alpha, \beta$, we have

$$
\begin{aligned}
\sup _{\xi}\left|\xi^{\beta} D_{\xi}^{\alpha} \hat{\varphi}\right| & =\sup _{\xi}\left|\xi^{\beta} \mathcal{F}\left((-x)^{\alpha} \varphi\right)\right| \\
& =\sup _{\xi} \mid \mathcal{F}\left(D_{x}^{\beta}\left((-x)^{\alpha} \varphi\right) \mid\right. \\
& =\sup _{\xi}\left|\int_{\mathbb{R}^{n}} e^{-i\langle x, \xi\rangle} \frac{1}{\langle x\rangle^{n+1}}\langle x\rangle^{n+1} D_{x}^{\beta}\left((-x)^{\alpha} \varphi\right) d x\right| \\
& \leq \sup _{x}\left|\langle x\rangle^{n+1} D_{x}^{\beta}\left((-x)^{\alpha} \varphi\right)\right| \int_{\mathbb{R}^{n}} \frac{d x}{\langle x\rangle^{n+1}}<\infty .
\end{aligned}
$$

Hence $\mathcal{F}: \mathcal{S} \rightarrow \mathcal{S}$, and a similar calculation shows that $\varphi_{i} \rightarrow \varphi$ in $\mathcal{S}$ implies $\mathcal{F}\left(\varphi_{j}\right) \rightarrow \mathcal{F}(\varphi)$.
3. To show $\mathcal{F}$ is invertible, note that

$$
\begin{aligned}
R \circ \mathcal{F} \circ \mathcal{F} \circ D_{x_{j}} & =R \circ \mathcal{F} \circ M_{\xi_{j}} \circ \mathcal{F} \\
& =R \circ\left(-D_{x_{j}}\right) \circ \mathcal{F} \circ \mathcal{F} \\
& =D_{x_{j}} \circ R \circ \mathcal{F} \circ \mathcal{F},
\end{aligned}
$$

where $M_{\xi_{j}}$ denotes multiplication by $\xi_{j}$. Thus $R \circ \mathcal{F} \circ \mathcal{F}$ commutes with $D_{x_{j}}$ and it likewise commutes with the multiplication operator $M_{\lambda}$. According to Lemma 3.3, stated and proved below, $R \circ \mathcal{F} \circ \mathcal{F}$ is a multiple of the identity operator:

$$
\begin{equation*}
R \circ \mathcal{F} \circ \mathcal{F}=c I \tag{3.10}
\end{equation*}
$$

From the example above, we know that

$$
\mathcal{F}\left(e^{-\frac{|x|^{2}}{2}}\right)=(2 \pi)^{n / 2} e^{-\frac{|\xi|^{2}}{2}} .
$$

Thus $\mathcal{F}\left(e^{-\frac{|\xi|^{2}}{2}}\right)=(2 \pi)^{n / 2} e^{-\frac{|x|^{2}}{2}}$. Consequently $c=(2 \pi)^{n}$, and hence

$$
\mathcal{F}^{-1}=\frac{1}{(2 \pi)^{n}} R \circ \mathcal{F} .
$$

4. Lastly, since

$$
\varphi(x)=\frac{1}{(2 \pi)^{n}} \int_{\mathbb{R}^{n}} e^{i\langle x, \xi\rangle} \hat{\varphi}(\xi) d \xi, \quad \psi(x)=\frac{1}{(2 \pi)^{n}} \int_{\mathbb{R}^{n}} e^{i\langle x, \eta\rangle} \hat{\psi}(\eta) d \eta
$$

we have

$$
\begin{aligned}
\varphi \psi & =\frac{1}{(2 \pi)^{2 n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{i\langle x, \xi+\eta\rangle} \hat{\varphi}(\xi) \hat{\psi}(\eta) d \xi d \eta \\
& =\frac{1}{(2 \pi)^{2 n}} \int_{\mathbb{R}^{n}} e^{i\langle x, \rho\rangle}\left(\int_{\mathbb{R}^{n}} \hat{\varphi}(\xi) \hat{\psi}(\rho-\xi) d \rho\right) d \xi \\
& =\frac{1}{(2 \pi)^{n}} \mathcal{F}^{-1}(\hat{\varphi} * \hat{\psi}) .
\end{aligned}
$$

But $\varphi \psi=\mathcal{F}^{-1} \mathcal{F}(\varphi \psi)$, and so assertion (iv) follows.

LEMMA 3.3 (Commutativity). Let $M_{f}: g \mapsto f g$ be the multiplication operator. Suppose that $L: \mathcal{S} \rightarrow \mathcal{S}$ is linear, and that

$$
\begin{equation*}
L \circ M_{x_{j}}=M_{x_{j}} \circ L, L \circ D_{x_{j}}=D_{x_{j}} \circ L \tag{3.11}
\end{equation*}
$$

$j=1, \ldots, n$. Then

$$
L=c I
$$

for some constant $c$, where I denotes the identity operator.

Proof. 1. Choose $\varphi \in \mathcal{S}$, fix $y \in \mathbb{R}^{n}$, and write

$$
\varphi(x)-\varphi(y)=\sum_{j=1}^{n}\left(x_{j}-y_{j}\right) \psi_{j}(x)
$$

for

$$
\psi_{j}(x):=\int_{0}^{1} \varphi_{x_{j}}(y+t(x-y)) d t
$$

Since typically $\psi_{j} \notin \mathcal{S}$, we select a smooth function $\chi$ with compact support such that $\chi \equiv 1$ for $x$ near $y$. Write

$$
\varphi_{j}(x):=\chi(x) \psi_{j}(x)+\frac{\left(x_{j}-y_{j}\right)}{|x-y|^{2}}(1-\chi(x)) \varphi(x)
$$

Then

$$
\begin{equation*}
\varphi(x)-\varphi(y)=\sum_{j=1}^{n}\left(x_{j}-y_{j}\right) \varphi_{j}(x) \tag{3.12}
\end{equation*}
$$

with $\varphi_{j} \in \mathcal{S}$.
2. We claim next that if $\varphi(y)=0$, then $L \varphi(y)=0$. This follows from (3.12), since

$$
L \varphi(x)=\sum_{j=1}^{n}\left(x_{j}-y_{j}\right) L \varphi_{j}=0
$$

at $x=y$.

Therefore $L \varphi(x)=c(x) \varphi(x)$ for some function $c$. Taking $\varphi(x)=$ $e^{-|x|^{2}}$, we deduce that $c \in C^{\infty}$. Finally, since $L$ commutes with differentiation, we conclude that $c$ must be a constant.

THEOREM 3.4 (Integral identities). If $\varphi, \psi \in \mathcal{S}$, then

$$
\begin{equation*}
\int_{\mathbb{R}^{n}} \hat{\varphi} \psi d x=\int_{\mathbb{R}^{n}} \varphi \hat{\psi} d y \tag{3.13}
\end{equation*}
$$

and

$$
\begin{equation*}
\int_{\mathbb{R}^{n}} \varphi \bar{\psi} d x=\frac{1}{(2 \pi)^{n}} \int_{\mathbb{R}^{n}} \hat{\varphi} \overline{\hat{\psi}} d \xi \tag{3.14}
\end{equation*}
$$

In particular,

$$
\begin{equation*}
\|\varphi\|_{L^{2}}^{2}=\frac{1}{(2 \pi)^{n}}\|\hat{\varphi}\|_{L^{2}}^{2} . \tag{3.15}
\end{equation*}
$$

Proof. Note first that

$$
\begin{aligned}
\int_{\mathbb{R}^{n}} \hat{\varphi} \psi d x & =\int_{\mathbb{R}^{n}}\left(\int_{\mathbb{R}^{n}} e^{-i\langle x, y\rangle} \varphi(y) d y\right) \psi(x) d x \\
& =\int_{\mathbb{R}^{n}}\left(\int_{\mathbb{R}^{n}} e^{-i\langle y, x\rangle} \psi(x) d x\right) \varphi(y) d y=\int_{\mathbb{R}^{n}} \hat{\psi} \varphi d y
\end{aligned}
$$

Replace $\psi$ by $\overline{\hat{\psi}}$ in (3.13):

$$
\int_{\mathbb{R}^{n}} \hat{\varphi} \overline{\hat{\psi}} d \xi=\int_{\mathbb{R}^{n}} \varphi(\overline{\hat{\psi}})^{\wedge} d x
$$

But $\overline{\hat{\psi}}=\int_{\mathbb{R}^{n}} e^{i\langle x, \xi\rangle} \bar{\psi}(x) d x=(2 \pi)^{n} \mathcal{F}^{-1}(\bar{\psi})$ and so $(\overline{\hat{\psi}})^{\wedge}=(2 \pi)^{n} \bar{\psi}$.
We record next some elementary estimates that we will need later:

## LEMMA 3.5 (Useful estimates).

(i) We have the bounds

$$
\begin{equation*}
\|\hat{u}\|_{L^{\infty}} \leq\|u\|_{L^{1}} \tag{3.16}
\end{equation*}
$$

and

$$
\begin{equation*}
\|u\|_{L^{\infty}} \leq \frac{1}{(2 \pi)^{n}}\|\hat{u}\|_{L^{1}} \tag{3.17}
\end{equation*}
$$

(ii) There exists a constant $C$ such that

$$
\begin{equation*}
\|\hat{u}\|_{L^{1}} \leq C \sup _{|\alpha| \leq n+1}\left\|\partial^{\alpha} u\right\|_{L^{1}} \tag{3.18}
\end{equation*}
$$

Proof. Estimates (3.16) and (3.17) follow easily from (3.1) and (3.5).
Furthermore,

$$
\begin{aligned}
\|\hat{u}\|_{L^{1}} & =\int_{\mathbb{R}^{n}}|\hat{u}|\langle\xi\rangle^{n+1}\langle\xi\rangle^{-n-1} d \xi \leq C \sup \left(|\hat{u}|\langle\xi\rangle^{n+1}\right) \\
& \leq C \sup _{|\alpha| \leq n+1}\left|\xi^{\alpha} \hat{u}\right|=C \sup _{|\alpha| \leq n+1}\left|\left(\partial^{\alpha} u\right)^{\wedge}\right| \leq C \sup _{|\alpha| \leq n+1}\left\|\partial^{\alpha} u\right\|_{L^{1}}
\end{aligned}
$$

This proves (3.18).
We close this section with an application showing that we can sometimes use the Fourier transform to solve PDE with variable coefficients.

EXAMPLE: Solving a PDE. Consider the initial-value problem

$$
\left\{\begin{array}{cl}
\partial_{t} u=x \partial_{y} u+\partial_{x}^{2} u & \text { on } \mathbb{R}^{2} \times(0, \infty) \\
u=\delta_{\left(x_{0}, y_{0}\right)} & \\
\text { on } \mathbb{R}^{2} \times\{t=0\} .
\end{array}\right.
$$

Let $\hat{u}:=\mathcal{F} u$ denote the Fourier transform of $u$ in the variables $x, y$ (but not in $t$ ). Then

$$
\left(\partial_{t}+\eta \partial_{\xi}\right) \hat{u}=-\xi^{2} \hat{u} .
$$

This is a linear first-order PDE we can solve by the method of characteristics:

$$
\begin{aligned}
\hat{u}(t, \xi+t \eta, \eta) & =\hat{u}(0, \xi, \eta) e^{-\int_{0}^{t}(\xi+s \eta)^{2} d s} \\
& =\hat{u}(0, \xi, \eta) e^{-\xi^{2} t-\xi \eta t^{2}-\frac{\eta^{2} t^{3}}{3}} \\
& =\hat{u}(0, \xi, \eta) e^{-\frac{1}{2}\left\langle B_{t}(\xi, \eta),(\xi, \eta)\right\rangle}
\end{aligned}
$$

for

$$
B_{t}:=\left(\begin{array}{cc}
2 t & t^{2} \\
t^{2} & 2 t^{3} / 3
\end{array}\right) .
$$

Furthermore, $\hat{u}(0, \xi, \eta)=\hat{\delta}_{\left(x_{0}, y_{0}\right)}$. Taking $\mathcal{F}^{-1}$, we find

$$
\begin{aligned}
& u(t, x, y-t x)=\delta_{\left(x_{0}, y\right)} * \mathcal{F}^{-1}\left(e^{-\frac{1}{2}\left\langle B_{t}(\xi, \eta),(\xi, \eta)\right\rangle}\right) \\
& \quad=\frac{\sqrt{3}}{2 \pi t^{3}} \exp \left(-\frac{\left(x-x_{0}\right)^{2}}{t}+\frac{3\left(x-x_{0}\right)\left(y-y_{0}\right)}{t^{2}}-\frac{3\left(y-y_{0}\right)^{2}}{t^{3}}\right)
\end{aligned}
$$

and hence

$$
\begin{aligned}
& u(t, x, y) \\
& =\frac{\sqrt{3}}{2 \pi t^{3}} \exp \left(-\frac{\left(x-x_{0}\right)^{2}}{t}+\frac{3\left(x-x_{0}\right)\left(y+t x-y_{0}\right)}{t^{2}}-\frac{3\left(y+t x-y_{0}\right)^{2}}{t^{3}}\right) .
\end{aligned}
$$

### 3.2 FOURIER TRANSFORM ON $\mathcal{S}^{\prime}$

Next we extend the Fourier transform to $\mathcal{S}^{\prime}$, the dual space of $\mathcal{S}$. We will then be able to study the Fourier transforms of various important, but nonsmooth, expressions.

## DEFINITIONS.

(i) We write $\mathcal{S}^{\prime}=\mathcal{S}^{\prime}\left(\mathbb{R}^{n}\right)$ for the space of tempered distributions, which is the dual of $\mathcal{S}$. That is, $u \in \mathcal{S}^{\prime}$ provided $u: \mathcal{S} \rightarrow \mathbb{C}$ is linear and $\varphi_{j} \rightarrow \varphi$ in $\mathcal{S}$ implies $u\left(\varphi_{j}\right) \rightarrow u(\varphi)$.
(ii) We say

$$
u_{j} \rightarrow u \quad \text { in } \mathcal{S}^{\prime}
$$

if

$$
u_{j}(\varphi) \rightarrow u(\varphi) \quad \text { for all } \varphi \in \mathcal{S}
$$

DEFINITION. If $u \in \mathcal{S}^{\prime}$, we define

$$
D^{\alpha} u, x^{\alpha} u, \mathcal{F} u \in \mathcal{S}^{\prime}
$$

by the rules

$$
\begin{aligned}
D^{\alpha} u(\varphi) & :=(-1)^{|\alpha|} u\left(D^{\alpha} \varphi\right) \\
\left(x^{\alpha} u\right)(\varphi) & :=u\left(x^{\alpha} \varphi\right) \\
(\mathcal{F} u)(\varphi) & :=u(\mathcal{F} \varphi)
\end{aligned}
$$

for $\varphi \in \mathcal{S}$.
EXAMPLE 1: Dirac measure. It follows from the definitions that

$$
\hat{\delta}_{0}(\varphi)=\delta_{0}(\hat{\varphi})=\hat{\varphi}(0)=\int_{\mathbb{R}^{n}} \varphi d x .
$$

We interpret this calculation as saying that

$$
\hat{\delta}_{0} \equiv 1
$$

EXAMPLE 2: Exponential of an imaginary quadratic form. The signature of a real, symmetric, nonsingular matrix $Q$ is
$\operatorname{sgn} Q:=$ number of positive eigenvalues of $Q$

- number of negative eigenvalues of $Q$.

THEOREM 3.6 (Transform of an imaginary exponential). Let $Q$ be a real, symmetric, nonsingular $n \times n$ matrix. Then

$$
\begin{equation*}
\mathcal{F}\left(e^{\frac{i}{2}\langle Q x, x\rangle}\right)=\frac{(2 \pi)^{n / 2} e^{\frac{i \pi}{4} \operatorname{sgn}(Q)}}{|\operatorname{det} Q|^{1 / 2}} e^{-\frac{i}{2}\left\langle Q^{-1} \xi, \xi\right\rangle} \tag{3.20}
\end{equation*}
$$

Compare this carefully with the earlier formula (3.2). The extra phase shift term $e^{\frac{i \pi}{4} \operatorname{sgn} Q}$ in (3.20) arises from the complex exponential.

Proof. 1. Let $\epsilon>0, Q_{\epsilon}:=Q+\epsilon i I$. Then

$$
\begin{aligned}
\mathcal{F}\left(e^{\frac{i}{2}\left\langle Q_{\epsilon} x, x\right\rangle}\right) & =\int_{\mathbb{R}^{n}} e^{\frac{i}{2}\left\langle Q_{\epsilon} x, x\right\rangle-i\langle x, \xi\rangle} d x \\
& =\int_{\mathbb{R}^{n}} e^{\frac{i}{2}\left\langle Q_{\epsilon}\left(x-Q_{\epsilon}^{-1} \xi\right), x-Q_{\epsilon}^{-1} \xi\right\rangle} e^{-\frac{i}{2}\left\langle Q_{\epsilon}^{-1} \xi, \xi\right\rangle} d x \\
& =e^{-\frac{i}{2}\left\langle Q_{\epsilon}^{-1} \xi, \xi\right\rangle} \int_{\mathbb{R}^{n}} e^{\frac{i}{2}\left\langle Q_{\epsilon} y, y\right\rangle} d y .
\end{aligned}
$$

Now change variables, to write $Q$ in the form $\operatorname{diag}\left(\lambda_{1}, \ldots, \lambda_{n}\right)$, with $\lambda_{1}, \ldots, \lambda_{r}>0$ and $\lambda_{r+1}, \ldots, \lambda_{n}<0$. Then

$$
\int_{\mathbb{R}^{n}} e^{\frac{i}{2}\left\langle Q_{\epsilon} y, y\right\rangle} d y=\int_{\mathbb{R}^{n}} e^{\sum_{k=1}^{n} \frac{1}{2}\left(i \lambda_{k}-\epsilon\right) w_{k}^{2}} d w=\prod_{k=1}^{n} \int_{-\infty}^{\infty} e^{\frac{1}{2}\left(i \lambda_{k}-\epsilon\right) w^{2}} d w .
$$

2. If $1 \leq k \leq r$, then $\lambda_{k}>0$ and we set $z=\left(\epsilon-i \lambda_{k}\right)^{1 / 2} w$, and we take the branch of the square root so that $\operatorname{Im}\left(\epsilon-i \lambda_{k}\right)^{1 / 2}<0$. Then

$$
\int_{-\infty}^{\infty} e^{\frac{1}{2}\left(i \lambda_{k}-\epsilon\right) w^{2}} d w=\frac{1}{\left(\epsilon-i \lambda_{k}\right)^{1 / 2}} \int_{\Gamma_{k}} e^{-\frac{z^{2}}{2}} d z
$$

for the contour $\Gamma_{k}$ as drawn.
Since $e^{-\frac{z^{2}}{2}}=e^{\frac{y^{2}-x^{2}}{2}-i x y}$ and $x^{2}>y^{2}$ on $\Gamma_{k}$, we can deform $\Gamma_{k}$ into the real axis.


Hence

$$
\int_{\Gamma_{k}} e^{-\frac{z^{2}}{2}} d z=\int_{-\infty}^{\infty} e^{-\frac{x^{2}}{2}} d x=\sqrt{2 \pi}
$$

Thus

$$
\prod_{k=1}^{r} \int_{-\infty}^{\infty} e^{\frac{1}{2}\left(i \lambda_{k}-\epsilon\right) w^{2}} d w=(2 \pi)^{r / 2} \prod_{k=1}^{r} \frac{1}{\left(\epsilon-i \lambda_{k}\right)^{1 / 2}}
$$

Also for $1 \leq k \leq r$ :

$$
\lim _{\epsilon \rightarrow 0^{+}} \frac{1}{\left(\epsilon-i \lambda_{k}\right)^{1 / 2}}=\frac{1}{(-i)^{1 / 2} \lambda_{k}^{1 / 2}}=\frac{e^{\frac{i \pi}{4}}}{\lambda_{k}^{1 / 2}},
$$

since we take the branch of the square root with $(-i)^{1 / 2}=e^{-i \pi / 4}$.
3. Similarly for $r+1 \leq k \leq n$, we set $z=\left(\epsilon-i \lambda_{k}\right)^{1 / 2} w$, but now take the branch of square root with $\operatorname{Im}\left(\epsilon-i \lambda_{k}\right)^{1 / 2}>0$. Hence

$$
\prod_{k=r+1}^{n} \int_{-\infty}^{\infty} e^{\frac{1}{2}\left(i \lambda_{k}-\epsilon\right) w^{2}} d w=(2 \pi)^{\frac{n-r}{2}} \prod_{k=r+1}^{n} \frac{1}{\left(\epsilon-i \lambda_{k}\right)^{1 / 2}}
$$

and for $r+1 \leq k \leq n$

$$
\lim _{\epsilon \rightarrow 0^{+}} \frac{1}{\left(\epsilon-i \lambda_{k}\right)^{1 / 2}}=\frac{1}{\left(-i \lambda_{k}\right)^{1 / 2}}=\frac{e^{-\frac{i \pi}{4}}}{\left|\lambda_{k}\right|^{1 / 2}}
$$

since we take the branch of the square root with $i^{1 / 2}=e^{\frac{i \pi}{4}}$.
4. Combining the foregoing calculations gives us

$$
\begin{aligned}
\mathcal{F}\left(e^{\frac{i}{2}\langle Q x, x\rangle}\right) & =\lim _{\epsilon \rightarrow 0} \mathcal{F}\left(e^{\frac{i}{2}\left\langle Q_{\epsilon} x, x\right\rangle}\right) \\
& =e^{-\frac{i}{2}\left\langle Q^{-1} \xi, \xi\right\rangle} \frac{(2 \pi)^{n / 2} e^{\frac{i \pi}{4}(r-(n-r))}}{\left|\lambda_{1} \lambda_{2} \ldots \lambda_{n}\right|^{1 / 2}} \\
& =e^{-\frac{i}{2}\left\langle Q^{-1} \xi, \xi\right\rangle} \frac{(2 \pi)^{n / 2} e^{\frac{i \pi}{4} \operatorname{sgn} Q}}{|\operatorname{det} Q|^{1 / 2}} .
\end{aligned}
$$

### 3.3 SEMICLASSICAL FOURIER TRANSFORM

DEFINITION. The semiclassical Fourier transform for $h>0$ is

$$
\begin{equation*}
\hat{\varphi}(\xi)=\mathcal{F}_{h} \varphi(\xi):=\int_{\mathbb{R}^{n}} e^{-\frac{i}{h}\langle x, \xi\rangle} \varphi(x) d x \tag{3.21}
\end{equation*}
$$

and its inverse is

$$
\begin{equation*}
\mathcal{F}_{h}^{-1} \psi(x):=\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}\langle x, \xi\rangle} \psi(\xi) d \xi \tag{3.22}
\end{equation*}
$$

Consequently

$$
\begin{equation*}
\delta_{\{y=x\}}=\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}\langle x-y, \xi\rangle} d \xi \quad \text { in } \mathcal{S}^{\prime} \tag{3.23}
\end{equation*}
$$

This is a rescaled version of (3.9).
We record for future reference some formulas involving the parameter $h$ :

THEOREM 3.7 (Properties of $\mathcal{F}_{\boldsymbol{h}}$ ). We have

$$
\begin{gather*}
\left(h D_{\xi}\right)^{\alpha} \mathcal{F}_{h} \varphi=\mathcal{F}_{h}\left((-x)^{\alpha} \varphi\right)  \tag{3.24}\\
\mathcal{F}_{h}\left(\left(h D_{x}\right)^{\alpha} \varphi\right)=\xi^{\alpha} \mathcal{F}_{h} \varphi ; \tag{3.25}
\end{gather*}
$$

and

$$
\begin{equation*}
\|\varphi\|_{L^{2}}=\frac{1}{(2 \pi h)^{n / 2}}\left\|\mathcal{F}_{h} \varphi\right\|_{L^{2}} \tag{3.26}
\end{equation*}
$$

We present next a scaled version of the uncertainty principle, which in its various guises limits the extent to which we can simultaneously localize our calculations in both the $x$ and $\xi$ variables.

THEOREM 3.8 (Uncertainty principle). We have

$$
\begin{equation*}
\frac{h}{2}\|f\|_{L^{2}}\left\|\mathcal{F}_{h} f\right\|_{L^{2}} \leq\left\|x_{j} f\right\|_{L^{2}}\left\|\xi_{j} \mathcal{F}_{h} f\right\|_{L^{2}} \quad(j=1, \cdots, n) \tag{3.27}
\end{equation*}
$$

Proof. To see this, note first that

$$
\xi_{j} \mathcal{F}_{h} f(\xi)=\mathcal{F}_{h}\left(h D_{x_{j}} f\right)
$$

Also, if $A, B$ are self-adjoint opertors, then

$$
\operatorname{Im}\langle A f, B f\rangle=\frac{1}{2 i}\langle[B, A] f, f\rangle
$$

Let $A=h D, B=x$. Therefore

$$
[x, h D] f=\frac{h}{i}[\langle x, \partial f\rangle-\partial(x f)]=\operatorname{inh} f .
$$

Thus

$$
\begin{aligned}
\left\|x_{j} f\right\|_{L^{2}}\left\|\xi_{j} \mathcal{F}_{h} f\right\|_{L^{2}} & =\left\|x_{j} f\right\|_{L^{2}}\left\|\mathcal{F}_{h}\left(h D_{x_{j}} f\right)\right\|_{L^{2}} \\
& =(2 \pi h)^{n / 2}\left\|x_{j} f\right\|_{L^{2}}\left\|h D_{x_{j}} f\right\|_{L^{2}} \\
& \geq(2 \pi h)^{n / 2}\left|\left\langle h D_{x_{j}} f, x_{j} f\right\rangle\right| \\
& \geq(2 \pi h)^{n / 2}\left|\operatorname{Im}\left\langle h D_{x_{j}} f, x_{j} f\right\rangle\right| \\
& =\frac{(2 \pi h)^{n / 2}}{2}\left|\left\langle\left[x_{j}, h D_{x_{j}}\right] f, f\right\rangle\right| \\
& =\frac{(2 \pi h)^{n / 2}}{2} h\|f\|_{L^{2}}^{2} \\
& =\frac{h}{2}\|f\|_{L^{2}}\left\|\mathcal{F}_{h} f\right\|_{L^{2}} .
\end{aligned}
$$

### 3.4 STATIONARY PHASE IN ONE DIMENSION

Understanding the right hand side of (3.21) in the limit $h \rightarrow 0$ requires our studying integral expressions with rapidly oscillating integrands. We begin with one dimensional problems.
DEFINITION. Given functions $a \in C_{\mathrm{c}}^{\infty}(\mathbb{R}), \varphi \in C^{\infty}(\mathbb{R})$, we define for $h>0$ the oscillatory integral

$$
I_{h}=I_{h}(a, \varphi):=\int_{-\infty}^{\infty} e^{\frac{i \varphi}{\hbar}} a d x .
$$

LEMMA 3.9 (Rapid decay). If $\varphi^{\prime} \neq 0$ on $K:=\operatorname{spt}(a)$, then

$$
\begin{equation*}
I_{h}=O\left(h^{\infty}\right) \quad \text { as } h \rightarrow 0 . \tag{3.28}
\end{equation*}
$$

NOTATION. As explained in Appendix A, the identity (3.28) means that for each positive integer $N$, there exists a constant $C_{N}$ such that

$$
\left|I_{h}\right| \leq C_{N} h^{N} \quad \text { for all } 0<h \leq 1
$$

Proof. We will integrate by parts $N$ times. For this, observe that the operator

$$
L:=\frac{h}{i} \frac{1}{\varphi^{\prime}} \partial_{x}
$$

is defined on $K$, since $\varphi^{\prime} \neq 0$ there. Notice also that

$$
L\left(e^{\frac{i \varphi}{h}}\right)=e^{\frac{i \varphi}{h}}
$$

Hence $L^{N}\left(e^{i \varphi / h}\right)=e^{i \varphi / h}$, for $N=1,2, \ldots$ Consequently

$$
\left|I_{h}\right|=\left|\int_{-\infty}^{\infty} L^{N}\left(e^{\frac{i \varphi}{h}}\right) a d x\right|=\left|\int_{-\infty}^{\infty} e^{i \varphi / h}\left(L^{*}\right)^{N} a d x\right|,
$$

$L^{*}$ denoting the adjoint of $L$. Since $a$ is smooth, $L^{*} a=-\frac{h}{i} \partial_{x}\left(\frac{a}{\varphi^{\prime}}\right)$ is of order $h$. We deduce that $\left|I_{h}\right| \leq C_{N} h^{N}$.

Suppose next that $\varphi^{\prime}$ vanishes at some point within $K:=\operatorname{spt}(a)$, in which case the oscillatory integral is no longer of order $h^{\infty}$. We instead want to expand $I_{h}$ in an asymptotic expansion in powers of $h$ :

THEOREM 3.10 (Stationary phase). Let $a \in C_{c}^{\infty}(\mathbb{R})$. Suppose that $x_{0} \in K=\operatorname{spt}(a)$ and

$$
\varphi^{\prime}\left(x_{0}\right)=0, \varphi^{\prime \prime}\left(x_{0}\right) \neq 0
$$

Assume further that $\varphi^{\prime}(x) \neq 0$ on $K-\left\{x_{0}\right\}$.
(i) There exist for each $k=0,1, \ldots$ differential operators $A_{2 k}(x, D)$, of order less than or equal to $2 k$, such that for all $N$

$$
\begin{align*}
&\left|I_{h}-\left(\sum_{k=0}^{N-1} A_{2 k}(x, D) a\left(x_{0}\right) h^{k+\frac{1}{2}}\right) e^{\frac{i}{h} \varphi\left(x_{0}\right)}\right|  \tag{3.29}\\
& \leq C_{N} h^{N+\frac{1}{2}} \sum_{0 \leq m \leq 2 N+2} \sup _{\mathbb{R}}\left|a^{(m)}\right| .
\end{align*}
$$

(i) In particular,

$$
\begin{equation*}
A_{0}=(2 \pi)^{1 / 2}\left|\varphi^{\prime \prime}\left(x_{0}\right)\right|^{-1 / 2} e^{\frac{i \pi}{4} \operatorname{sgn} \varphi^{\prime \prime}\left(x_{0}\right)} ; \tag{3.30}
\end{equation*}
$$

and consequently

$$
\begin{equation*}
I_{h}=(2 \pi h)^{1 / 2}\left|\varphi^{\prime \prime}\left(x_{0}\right)\right|^{-1 / 2} e^{\frac{i \pi}{4} \operatorname{sgn} \varphi^{\prime \prime}\left(x_{0}\right)} e^{\frac{i \varphi\left(x_{0}\right)}{h}} a\left(x_{0}\right)+O\left(h^{3 / 2}\right) \tag{3.31}
\end{equation*}
$$

as $h \rightarrow 0$.
NOTATION. We will sometimes write (3.29) in the less precise form

$$
\begin{equation*}
I_{h} \sim e^{\frac{i}{h} \varphi\left(x_{0}\right)} \sum_{k=0}^{\infty} A_{2 k}(x, D) a\left(x_{0}\right) h^{k+\frac{1}{2}} . \tag{3.32}
\end{equation*}
$$

We present two proofs of this important theorem. The second proof is more complicated, but provides us with explicit expressions for the terms of the expansion (3.32): see (3.35).

First proof of Theorem 3.10. 1. We may without loss assume $x_{0}=0$, $\varphi(0)=0$. Then $\varphi(x)=\frac{1}{2} \psi(x) x^{2}$, for

$$
\psi(x):=2 \int_{0}^{1}(1-t) \varphi^{\prime \prime}(t x) d t
$$

Notice that $\psi(0)=\varphi^{\prime \prime}(0) \neq 0$. We change variables by writing

$$
y:=|\psi(x)|^{1 / 2} x
$$

for $x$ near 0 . Thus

$$
\partial_{y} x=\left|\varphi^{\prime \prime}(0)\right|^{-1 / 2} \quad \text { at } x=y=0 .
$$

Now select a smooth function $\chi: \mathbb{R} \rightarrow \mathbb{R}$ such that $0 \leq \chi \leq 1, \chi \equiv 1$ near 0 , and $\operatorname{sgn} \varphi^{\prime \prime}(x)=\operatorname{sgn} \varphi^{\prime \prime}(0) \neq 0$ on the support of $\chi$. Then Lemma 3.9 implies

$$
\begin{aligned}
I_{h} & =\int_{-\infty}^{\infty} e^{i \varphi(x) / h} \chi(x) a(x) d x+\int_{-\infty}^{\infty} e^{i \varphi(x) / h}(1-\chi(x)) a(x) d x \\
& =\int_{-\infty}^{\infty} e^{\frac{i \epsilon}{2 h} y^{2}} u(y) d y+O\left(h^{\infty}\right),
\end{aligned}
$$

for $\epsilon:=\operatorname{sgn} \varphi^{\prime \prime}(0)= \pm 1, u(y):=\chi(x(y)) a(x(y))\left|\operatorname{det} \partial_{y} x\right|$.
2. The Fourier transform formula (3.20) tells us that

$$
\mathcal{F}\left(e^{-\frac{i \epsilon y^{2}}{2 h}}\right)=(2 \pi h)^{1 / 2} e^{-\frac{i \pi \epsilon}{4}} e^{\frac{i \epsilon h \xi^{2}}{2}} .
$$

Applying (3.14), we see that consequently

$$
I_{h}=\left(\frac{h}{2 \pi}\right)^{1 / 2} e^{\frac{i \pi \epsilon}{4}} \int_{-\infty}^{\infty} e^{-\frac{i \epsilon h \xi^{2}}{2}} \hat{u}(\xi) d \xi+O\left(h^{\infty}\right)
$$

The advantage is that the small parameter $h$, and not $h^{-1}$, occurs in the exponential.
3. Next, write

$$
J(h, u):=\int_{-\infty}^{\infty} e^{-\frac{i \epsilon h \xi^{2}}{2}} \hat{u}(\xi) d \xi, \quad J(0, u)=2 \pi u(0)
$$

Then

$$
\partial_{h} J(h, u)=\int_{-\infty}^{\infty} e^{-\frac{i \epsilon h \xi^{2}}{2}}\left(\frac{\epsilon \xi^{2}}{2 i} \hat{u}(\xi)\right) d \xi=J(h, P u)
$$

for $P:=(i \epsilon / 2) \partial^{2}$. Continuing, we discover

$$
\partial_{h}^{k} J(h, u)=J\left(h, P^{k} u\right)
$$

Therefore

$$
J(h, u)=\sum_{k=0}^{N-1} \frac{h^{k}}{k!} J\left(0, P^{k} u\right)+\frac{h^{N}}{N!} R_{N}(h, u),
$$

for the remainder term

$$
R_{N}(h, u):=N \int_{0}^{1}(1-t)^{N-1} J\left(t h, P^{N} u\right) d t
$$

Thus Lemma 3.5 implies

$$
\left|R_{N}\right| \leq C_{N}\left\|\widehat{P^{N} u}\right\|_{L^{1}} \leq C_{N} \sum_{0 \leq k \leq 2} \sup _{\mathbb{R}}\left|\partial^{k}\left(P^{N} u\right)\right|
$$

4. Since the definition of $J$ gives

$$
h^{k} J\left(0, P^{k} u\right)=h^{2} P^{k} u(0)=(h / 2 i)^{k} u^{(2 k)}(0)
$$

and since $u=\chi(x(y)) a(x(y)) \mid \operatorname{det} \partial_{y} x$, the expansion follows.
The second proof of stationary phase asymptotics will employ this
LEMMA 3.11 (More on rapid decay). Suppose that $a \in C_{\mathrm{c}}^{\infty}(\mathbb{R})$ and that $\varphi \in C^{\infty}(\mathbb{R})$. For each positive integer $m$, there exists a constant $C_{m}$ such that

$$
\begin{equation*}
\left|\int_{-\infty}^{\infty} e^{i \varphi / h} a d x\right| \leq C_{m} h^{m} \sum_{0 \leq k \leq m} \sup _{\mathbb{R}}\left(\left|a^{(k)} \| \varphi^{\prime}\right|^{k-2 m}\right) \tag{3.33}
\end{equation*}
$$

This inequality will be useful at points where $\varphi^{\prime}$ is small, provided $a^{(m)}$ is also small.

Proof. The proof is an induction on $m$, the case $m=0$ being obvious. Assume the assertion for $m-1$. Then

$$
\begin{aligned}
\int_{-\infty}^{\infty} e^{i \varphi / h} a d x & =\frac{h}{i} \int_{-\infty}^{\infty}\left(e^{i \varphi / h}\right)^{\prime} \frac{a}{\varphi^{\prime}} d x \\
& =-\frac{h}{i} \int_{-\infty}^{\infty} e^{i \varphi / h}\left(\frac{a}{\varphi^{\prime}}\right)^{\prime} d x=-\frac{h}{i} \int_{-\infty}^{\infty} e^{i \varphi / h} \tilde{a} d x
\end{aligned}
$$

for

$$
\tilde{a}:=\left(a / \varphi^{\prime}\right)^{\prime}
$$

Observe that

$$
\left|\tilde{a}^{(k)}\right|=\left|\left(a / \varphi^{\prime}\right)^{(k+1)}\right| \leq C \sum_{0 \leq j \leq k+1} a^{(j)}\left|\varphi^{\prime}\right|^{j-k-2}
$$

The induction hypothesis therefore implies

$$
\begin{aligned}
\left|\int_{-\infty}^{\infty} e^{i \varphi(x) / h} a d x\right| & \leq h\left|\int_{-\infty}^{\infty} e^{i \varphi(x) / h} \tilde{a} d x\right| \\
& \leq h C_{m-1} h^{m-1} \sum_{0 \leq k \leq m-1} \sup _{\mathbb{R}}\left(\left|\tilde{a}^{(k)} \| \varphi^{\prime}\right|^{k-2(m-1)}\right) \\
& \leq C_{m} h^{m} \sum_{0 \leq j \leq m} \sup _{\mathbb{R}}\left(\left|a^{(j)} \| \varphi^{\prime}\right|^{j-2 m}\right) .
\end{aligned}
$$

Second proof of Theorem 3.10. 1. As before, we may assume $x_{0}=0$, $\varphi(0)=\varphi^{\prime}(0)=0, \varphi^{\prime \prime}(0) \neq 0$. To find the expansion in $h$ of our integral

$$
I_{h}=\int_{-\infty}^{\infty} e^{i \varphi / h} a d x
$$

we write

$$
\varphi_{s}(x):=\varphi^{\prime \prime}(0) x^{2} / 2+s g(x)
$$

for $0 \leq s \leq 1$, where

$$
g(x):=\varphi(x)-\varphi^{\prime \prime}(0) x^{2} / 2
$$

Then $\varphi=\varphi_{1}$ and $g=O\left(x^{3}\right)$ as $x \rightarrow 0$. Furthermore,

$$
\varphi_{s}^{\prime}(x)=\varphi^{\prime \prime}(0) x+O\left(x^{2}\right)
$$

and therefore

$$
|x| \leq\left|\varphi^{\prime \prime}(0)\right|^{-1}\left|\varphi_{s}^{\prime}(x)+O\left(x^{2}\right)\right| \leq 2\left|\varphi^{\prime \prime}(0)\right|^{-1}\left|\varphi^{\prime}(x)\right|
$$

for sufficiently small $x$. Consequently, using a cutoff function $\chi$ as in the first proof, we may assume that

$$
\begin{equation*}
\frac{x}{\varphi_{s}^{\prime}(x)} \text { is bounded on } K=\operatorname{spt}(a) \tag{3.34}
\end{equation*}
$$

2.We also write

$$
I_{h}(s):=\int_{-\infty}^{\infty} e^{i \varphi_{s} / h} a d x .
$$

Let us calculate

$$
\frac{d^{2 m}}{d s^{2 m}} I_{h}(s)=(i / h)^{2 m} \int_{-\infty}^{\infty} e^{i \varphi_{s} / h} g^{2 m} a d x
$$

Lemma 3.11, with $3 m$ replacing $m$, implies

$$
\left|I_{h}^{(2 m)}(s)\right| \leq \frac{C}{h^{2 m}} h^{3 m} \sum_{0 \leq k \leq 3 m} \sup _{\mathbb{R}}\left(\left|\left(a g^{2 m}\right)^{(k)} \| \varphi_{s}^{\prime}\right|^{k-6 m}\right) .
$$

Now the amplitude $a g^{2 m}$ vanishes to order $6 m$ at $x=0$. Consequently, for each $0 \leq k \leq 3 m$ we recall (3.34) to estimate

$$
\left|\left(a g^{2 m}\right)^{(k)}\right|\left|\varphi_{s}^{\prime}\right|^{k-6 m} \leq C|x|^{6 m-k}|x|^{k-6 m} \leq C .
$$

Therefore

$$
\left|I_{h}^{(2 m)}(s)\right| \leq M h^{m} .
$$

It follows that

$$
\begin{aligned}
I_{h}=I_{h}(1) & =\sum_{l=0}^{2 m-1} I_{h}^{(l)}(0) / l!+\frac{1}{(2 m-1)!} \int_{0}^{1}(1-s)^{2 m-1} I_{h}^{(2 m)}(s) d s \\
& =\sum_{l=0}^{2 m-1} I_{h}^{(l)}(0) / l!+O\left(h^{m}\right)
\end{aligned}
$$

3. It remains to compute the expansions in $h$ of the terms

$$
I_{h}^{(l)}(0)=(i / h)^{l} \int_{-\infty}^{\infty} e^{i \varphi_{0} / h} g^{l} a d x
$$

for $l=0, \ldots, 2 m-1$. But this follows as in the first proof, since the phase $\varphi_{0}(x)=\varphi^{\prime \prime}(0) x^{2} / 2$ is purely quadratic. Up to constants, the terms in the expansion are

$$
h^{\frac{1}{2}+k-l}\left(g^{l} a\right)^{(2 k)}(0)
$$

for $l<2 m$ and $k=0,1, \cdots$.
This at first first looks discouraging because of $-l$ in the power of $h$.
Recall however that $g=O\left(x^{3}\right)$ near 0 ; so that $\left(g^{l} a\right)^{(2 k)}(0)=0$ unless $2 k \geq 3 l$. Also, if $k-l=j$, then

$$
3 j=3 k-3 l \geq k, \quad 2 j=2 k-2 l \geq l .
$$

This means that there are at most finitely many values of $k$ and $l$ in the expansion corresponding to the term $h^{\frac{1}{2}+j}=h^{\frac{1}{2}+k-l}$.

REMARK. This second proof is a one dimensional, variant of that in Hörmander [H1, Section 7.7]. It avoids the Morse Lemma (see Theorem 3.14 below), but at some considerable technical expense. But this proof in fact provides the explicit expansion

$$
\begin{align*}
\int_{\mathbb{R}} e^{i \varphi / h} a d x \sim & e^{\frac{i \pi}{4} \operatorname{sgn} \varphi^{\prime \prime}\left(x_{0}\right)}\left(\frac{2 \pi h}{\left|\varphi^{\prime \prime}(0)\right|}\right)^{\frac{1}{2}} \\
& \sum_{k=0}^{\infty} \sum_{l=0}^{\infty}\left(\frac{h}{2 i \varphi^{\prime \prime}(0)}\right)^{k} \frac{1}{l!} \frac{1}{k!} \frac{d^{2 k}}{d x^{2 k}}\left((i / h)^{l} g^{l} a\right)(0) . \tag{3.35}
\end{align*}
$$

### 3.5 STATIONARY PHASE IN HIGHER DIMENSIONS

We turn next to $n$-dimensional integrals.
DEFINITION. We call the expression

$$
I_{h}=I_{h}(a, \varphi)=\int_{\mathbb{R}^{n}} e^{i \varphi / h} a d x
$$

where $a \in C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{n}\right), \varphi \in C^{\infty}\left(\mathbb{R}^{n}\right)$ are real-valued.
3.5.1 Quadratic phase function. We begin with the case of a quadratic phase

$$
\varphi(x)=\frac{1}{2}\langle Q x, x\rangle,
$$

where $Q$ is a nonsingular, symmetric matrix.

THEOREM 3.12 (Quadratic phase asymptotics). For each postive integer $N$, we have the expansion

$$
\begin{equation*}
(2 \pi h)^{\frac{n}{2}} \frac{e^{\frac{i \pi}{4} \operatorname{sgn} Q}}{|\operatorname{det} Q|^{\frac{1}{2}}}\left(\sum_{k=0}^{N-1} \frac{h^{k}}{k!}\left(\frac{\left\langle Q^{-1} D, D\right\rangle}{2 i}\right)^{k} a(0)+O\left(h^{N}\right)\right) . \tag{3.36}
\end{equation*}
$$

Proof. 1. The Fourier transform formulas (3.20) and (3.14) imply

$$
I_{h}=\left(\frac{h}{2 \pi}\right)^{n / 2} \frac{e^{\frac{i \pi}{4} \operatorname{sgn} Q}}{|\operatorname{det} Q|^{\frac{1}{2}}} \int_{\mathbb{R}^{n}} e^{-\frac{i h}{2}\left\langle Q^{-1} \xi, \xi\right\rangle} \hat{a}(\xi) d \xi
$$

Write

$$
J(h, a):=\int_{\mathbb{R}^{n}} e^{-\frac{i h}{2}\left\langle Q^{-1} \xi, \xi\right\rangle} \hat{a}(\xi) d \xi
$$

then

$$
\partial_{h} J(h, a)=\int_{\mathbb{R}^{n}} e^{-\frac{i h}{2}\left\langle Q^{-1} \xi, \xi\right\rangle}\left(-\frac{i}{2}\left\langle Q^{-1} \xi, \xi\right\rangle \hat{a}(\xi)\right) d \xi=J(h, P a)
$$

for

$$
P:=-\frac{i}{2}\left\langle Q^{-1} D, D\right\rangle
$$

Therefore

$$
J(h, a)=\sum_{k=0}^{N-1} \frac{h^{k}}{k!} J\left(0, P^{k} a\right)+\frac{h^{N}}{N!} R_{N}(h, a)
$$

for the remainder term

$$
R_{N}(h, a):=N \int_{0}^{1}(1-t)^{N-1} J\left(t h, P^{N} a\right) d t
$$

2. Now (3.5) gives

$$
J\left(0, P^{k} a\right)=\int_{\mathbb{R}^{n}}\left(-\frac{i}{2}\left\langle Q^{-1} \xi, \xi\right\rangle\right)^{k} \hat{a}(\xi) d \xi=(2 \pi)^{n} P^{k} a(0)
$$

Furthermore, Lemma 3.5,(ii) implies

$$
\left|R_{N}\right| \leq C_{N}\left\|\widehat{P^{N}} a\right\|_{L^{1}} \leq C_{N} \sup _{|\alpha| \leq 2 N+n+1}\left|\partial^{\alpha} a\right|
$$

3.5.2 General phase function. Assume next that the phase $\varphi$ is a smooth function.

LEMMA 3.13 (Rapid decay again). If $\partial \varphi \neq 0$ on $K:=\operatorname{spt}(a)$, then

$$
I_{h}=O\left(h^{\infty}\right) .
$$

In particular, for each positive integer $N$

$$
\begin{equation*}
\left|I_{h}\right| \leq C h^{N} \sum_{|\alpha| \leq N} \sup _{\mathbb{R}^{n}}\left|\partial^{\alpha} a\right|, \tag{3.37}
\end{equation*}
$$

where $C$ depends upon only $K$ and $n$.

Proof. Define the operator

$$
L:=\frac{h}{i} \frac{1}{|\partial \varphi|^{2}}\langle\partial \varphi, \partial\rangle
$$

for $x \in K$, and observe that

$$
L\left(e^{i \varphi / h}\right)=e^{i \varphi / h}
$$

Hence $L^{N}\left(e^{i \varphi / h}\right)=e^{i \varphi / h}$, and consequently

$$
\left|I_{h}\right|=\left|\int_{\mathbb{R}^{n}} L^{N}\left(e^{i \varphi / h}\right) a d x\right|=\left|\int_{\mathbb{R}^{n}} e^{i \varphi / h}\left(L^{*}\right)^{N} a d x\right| \leq C h^{N} .
$$

DEFINITION. We say $\varphi: \mathbb{R}^{n} \rightarrow \mathbb{R}$ has a nondegenerate critical point at $x_{0}$ if

$$
\partial \varphi\left(x_{0}\right)=0, \operatorname{det} \partial^{2} \varphi\left(x_{0}\right) \neq 0 .
$$

We also write
$\operatorname{sgn} \partial^{2} \varphi\left(x_{0}\right):=$ number of postive eigenvalues of $\partial^{2} \varphi\left(x_{0}\right)$

- number of negative eigenvalues of $\partial^{2} \varphi\left(x_{0}\right)$.

Next we change variables locally to convert the phase function $\varphi$ into a quadratic:

THEOREM 3.14 (Morse Lemma). Let $\varphi: \mathbb{R}^{n} \rightarrow \mathbb{R}$ be smooth, with a nondegenerate critical point at $x_{0}$. Then there exist neighborhoods $U$ of 0 and $V$ of $x_{0}$ and a diffeomorphism

$$
\gamma: V \rightarrow U
$$

such that

$$
\begin{equation*}
\left(\varphi \circ \gamma^{-1}\right)(x)=\varphi\left(x_{0}\right)+\frac{1}{2}\left(x_{1}^{2}+\cdots+x_{r}^{2}-x_{r+1}^{2} \cdots-x_{n}^{2}\right), \tag{3.38}
\end{equation*}
$$

where $r$ is the number of positive eigenvalues of $\partial^{2} \varphi\left(x_{0}\right)$.

Proof. 1. As usual, we suppose $x_{0}=0, \varphi(0)=0$. After a linear change of variables, we have

$$
\varphi(x)=\frac{1}{2}\left(x_{1}^{2}+\cdots+x_{r}^{2}-x_{r+1}^{2} \cdots-x_{n}^{2}\right)+O\left(|x|^{3}\right)
$$

and so the problem is to design a further change of variables that removes the cubic and higher terms.
2. Now

$$
\varphi(x)=\int_{0}^{1}(1-t) \partial_{t}^{2} \varphi(t x) d t=\frac{1}{2}\langle x, Q(x) x\rangle,
$$

where

$$
Q(0)=\partial^{2} \varphi(0)=\left(\begin{array}{cc}
I_{r} & O \\
O & -I_{n-r}
\end{array}\right) .
$$

In this expression the upper indentity matrix is $r \times r$ and the lower identity matrix is $(n-r) \times(n-r)$. We want to find a smooth mapping $A$ from $\mathbb{R}^{n}$ to $\mathbb{M}^{n \times n}$ such that

$$
\begin{equation*}
\langle A(x) x, Q(0) A(x) x\rangle=\langle x, Q(x) x\rangle . \tag{3.39}
\end{equation*}
$$

Then

$$
\gamma(x)=A(x) x
$$

is the desired change of variable.
Formula (3.39) will hold provided

$$
\begin{equation*}
A^{T}(x) Q(0) A(x)=Q(x) \tag{3.40}
\end{equation*}
$$

Let $F: \mathbb{M}^{n \times n} \rightarrow \mathbb{S}^{n \times n}$ be defined by

$$
F(A)=A^{T} Q(0) A
$$

We want to find a right inverse $G: \mathbb{S}^{n \times n} \rightarrow \mathbb{M}^{n \times n}$, so that

$$
F \circ G=I \quad \text { near } Q(0)
$$

Then

$$
A(x):=G(Q(x))
$$

will solve (3.40).
3. We will apply a version of the Implicit Function Theorem (Theorem C.2). To do so, it suffices to find $A \in L\left(\mathbb{S}^{n \times n}, \mathbb{M}^{n \times n}\right)$ such that

$$
\partial F(I) A=I
$$

Now

$$
\partial F(I)(C)=C^{T} Q(0)+Q(0) C
$$

Define

$$
A(D):=\frac{1}{2} Q(0)^{-1} D
$$

for $D \in \mathbb{S}^{n \times n}$. Then

$$
\begin{aligned}
\partial F(I) A(D) & =\frac{1}{2} \partial F(I)\left(Q^{-1}(0) D\right) \\
& =\frac{1}{2}\left[\left(Q(0)^{-1} D\right)^{T} Q(0)+Q(0)\left(Q(0)^{-1} D\right)\right] \\
& =D
\end{aligned}
$$

Given now a general phase function $\varphi$, we apply the Morse Lemma to convert locally to a quadratic phase for which the asymptotics provided by Theorem 3.12 apply:

THEOREM 3.15 (Stationary phase asymptotics). Assume that $a \in C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{n}\right)$. Suppose $x_{0} \in K:=\operatorname{spt}(a)$ and

$$
\partial \varphi\left(x_{0}\right)=0, \operatorname{det} \partial^{2} \varphi\left(x_{0}\right) \neq 0
$$

Assume further that $\partial \varphi(x) \neq 0$ on $K-\left\{x_{0}\right\}$.
(i) Then there exist for $k=0,1, \ldots$ differential operators $A_{2 k}(x, D)$ of order less than or equal to $2 k$, such that for each $N$

$$
\begin{align*}
&\left|I_{h}-\left(\sum_{k=0}^{N-1} A_{2 k}(x, D) a\left(x_{0}\right) h^{k+\frac{n}{2}}\right) e^{\frac{i \varphi\left(x_{0}\right)}{h}}\right|  \tag{3.41}\\
& \leq C_{N} h^{N+\frac{n}{2}} \sum_{|\alpha| \leq 2 N+n+1} \sup _{\mathbb{R}^{n}}\left|\partial^{\alpha} a\right| .
\end{align*}
$$

(ii) In particular,

$$
\begin{equation*}
A_{0}=(2 \pi)^{n / 2}\left|\operatorname{det} \partial^{2} \varphi\left(x_{0}\right)\right|^{-1 / 2} e^{\frac{i \pi}{4} \operatorname{sgn} \partial^{2} \varphi\left(x_{0}\right)} \tag{3.42}
\end{equation*}
$$

and therefore

$$
\begin{align*}
& I_{h}= \\
& \qquad(2 \pi h)^{n / 2}\left|\operatorname{det} \partial^{2} \varphi\left(x_{0}\right)\right|^{-1 / 2} e^{\frac{i \pi}{4} \operatorname{sgn} \partial^{2} \varphi\left(x_{0}\right)} e^{\frac{i \varphi\left(x_{0}\right)}{h}} a\left(x_{0}\right)+O\left(h^{\frac{n+2}{2}}\right) \tag{3.43}
\end{align*}
$$

as $h \rightarrow 0$.

Proof. Without loss $x_{0}=0, \varphi\left(x_{0}\right)=\partial \varphi\left(x_{0}\right)=0$. Introducing a cutoff function $\chi$ and applying the Morse Lemma, Theorem 3.14, and then Lemma 3.13, we can write

$$
I_{h}=\int_{\mathbb{R}^{n}} e^{i \varphi(x) / h} a d x=\int_{\mathbb{R}^{n}} e^{\frac{i}{2 h}\langle Q x, x\rangle} u d x+O\left(h^{\infty}\right)
$$

where

$$
Q=\left(\begin{array}{cc}
I_{r} & O \\
O & -I_{n-r}
\end{array}\right)
$$

and

$$
u(x):=a\left(\kappa^{-1}(x)\right)\left|\operatorname{det} \partial \kappa^{-1}(x)\right|, \quad \operatorname{det} \partial \kappa^{-1}(0)=\operatorname{det} \partial \phi\left(x_{0}\right)^{-1}
$$

and

$$
Q=\left(\begin{array}{cc}
I_{r} & O \\
O & -I_{n-r}
\end{array}\right)
$$

Note that $\operatorname{sgn} Q=\operatorname{sgn} \partial^{2} \varphi\left(x_{0}\right)$ and $|\operatorname{det} Q|=1$. We invoke Theorem 3.12 to finish the proof.

### 3.6 IMPORTANT EXAMPLES.

In Chapter 4 we will be interested in asymptotics in $h$ of various expressions involving the Fourier transform. These involve the particular phase function

$$
\varphi(x, y)=\langle x, y\rangle
$$

on $\mathbb{R}^{n} \times \mathbb{R}^{n}$, corresponding to the Euclidean inner product. We will also encounter important applications with the phase

$$
\varphi(z, w)=\sigma(z, w)=\langle J z, w\rangle
$$

on $\mathbb{R}^{2 n} \times \mathbb{R}^{2 n}$, corresponding to the symplectic structure. We therefore record in this section the stationary phase expansions corresponding to these special cases.

THEOREM 3.16 (Important phase functions).
(i) Assume that $a \in C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{2 n}\right)$. Then for each postive integer $N$,

$$
\begin{align*}
& \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}\langle x, y\rangle} a(x, y) d x d y=  \tag{3.44}\\
& \quad(2 \pi h)^{n}\left(\sum_{k=0}^{N-1} \frac{h^{k}}{k!}\left(\frac{\left\langle D_{x}, D_{y}\right\rangle}{i}\right)^{k} a(0,0)+O\left(h^{N}\right)\right)
\end{align*}
$$

as $h \rightarrow 0$.
(ii) Assume that $a \in C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{4 n}\right)$. Then for each postive integer $N$,

$$
\begin{align*}
& \int_{\mathbb{R}^{2 n}} \int_{\mathbb{R}^{2 n}} e^{\frac{i}{h} \sigma(z, w)} a(z, w) d z d w=  \tag{3.45}\\
& \quad(2 \pi h)^{2 n}\left(\sum_{k=0}^{N-1} \frac{h^{k}}{k!}\left(\frac{\sigma\left(D_{x}, D_{\xi}, D_{y}, D_{\eta}\right)}{i}\right)^{k} a(0,0)+O\left(h^{N}\right)\right)
\end{align*}
$$

where $z=(x, \xi), w=(y, \eta)$, and

$$
\sigma\left(D_{x}, D_{\xi}, D_{y}, D_{\eta}\right):=\left\langle D_{\xi}, D_{y}\right\rangle-\left\langle D_{x}, D_{\eta}\right\rangle .
$$

Proof. 1. We write $(x, y)$ to denote a typical point of $\mathbb{R}^{2 n}$, and let

$$
Q:=\left(\begin{array}{cc}
O & I_{n} \\
I_{n} & O
\end{array}\right) .
$$

Then $Q$ is symmetric, $Q^{-1}=Q,|\operatorname{det} Q|=1, \operatorname{sgn}(Q)=0$ and $Q(x, y)=$ $(y, x)$. Consequently $\frac{1}{2}\langle Q(x, y),(x, y)\rangle=\langle x, y\rangle$.

Furthermore, since $D=\left(D_{x}, D_{y}\right)$,

$$
\frac{1}{2}\left\langle Q^{-1} D, D\right\rangle=\left\langle D_{x}, D_{y}\right\rangle .
$$

Hence Theorem 3.12 gives (3.44).
2. We write $(z, w)$ to denote a typical point of $\mathbb{R}^{4 n}$, where $z=(x, \xi)$, $w=(y, \eta)$. Set

$$
Q:=\left(\begin{array}{cc}
O & -J \\
J & O
\end{array}\right) .
$$

Then $Q$ is symmetric, $Q^{-1}=Q,|\operatorname{det} Q|=1, \operatorname{sgn}(Q)=0$ and $Q(z, w)=$ $(-J w, J z)$. Consequently $\frac{1}{2}\langle Q(z, w),(z, w)\rangle=\langle J z, w\rangle=\sigma(z, w)$.

We have $D=\left(D_{z}, D_{w}\right)=\left(D_{x}, D_{\xi}, D_{y}, D_{\eta}\right)$, and therefore

$$
\frac{1}{2}\left\langle Q^{-1} D, D\right\rangle=\sigma\left(D_{x}, D_{\xi}, D_{y}, D_{\eta}\right) .
$$

Theorem 3.12 now provides us with the expansion (3.45).

## 4. Quantization of symbols

4.1 Quantization formulas
4.2 Composition, asymptotic expansions
4.3 General symbol classes
4.4 Operators on $L^{2}$
4.5 Inverses
4.6 Gårding inequalities

The Fourier transform and its inverse allow us to move at will between the position $x$ and momentum $\xi$ variables, but what we really want is to deal with both sets of variables simultaneously. This chapter therefore introduces the quantization of symbols, that is, of appropriate functions of both $x$ and $\xi$. The resulting operators applied to functions entail information in the full $(x, \xi)$ phase space, and particular choices of the symbol will later prove very useful, allowing us for example to "localize" in phase space.

The plan is to introduce quantization and then to work out the resulting symbol calculus, meaning the systematic rules for manipulating symbols and their associated operators.

### 4.1 QUANTIZATION FORMULAS

NOTATION. For this section we take $h>0$ and $a \in \mathcal{S}\left(\mathbb{R}^{2 n}\right), a=$ $a(x, \xi)$. We hereafter call $a$ a symbol.

To quantize this symbol means to associate with it an $h$-dependent linear operator acting on functions $u=u(x)$. There are several standard ways to do so:

## DEFINITIONS.

(i) We define the Weyl quantization to be the operator $a^{\mathrm{w}}(x, h D)$ acting on $u \in \mathcal{S}\left(\mathbb{R}^{n}\right)$ by the formula

$$
\begin{equation*}
a^{\mathrm{w}}(x, h D) u(x):=\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}\langle x-y, \xi\rangle} a\left(\frac{x+y}{2}, \xi\right) u(y) d y d \xi \tag{4.1}
\end{equation*}
$$

(ii) We define also the standard quantization

$$
\begin{equation*}
a(x, h D) u(x):=\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}\langle x-y, \xi\rangle} a(x, \xi) u(y) d y d \xi \tag{4.2}
\end{equation*}
$$

for $u \in \mathcal{S}$.
(iii) More generally, for $u \in \mathcal{S}$ and $0 \leq t \leq 1$, we set

$$
\begin{align*}
& \mathrm{Op}_{t}(a) u(x):=  \tag{4.3}\\
& \frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}\langle x-y, \xi\rangle} a(t x+(1-t) y, \xi) u(y) d y d \xi .
\end{align*}
$$

Hence

$$
\mathrm{Op}_{\frac{1}{2}}(a)=a^{\mathrm{w}}(x, h D), \quad \mathrm{Op}_{1}(a)=a(x, h D) .
$$

REMARKS. (i) Observe that

$$
a(x, h D) u=\mathcal{F}_{h}^{-1}\left(a \mathcal{F}_{h} u\right) .
$$

This simple expression makes most of the subsequent calculations much easier for the standard quantization, as opposed to the Weyl quantization. However the later has many better properties and will be our principal concern.
(ii) We will only rarely be directly interested in the operators $\mathrm{Op}_{t}$ for $t \neq \frac{1}{2}, 1$; but they will prove useful for interpolating between the Weyl and standard quantizations.

## EXAMPLES.

(i) If $a(x, \xi)=\xi^{\alpha}$, then

$$
\mathrm{Op}_{t}(a) u=(h D)^{\alpha} u \quad(0 \leq t \leq 1) .
$$

(ii) If $a(x, \xi)=V(x)$, then

$$
\mathrm{Op}_{t}(a) u=V(x) u \quad(0 \leq t \leq 1) .
$$

(iii) If $a(x, \xi)=\langle x, \xi\rangle$, then

$$
\mathrm{Op}_{t}(a) u=(1-t)\langle h D, x u\rangle+t\langle x, h D u\rangle \quad(0 \leq t \leq 1)
$$

(iv) If $a(x, \xi)=\sum_{|\alpha| \leq N} a_{\alpha}(x) \xi^{\alpha}$ and $t=1$, then

$$
a(x, h D)=\sum_{|\alpha| \leq N} a_{\alpha}(x)(h D)^{\alpha} u
$$

These formulas follow straightforwardly from the definitions.

THEOREM 4.1 (Schwartz class symbols). Assume $a \in \mathcal{S}$.
(i) Then for each $0 \leq t \leq 1, \mathrm{Op}_{t}(a)$ can be defined as an operator mapping $\mathcal{S}^{\prime}$ to $\mathcal{S}$; and furthermore

$$
\mathrm{Op}_{t}(a): \mathcal{S}^{\prime} \rightarrow \mathcal{S}
$$

is continuous.
(ii) We have

$$
\begin{equation*}
\mathrm{Op}_{t}(a)^{*}=\mathrm{Op}_{1-t}(\bar{a}) \quad(0 \leq t \leq 1) ; \tag{4.4}
\end{equation*}
$$

and in particular the Weyl quantization of a real symbol is self-adjoint:

$$
\begin{equation*}
a^{\mathrm{w}}(x, h D)^{*}=a^{\mathrm{w}}(x, h D) \quad \text { if } a \text { is real. } \tag{4.5}
\end{equation*}
$$

Proof. (i) We have

$$
\mathrm{Op}_{t}(a) u(x)=\int_{\mathbb{R}^{n}} K_{t}(x, y) u(y) d y
$$

for the kernel

$$
\begin{aligned}
K_{t}(x, y) & :=\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}\langle x-y, \xi\rangle} a(t x+(1-t) y, \xi) d \xi \\
& =\mathcal{F}_{h}^{-1}(a(t x+(1-t) y, \cdot))(x-y)
\end{aligned}
$$

Thus $K_{t} \in \mathcal{S}$, and so

$$
\mathrm{Op}_{t}(a) u(x)=u\left(K_{t}(x, \cdot)\right)
$$

maps $\mathcal{S}^{\prime}$ continuously into $\mathcal{S}$.
(ii) The kernel of $\mathrm{Op}_{t}(a)^{*}$ is $K_{t}^{*}(x, y):=\bar{K}_{t}(y, x)=\bar{K}_{1-t}(x, y)$, which is the kernel of $\mathrm{Op}_{1-t}(\bar{a})$.

We next observe that the formulas (4.1)-(4.3) make sense if $a$ is merely a distribution:

THEOREM 4.2 (Distributional symbols). If $a \in \mathcal{S}^{\prime}$, then $\mathrm{Op}_{t}(a)$ can be defined as an operator mapping $\mathcal{S}$ to $\mathcal{S}^{\prime}$; and furthermore

$$
\mathrm{Op}_{t}(a): \mathcal{S} \rightarrow \mathcal{S}^{\prime} \quad(0 \leq t \leq 1)
$$

is continuous.

Proof. The formula for the distributional kernel $K_{t}$ of $\mathrm{Op}_{t}(a)$ shows that $K_{t} \in \mathcal{S}^{\prime}\left(\mathbb{R}^{n} \times \mathbb{R}^{n}\right)$. Hence $\mathrm{Op}_{t}(a)$ is well defined as an operator from $\mathcal{S}$ to $\mathcal{S}^{\prime}$ : if $u, v \in \mathcal{S}$ then

$$
\left(\mathrm{Op}_{t}(a) u\right)(v):=K_{t}(u \otimes v)
$$

### 4.2 COMPOSITION, ASYMPTOTIC EXPANSIONS

We begin now a careful study of the properties of the quantized operators defined above, especially the Weyl quantization. Our particular
goal in this section is showing that if $a$ and $b$ are symbols, then there exists a symbol $c=a \# b$ such that

$$
a^{\mathrm{w}}(x, h D) b^{\mathrm{w}}(x, h D)=c^{\mathrm{w}}(x, h D) .
$$

4.2.1 Linear symbols. We begin with linear symbols of the form

$$
\begin{equation*}
l(x, \xi):=\left\langle x^{*}, x\right\rangle+\left\langle\xi^{*}, \xi\right\rangle \tag{4.6}
\end{equation*}
$$

for $\left(x^{*}, \xi^{*}\right) \in \mathbb{R}^{2 n}$. To simplify calculations later on, we will often identify the linear symbol $l$ with the point $\left(x^{*}, \xi^{*}\right)$.

LEMMA 4.3 (Quantizing linear symbols). Let $l$ be given by (4.6). Then

$$
\mathrm{Op}_{t}(l)=\left\langle x^{*}, x\right\rangle+\left\langle\xi^{*}, h D\right\rangle \quad(0 \leq t \leq 1)
$$

NOTATION. In view of this result, we hereafter write

$$
\begin{equation*}
l(x, h D)=l^{\mathrm{w}}(x, h D)=\left\langle x^{*}, x\right\rangle+\left\langle\xi^{*}, h D\right\rangle \tag{4.7}
\end{equation*}
$$

Proof. Let $u \in \mathcal{S}$ and compute the derivative

$$
\begin{aligned}
\frac{d}{d t} \mathrm{Op}_{t}(l) u & =\frac{1}{(2 \pi h)^{n}} \frac{d}{d t} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}\langle x-y, \xi\rangle}\left(\left\langle x^{*}, t x+(1-t) y\right\rangle\right. \\
& \left.=\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}\langle x-y, \xi\rangle}\left\langle\xi^{*}, \xi\right\rangle\right) u(y) d y d \xi \\
& \left.=\frac{h}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}}\left\langle x^{*}, D_{\xi} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}\langle x-y, \xi\rangle} u(y) d y\right\rangle d \xi\right) d y d \xi \\
& =\frac{h}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}}\left\langle x^{*}, D_{\xi}\left(e^{\frac{i}{h}\langle x, \xi\rangle} \hat{u}(\xi)\right)\right\rangle d \xi
\end{aligned}
$$

Since $\hat{u}(\xi) \rightarrow 0$ rapidly as $|\xi| \rightarrow \infty$, the last expression vanishes. Therefore $\mathrm{Op}_{t}(l)$ does not in fact depend upon $t$; and consequently for all $0 \leq t \leq 1, \mathrm{Op}_{t}(l) u=\mathrm{Op}_{1}(l) u=\left\langle x^{*}, x\right\rangle u+\left\langle\xi^{*}, h D\right\rangle u$.

Next we compute the Weyl quantization of $e^{\frac{i}{h} l}$.
THEOREM 4.4 (Quantizing exponentials of linear symbols).
(i) For each linear symbol l we have the identity

$$
\begin{equation*}
\left(e^{\frac{i}{h} l}\right)^{\mathrm{w}}(x, h D)=e^{\frac{i}{h} l(x, h D)}, \tag{4.8}
\end{equation*}
$$

where

$$
\begin{equation*}
e^{\frac{i}{h} l(x, h D)} u(x):=e^{\frac{i}{h}\left\langle x^{*}, x\right\rangle+\frac{i}{2 h}\left\langle x^{*}, \xi^{*}\right\rangle} u\left(x+\xi^{*}\right) . \tag{4.9}
\end{equation*}
$$

(ii) If $l, m \in \mathbb{R}^{2 n}$, then

$$
\begin{equation*}
e^{\frac{i}{h} l(x, h D)} e^{\frac{i}{h} m(x, h D)}=e^{\frac{i}{2 h} \sigma(l, m)} e^{\frac{i}{h}(l+m)(x, h D)} . \tag{4.10}
\end{equation*}
$$

Proof. 1. Consider for $u \in \mathcal{S}$ the PDE

$$
\left\{\begin{aligned}
i h \partial_{t} v+l(x, h D) v & =0 & & (t \in \mathbb{R}) \\
v(0) & =u & & (t=0) .
\end{aligned}\right.
$$

Its unique solution is denoted

$$
v(x, t)=e^{\frac{i t}{h} l(x, h D)} u,
$$

this formula defining the operators $e^{\frac{i t}{h} l(x, h D)}$ for $t \in \mathbb{R}$. But we can check by a direct calculation using (4.7) that

$$
v(x, t)=e^{\frac{i t}{h}\left\langle x^{*}, x\right\rangle+\frac{i t^{2}}{2 h}\left\langle x^{*}, \xi^{*}\right\rangle} u\left(x+t \xi^{*}\right) ;
$$

and therefore (4.9) holds.
2. Furthermore,

$$
\begin{aligned}
\left(e^{\frac{i l}{h}}\right)^{\mathrm{w}} u= & \frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}\langle x-y, \xi\rangle} e^{\frac{i}{h}\left(\left\langle\xi^{*}, \xi\right\rangle+\left\langle x^{*}, \frac{x+y}{2}\right\rangle\right)} u(y) d y d \xi \\
& =\frac{e^{\frac{i}{2 h}\left\langle x^{*}, x\right\rangle}}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}\left\langle x-y+\xi^{*}, \xi\right\rangle}\left(e^{\frac{i}{2 h}\left\langle x^{*}, y\right\rangle} u(y)\right) d y d \xi \\
& =\frac{e^{\frac{i}{2 h}\left\langle x^{*}, x\right\rangle}}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}\langle x-y, \xi\rangle}\left(e^{\frac{i}{2 h}\left\langle x^{*}, y+\xi^{*}\right\rangle} u\left(y+\xi^{*}\right)\right) d y d \xi \\
& =e^{\frac{i}{h}\left\langle x^{*}, x\right\rangle+\frac{i}{2 h}\left\langle x^{*}, \xi^{*}\right\rangle} u\left(x+\xi^{*}\right),
\end{aligned}
$$

since

$$
\delta_{\{y=x\}}=\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}\langle x-y, \xi\rangle} d \xi \quad \text { in } \mathcal{S}^{\prime},
$$

according to (3.23). This proves (4.8).
3. Suppose $l(x, \xi)=\left\langle x_{1}^{*}, x\right\rangle+\left\langle\xi_{1}^{*}, \xi\right\rangle$ and $m(x, \xi)=\left\langle x_{2}^{*}, x\right\rangle+\left\langle\xi_{2}^{*}, \xi\right\rangle$. According to (4.9),

$$
e^{\frac{i}{h} m(x, h D)} u(x)=e^{\frac{i}{h}\left\langle x_{2}^{*}, x\right\rangle+\frac{i}{2 h}\left\langle x_{2}^{*}, \xi_{2}^{*}\right\rangle} u\left(x+\xi_{2}^{*}\right) ;
$$

and consequently

$$
\begin{aligned}
& e^{\frac{i}{h} l(x, h D)} e^{\frac{i}{h} m(x, h D)} u(x)= \\
& \quad e^{\frac{i}{h}\left\langle x_{1}^{*}, x\right\rangle+\frac{i}{2 h}\left\langle x_{1}^{*}, \xi_{1}^{*}\right\rangle} e^{\frac{i}{h}\left\langle x_{2}^{*}, x+\xi_{1}^{*}\right\rangle+\frac{i}{2 h}\left\langle x_{2}^{*}, \xi_{2}^{*}\right\rangle} u\left(x+\xi_{1}^{*}+\xi_{2}^{*}\right) .
\end{aligned}
$$

Furthermore, (4.9) implies also that

$$
e^{\frac{i}{h}(l+m)(x, h D)} u(x)=e^{\frac{i}{h}\left\langle x_{1}^{*}+x_{2}^{*}, x\right\rangle+\frac{i}{2 h}\left\langle x_{1}^{*}+x_{2}^{*}, \xi_{1}^{*}+\xi_{2}^{*}\right\rangle} u\left(x+\xi_{1}^{*}+\xi_{2}^{*}\right) .
$$

Using the formula above, we therefore compute

$$
e^{\frac{i}{h}(l+m)(x, h D)} u(x)=e^{\frac{i}{2 h}\left(\left\langle x_{1}^{*}, \xi_{2}^{*}\right\rangle-\left\langle x_{2}^{*}, \xi_{1}^{*}\right\rangle\right)} e^{\frac{i}{h} l(x, h D)} e^{\frac{i}{h} m(x, h D)} u(x) .
$$

This proves (4.10), since $\sigma(l, m)=\left\langle\xi_{1}^{*}, x_{2}^{*}\right\rangle-\left\langle x_{1}^{*}, \xi_{2}^{*}\right\rangle$.
4.2.2 Exponentials of quadratics. We record in the section some useful integral representation formulas for the quantization of certain quadratic exponentials. We will later apply the stationary phase expansions from Theorems 3.12 and 3.16 to these expressions.

## THEOREM 4.5 (Quantizing quadratic exponentials).

(i) Let $Q$ denote a nonsingular, symmetric, $n \times n$ matrix. Then if $u \in \mathcal{S}\left(\mathbb{R}^{n}\right)$,

$$
\begin{align*}
& e^{\frac{i \hbar}{2}\langle Q D, D\rangle} u(x)=  \tag{4.11}\\
& \frac{|\operatorname{det} Q|^{-\frac{1}{2}}}{(2 \pi h)^{\frac{n}{2}}} e^{\frac{i \pi}{4} \operatorname{sgn} Q} \int_{\mathbb{R}^{n}} e^{-\frac{i}{2 h}\left\langle Q^{-1} y, y\right\rangle} u(x+y) d y .
\end{align*}
$$

(ii) In particular, if $u \in \mathcal{S}\left(\mathbb{R}^{2 n}\right)$, $u=u(x, y)$, then

$$
\begin{align*}
& e^{i h\left\langle D_{x}, D_{y}\right\rangle} u(x, y)=  \tag{4.12}\\
& \frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{-\frac{i}{h}\left\langle x_{1}, y_{1}\right\rangle} u\left(x+x_{1}, y+y_{1}\right) d x_{1} d y_{1} .
\end{align*}
$$

(iii) Suppose that $u \in \mathcal{S}\left(\mathbb{R}^{4 n}\right)$, $u=u(z, w)$. Then

$$
\begin{align*}
& e^{i h \sigma\left(D_{z}, D_{z}\right)} u(z, w)=  \tag{4.13}\\
& \quad \frac{1}{(2 \pi h)^{2 n}} \int_{\mathbb{R}^{2 n}} \int_{\mathbb{R}^{2 n}} e^{-\frac{i}{h} \sigma\left(z_{1}, w_{1}\right)} u\left(z+z_{1}, w+w_{1}\right) d z_{1} d w_{1} .
\end{align*}
$$

Proof. 1. Observe first that

$$
\begin{aligned}
\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}\langle w, \xi\rangle} e^{\frac{i}{h}\langle Q \xi, \xi\rangle} d \xi & =\mathcal{F}_{h}^{-1}\left(e^{\frac{i}{h}\langle Q \xi, \xi\rangle}\right)(w) \\
& =\frac{|\operatorname{det} Q|^{-\frac{1}{2}}}{(2 \pi h)^{\frac{n}{2}}} e^{\frac{i \pi}{4} \operatorname{sgn} Q} e^{-\frac{i}{2 h}\left\langle Q^{-1} w, w\right\rangle}
\end{aligned}
$$

Therefore

$$
\begin{aligned}
e^{\frac{i \hbar}{2}\langle Q D, D\rangle} u(x) & =e^{\frac{i}{h}\langle Q h D, h D\rangle} u(x) \\
& =\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}\langle x-y, \xi\rangle} e^{\frac{i}{h}\langle Q \xi, \xi\rangle} u(y) d y d \xi \\
& =\frac{|\operatorname{det} Q|^{-\frac{1}{2}}}{(2 \pi h)^{\frac{n}{2}}} e^{\frac{i \pi}{4} \operatorname{sgn} Q} \int_{\mathbb{R}^{n}} e^{-\frac{i}{2 h}\left\langle Q^{-1}(x-y), x-y\right\rangle} u(y) d y \\
& =\frac{|\operatorname{det} Q|^{-\frac{1}{2}}}{(2 \pi h)^{\frac{n}{2}}} e^{\frac{i \pi}{4} \operatorname{sgn} Q} \int_{\mathbb{R}^{n}} e^{-\frac{i}{2 h}\left\langle Q^{-1} y, y\right\rangle} u(x+y) d y .
\end{aligned}
$$

2. Assertion (4.12) is a special case of (4.11), had by replacing $n$ by $2 n$ and taking

$$
Q:=\left(\begin{array}{cc}
O & I_{n} \\
I_{n} & O
\end{array}\right)
$$

See the proof of Theorem 3.16,(i). It is a useful exercise to give a direct derivation.
3. Similarly, assertion (4.13) is a special case of (4.11) obtained by replacing $n$ by $4 n$ and taking

$$
Q:=\left(\begin{array}{cc}
O & -J \\
J & O
\end{array}\right) .
$$

See the proof of Theorem 3.16,(ii).

### 4.2.3 Composing symbols.

Next we establish the fundamental formula $a^{\mathrm{w}} b^{\mathrm{w}}=(a \# b)^{\mathrm{w}}$, along with a recipe for computing the new symbol $a \# b$ :

## THEOREM 4.6 (Composition for Weyl quantization).

(i) Suppose that $a, b \in \mathcal{S}$. Then

$$
\begin{equation*}
a^{\mathrm{w}}(x, h D) b^{\mathrm{w}}(x, h D)=(a \# b)^{\mathrm{w}}(x, h D) \tag{4.14}
\end{equation*}
$$

for the symbol

$$
\begin{equation*}
a \# b(x, \xi):=\left.e^{\frac{i h}{2} \sigma\left(D_{x}, D_{\xi}, D_{y}, D_{\eta}\right)}(a(x, \xi) b(y, \eta))\right|_{\substack{y=x \\ \eta=\xi}} \tag{4.15}
\end{equation*}
$$

(ii) We have the integral representation formula

$$
\begin{aligned}
& a \# b(x, \xi)= \\
& \quad \frac{1}{(\pi h)^{2 n}} \int_{\mathbb{R}^{2 n}} \int_{\mathbb{R}^{2 n}} e^{-\frac{2 i}{h} \sigma\left(w_{1}, w_{2}\right)} a\left(z+w_{1}\right) b\left(z+w_{2}\right) d w_{1} d w_{2},
\end{aligned}
$$

where $z=(x, \xi)$.

Proof. 1. For $l \in \mathbb{R}^{2 n}$, define

$$
\hat{a}(l):=\int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{-\frac{i}{h} l(x, \xi)} a(x, \xi) d x d \xi ;
$$

then

$$
a(x, \xi)=\frac{1}{(2 \pi h)^{2 n}} \int_{\mathbb{R}^{2 n}} e^{\frac{i}{h} l(x, \xi)} \hat{a}(l) d l
$$

Therefore Theorem 4.4,(i) implies

$$
\begin{equation*}
a^{\mathrm{w}}(x, h D)=\frac{1}{(2 \pi h)^{2 n}} \int_{\mathbb{R}^{2 n}} \hat{a}(l) e^{\frac{i}{h} l(x, h D)} d l, \tag{4.17}
\end{equation*}
$$

and likewise

$$
b^{\mathrm{w}}(x, h D)=\frac{1}{(2 \pi h)^{2 n}} \int_{\mathbb{R}^{2 n}} \hat{b}(m) e^{\frac{i}{h} m(x, h D)} d m .
$$

Theorem 4.4,(ii) lets us next compute

$$
\begin{aligned}
a^{\mathrm{w}}(x, h D) & b^{\mathrm{w}}(x, h D) \\
& =\frac{1}{(2 \pi h)^{4 n}} \int_{\mathbb{R}^{2 n}} \int_{\mathbb{R}^{2 n}} \hat{a}(l) \hat{b}(m) e^{\frac{i}{h} l(x, h D)} e^{\frac{i}{h} m(x, h D)} d m d l \\
& =\frac{1}{(2 \pi h)^{4 n}} \int_{\mathbb{R}^{2 n}} \int_{\mathbb{R}^{2 n}} \hat{a}(l) \hat{b}(m) e^{\frac{i}{2 h} \sigma(l, m)} e^{\frac{i}{h}(l+m)(x, h D)} d l d m \\
& =\frac{1}{(2 \pi h)^{2 n}} \int_{\mathbb{R}^{2 n}} \hat{c}(r) e^{\frac{i}{h} r(x, h D)} d r
\end{aligned}
$$

for

$$
\begin{equation*}
\hat{c}(r):=\frac{1}{(2 \pi h)^{2 n}} \int_{\{l+m=r\}} \hat{a}(l) \hat{b}(m) e^{\frac{i \sigma(l, m)}{2 h}} d l . \tag{4.18}
\end{equation*}
$$

To get this, we changed variables by setting $r=m+l$.
2. We will show that $\hat{c}$ defined by (4.18) is the Fourier transform of the symbol $c$ defined by the right hand side of (4.15). We first simplify notation by writing $z=(x, \xi), w=(y, \eta)$. Then

$$
c(z)=\left.e^{\frac{i h}{2} \sigma\left(D_{z}, D_{w}\right)} a(z) b(w)\right|_{w=z}=\left.e^{\frac{i}{2 h} \sigma\left(h D_{z}, h D_{w}\right)} a(z) b(w)\right|_{w=z}
$$

and

$$
\begin{aligned}
a(z) & =\frac{1}{(2 \pi h)^{2 n}} \int_{\mathbb{R}^{2 n}} e^{\frac{i}{h} l(z)} \hat{a}(l) d l \\
b(w) & =\frac{1}{(2 \pi h)^{2 n}} \int_{\mathbb{R}^{2 n}} e^{\frac{i}{h} m(w)} \hat{b}(m) d m
\end{aligned}
$$

Furthermore, a direct calculation, the details of which we leave to the reader, demonstrates that

$$
e^{\frac{i}{2 h} \sigma\left(h D_{z}, h D_{w}\right)} e^{\frac{i}{h}(l(z)+m(w))}=e^{\frac{i}{h}(l(z)+m(w))+\frac{i}{2 h} \sigma(l, m)} .
$$

Consequently

$$
\begin{aligned}
c(z) & =\left.\frac{1}{(2 \pi h)^{4 n}} \int_{\mathbb{R}^{2 n}} \int_{\mathbb{R}^{2 n}} e^{\frac{i}{2 h} \sigma\left(h D_{z}, h D_{w}\right)} e^{\frac{i}{h}(l(z)+m(w))}\right|_{z=w} \hat{a}(l) \hat{b}(m) d l d m \\
& =\frac{1}{(2 \pi h)^{4 n}} \int_{\mathbb{R}^{2 n}} \int_{\mathbb{R}^{2 n}} e^{\frac{i}{h}(l(z)+m(z))+\frac{i}{2 h} \sigma(l, m)} \hat{a}(l) \hat{b}(m) d l d m .
\end{aligned}
$$

The semiclassical Fourier transform of $c$ is therefore
$\frac{1}{(2 \pi h)^{2 n}} \int_{\mathbb{R}^{2 n}} \int_{\mathbb{R}^{2 n}}\left(\frac{1}{(2 \pi h)^{2 n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}(l+m-r)(z)} d z\right) e^{\frac{i}{2 h} \sigma(l, m)} \hat{a}(l) \hat{b}(m) d l d m$.
According to (3.23), the term inside the parentheses is $\delta_{\{l+m=r\}}$ in $\mathcal{S}^{\prime}$. Thus the foregoing equals

$$
\frac{1}{(2 \pi h)^{2 n}} \int_{\{l+m=r\}} e^{\frac{i}{2 h} \sigma(l, m)} \hat{a}(l) \hat{b}(m) d l=\hat{c}(r),
$$

in view of (4.18).
3. Formula (4.16) follows from Theorem 4.5,(iii), with $\frac{h}{2}$ replacing $h$.
4.2.4 Asymptotics. We next apply stationary phase to derive a useful asymptotic expansion of $a \# b$ :

THEOREM 4.7 (Semiclassical expansions). Assume $a, b \in \mathcal{S}$.
(i) We have for $N=0,1, \ldots$,

$$
\begin{align*}
& a \# b(x, \xi)=  \tag{4.19}\\
& \left.\sum_{k=0}^{N} \frac{h^{k}}{k!}\left(\frac{i}{2} \sigma\left(D_{x}, D_{\xi}, D_{y}, D_{\eta}\right)\right)^{k}(a(x, \xi) b(y, \eta))\right|_{\substack{y=x \\
\eta=\xi}}+O\left(h^{N+1}\right)
\end{align*}
$$

as $h \rightarrow 0$, the error taken in $\mathcal{S}$.
(ii) In particular,

$$
\begin{equation*}
a \# b=a b+\frac{h}{2 i}\{a, b\}+O\left(h^{2}\right) ; \tag{4.20}
\end{equation*}
$$

and

$$
\begin{equation*}
\left[a^{\mathrm{w}}, b^{\mathrm{w}}\right]=\frac{h}{i}\{a, b\}^{\mathrm{w}}+O\left(h^{2}\right) \tag{4.21}
\end{equation*}
$$

(iii) If $\operatorname{spt}(a) \cap \operatorname{spt}(b)=\emptyset$, then

$$
a \# b=O\left(h^{\infty}\right)
$$

Proof. 1. To prove (4.19), we apply the stationary phase Theorem 3.16,(ii), with $\frac{h}{2}$ replacing $h$ and $-\sigma$ replacing $\sigma$, to the integral formula (4.16).
2. Next, compute

$$
\begin{aligned}
a \# b & =a b+\left.\frac{i h}{2} \sigma\left(D_{x}, D_{\xi}, D_{y}, D_{\eta}\right) a(x, \xi) b(y, \eta)\right|_{\substack{y=x \\
\eta=\xi}}+O\left(h^{2}\right) \\
& =a b+\left.\frac{i h}{2}\left(\left\langle D_{\xi} a, D_{y} b\right\rangle-\left\langle D_{x} a, D_{\eta} b\right\rangle\right)\right|_{\substack{y=x \\
\eta=\xi}}+O\left(h^{2}\right) \\
& =a b+\frac{h}{2 i}\left(\left\langle\partial_{\xi} a, \partial_{x} b\right\rangle-\left\langle\partial_{x} a, \partial_{\xi} b\right\rangle\right)+O\left(h^{2}\right) \\
& =a b+\frac{h}{2 i}\{a, b\}+O\left(h^{2}\right) .
\end{aligned}
$$

Consequently,

$$
\begin{aligned}
{\left[a^{\mathrm{w}}, b^{\mathrm{w}}\right] } & =a^{\mathrm{w}} b^{\mathrm{w}}-b^{\mathrm{w}} a^{\mathrm{w}}=(a \# b-b \# a)^{\mathrm{w}} \\
& =\left(a b+\frac{h}{2 i}\{a, b\}-\left(b a+\frac{h}{2 i}\{b, a\}\right)+O\left(h^{2}\right)\right)^{\mathrm{w}} \\
& =\frac{h}{i}\{a, b\}^{\mathrm{w}}+O\left(h^{2}\right) .
\end{aligned}
$$

3. If $\operatorname{spt}(a) \cap \operatorname{spt}(b)=\emptyset$, each term in the expansion (4.19) vanishes.
4.2.5 Standard quantization. Next we replace Weyl $\left(t=\frac{1}{2}\right)$ by standard $(t=1)$ quantization in our formulas. The proofs are simpler.

## THEOREM 4.8 (Composition for standard quantization).

(i) Let $a, b \in \mathcal{S}$. Then

$$
a(x, h D) b(x, h D)=c(x, h D)
$$

for the symbol

$$
\begin{equation*}
c(x, \xi)=\left.e^{i h\left\langle D_{\xi}, D_{y}\right\rangle}(a(x, \xi) b(y, \eta))\right|_{\substack{y=x \\ \eta=\xi}} . \tag{4.22}
\end{equation*}
$$

(ii) We have the integral representation formula

$$
\begin{align*}
& c(x, \xi):=  \tag{4.23}\\
& \quad \frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{-\frac{i}{h}\left\langle x_{1}, \xi_{1}\right\rangle} a\left(x, \xi+\xi_{1}\right) b\left(x+x_{1}, \xi\right) d x_{1} d \xi_{1} .
\end{align*}
$$

(iii) For each $N=0,1, \ldots$,

$$
\begin{equation*}
c(x, \xi)=\left.\sum_{k=0}^{N} \frac{h^{k}}{k!}\left(i\left\langle D_{\xi}, D_{y}\right\rangle\right)^{k}(a(x, \xi) b(y, \eta))\right|_{\substack{y=x \\ \eta=\xi}}+O\left(h^{N+1}\right) \tag{4.24}
\end{equation*}
$$

as $h \rightarrow 0$, the error taken in $\mathcal{S}$.

Proof. Let $u \in \mathcal{S}$. Then

$$
\begin{aligned}
& a(x, h D) b(x, h D) u(x) \\
&=\frac{1}{(2 \pi h)^{2 n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}(\langle x, \eta\rangle+\langle y, \xi-\eta\rangle)} a(x, \eta) b(y, \xi) \hat{u}(\xi) d y d \eta d \xi \\
& \quad=\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} c(x, \xi) e^{\frac{i}{h}\langle x, \xi\rangle} \hat{u}(\xi) d \xi, \\
& \quad=c(x, h D) u(x)
\end{aligned}
$$

for

$$
c(x, \xi)=\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{-\frac{i}{h}\langle x-y, \xi-\eta\rangle} a(x, \eta) b(y, \xi) d y d \eta
$$

Change variables by putting $x_{1}=y-x, \xi_{1}=\eta-\xi$, to rewrite $c$ in the form (4.23). Then (4.22) is a consequence of Theorem 4.5,(ii).

Finally, the stationary phase Theorem 3.16,(i) provides the asymptotic expansion (4.24)

THEOREM 4.9 (Adjoints for standard quantization). If $a \in \mathcal{S}$, then

$$
a(x, h D)^{*}=b(x, h D),
$$

for

$$
\begin{equation*}
b(x, \xi):=e^{i\left\langle\left\langle D_{x}, D_{\xi}\right\rangle\right.} \bar{a}(x, \xi) . \tag{4.25}
\end{equation*}
$$

Proof. 1. We first observe that, as in the proof of Theorem 4.4,

$$
\mathrm{Op}_{t}\left(e^{\frac{i}{h} l(x, \xi)}\right) u(x)=e^{\frac{i}{h}\left\langle x, x^{*}\right\rangle+\frac{i}{h}(1-t)\left\langle x^{*}, \xi^{*}\right\rangle} u\left(x+\xi^{*}\right)
$$

It follows that

$$
\begin{equation*}
\mathrm{Op}_{t}\left(e^{\frac{i}{h} l(x, \xi)}\right)=e^{\frac{i}{h}(s-t)\left\langle x^{*}, \xi^{*}\right\rangle} \mathrm{Op}_{s}\left(e^{\frac{i}{h} l(x, \xi)}\right) \tag{4.26}
\end{equation*}
$$

Next we record an interesting conversion formula, namely that if

$$
A=\mathrm{Op}_{t}\left(a_{t}\right) \quad(0 \leq t \leq 1)
$$

then

$$
\begin{equation*}
a_{t}(x, \xi)=e^{i(t-s) h\left\langle D_{x}, D_{\xi}\right\rangle} a_{s}(x, \xi) . \tag{4.27}
\end{equation*}
$$

To see this, notice that the decomposition formula (4.17) implies

$$
\mathrm{Op}_{t}\left(a_{t}\right)=\frac{1}{(2 \pi h)^{2 n}} \int_{\mathbb{R}^{2 n}} \hat{a}_{t}(l) \mathrm{Op}_{t}\left(e^{\frac{i}{h} l}\right) d l
$$

We apply (4.26) to derive (4.27).
2. Next recall (4.4):

$$
a(x, h D)^{*}=\mathrm{Op}_{1}(a)^{*}=\mathrm{Op}_{0}(\bar{a})
$$

We now invoke (4.27), to write

$$
\mathrm{Op}_{0}(\bar{a})=\mathrm{Op}_{1}(b),
$$

the symbol $b$ defined by (4.25).

### 4.3 GENERAL SYMBOL CLASSES

We next extend our calculus to symbols $a=a(x, \xi, h)$ which depend on the parameter $h$ and which can grow, along with their derivatives, as $|x|,|\xi| \rightarrow \infty$.

### 4.3.1 More definitions.

DEFINITION. A function $m: \mathbb{R}^{2 n} \rightarrow(0, \infty)$ is called an order function if there exist constants $C, N$ such that

$$
\begin{equation*}
m(w) \leq C\langle z-w\rangle^{N} m(z) \tag{4.28}
\end{equation*}
$$

for all $w, z \in \mathbb{R}^{n}$.
Observe that if $m_{1}, m_{2}$ are order functions, so is $m_{1} m_{2}$.
EXAMPLES. Standard examples are $m(z) \equiv 1$ and $m(z)=\langle z\rangle=$ $\left(1+|z|^{2}\right)^{1 / 2}$.

## DEFINITIONS.

(i) Given an order function $m$ on $\mathbb{R}^{2 n}$, we define the corresponding class of symbols:

$$
\begin{aligned}
& S(m):=\left\{a \in C^{\infty} \mid \text { for each multiindex } \alpha\right. \\
&\left.\quad \text { there exists a constant } C_{\alpha} \text { so that }\left|\partial^{\alpha} a\right| \leq C_{\alpha} m\right\} .
\end{aligned}
$$

(ii) We as well define

$$
S^{k}(m):=\left\{a \in C^{\infty}| | \partial^{\alpha} a \mid \leq C_{\alpha} h^{-k} m \text { for all multiindices } \alpha\right\}
$$

and

$$
S_{\delta}^{k}(m):=\left\{a \in C^{\infty}| | \partial^{\alpha} a \mid \leq C_{\alpha} h^{-\delta|\alpha|-k} m \text { for all multiindices } \alpha\right\}
$$

The index $k$ indicates how singular is the symbol $a$ as $h \rightarrow 0$; the index $\delta$ allows for increasing singularity of the higher derivatives. Notice that
the more negative $k$ is, the more rapidly $a$ and its derivatives vanish as $h \rightarrow 0$.
(iii) Write also

$$
S^{-\infty}(m):=\left\{a \in C^{\infty} \mid \text { for each } \alpha \text { and } N,\left|\partial^{\alpha} a\right| \leq C_{\alpha, N} h^{N} m\right\}
$$

So if $a$ is a symbol belonging to $S^{-\infty}(m)$, then $a$ and all of its derivatives are $O\left(h^{\infty}\right)$ as $h \rightarrow 0$.

NOTATION. If the order function is the constant function $m \equiv 1$, we will usually not write it:

$$
S^{k}:=S^{k}(1), S_{\delta}^{k}:=S_{\delta}^{k}(1)
$$

We will also omit zero superscripts. Thus

$$
\begin{aligned}
S & :=\left\{a \in C^{\infty}\left(\mathbb{R}^{2 n}\right)| | \partial^{\alpha} a \mid \leq C_{\alpha} \text { for all multiindices } \alpha\right\} \\
S_{\delta} & :=\left\{a \in C^{\infty}| | \partial^{\alpha} a \mid \leq C_{\alpha} h^{-\delta|\alpha|} \text { for all multiindices } \alpha\right\}
\end{aligned}
$$

## REMARKS: rescaling in $h$.

(i) We will show in the next subsection that if $a \in S_{\delta}$, then the quantization formula

$$
a^{\mathrm{w}}(x, h D) u(x):=\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}\langle x-y, \xi\rangle} a\left(\frac{x+y}{2}, \xi\right) u(y) d \xi d y
$$

makes sense for $u \in \mathcal{S}$. It is often convenient to rescale to the case $h=1$, by changing to the new variables

$$
\begin{equation*}
\tilde{x}:=h^{-\frac{1}{2}} x, \tilde{y}:=h^{-\frac{1}{2}} y, \tilde{\xi}:=h^{-\frac{1}{2}} \xi . \tag{4.29}
\end{equation*}
$$

Then

$$
\begin{aligned}
& a^{\mathrm{w}}(x, h D) u(x) \\
&=\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} a\left(\frac{x+y}{2}, \xi\right) e^{\frac{i}{h}\langle x-y, \xi\rangle} u(y) d y d \xi \\
&=\frac{1}{(2 \pi)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} a_{h}\left(\frac{\tilde{x}+\tilde{y}}{2}, \tilde{\xi}\right) e^{i\langle\tilde{x}-\tilde{y}, \tilde{\xi}\rangle} \tilde{u}(\tilde{y}) d \tilde{y} d \tilde{\xi} ;
\end{aligned}
$$

and therefore

$$
\begin{equation*}
a^{\mathrm{w}}(x, h D) u(x)=a_{h}^{\mathrm{w}}(\tilde{x}, D) \tilde{u}(\tilde{x}), \tag{4.30}
\end{equation*}
$$

for

$$
\begin{equation*}
\tilde{u}(\tilde{x}):=u(x)=u\left(h^{\frac{1}{2}} \tilde{x}\right), \quad a_{h}(\tilde{x}, \tilde{\xi}):=a(x, \xi)=a\left(h^{\frac{1}{2}} \tilde{x}, h^{\frac{1}{2}} \tilde{\xi}\right) . \tag{4.31}
\end{equation*}
$$

(ii) Observe also that if $a \in S_{\delta}$, then

$$
\begin{equation*}
\left|\partial^{\alpha} a_{h}\right|=h^{\left.\frac{|\alpha|}{2} \right\rvert\,}\left|\partial^{\alpha} a\right| \leq C_{\alpha} h^{|\alpha|\left(\frac{1}{2}-\delta\right)} \tag{4.32}
\end{equation*}
$$

for each multiindex $\alpha$. If $\delta>\frac{1}{2}$, the last term is unbounded as $h \rightarrow 0$; and consequently we will henceforth always assume

$$
0 \leq \delta \leq \frac{1}{2}
$$

We see also that the case

$$
\delta=\frac{1}{2}
$$

is critical, in that we do not then get decay as $h \rightarrow 0$ for the terms on the right hand side of (4.32) when $|\alpha|>0$.
4.3.2 Quantization. Next we discuss the Weyl quantization of symbols in the class $S_{\delta}(m)$ :

THEOREM 4.10 (Quantizing general symbols). If $a \in S_{\delta}(m)$, then

$$
a^{\mathrm{w}}(x, h D): \mathcal{S} \rightarrow \mathcal{S} .
$$

Proof. 1. We take $h=1$ for simplicity; so that

$$
a^{\mathrm{w}}(x, D) u(x)=\frac{1}{(2 \pi)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{i\langle x-y, \xi\rangle} a\left(\frac{x+y}{2}, \xi\right) u(y) d y d \xi
$$

for $u \in \mathcal{S}$. Observe next that $L_{1} e^{i\langle x-y, \xi\rangle}=e^{i\langle x-y, \xi\rangle}$, where

$$
L_{1}:=\frac{1+\left\langle x-y, D_{\xi}\right\rangle}{1+|x-y|^{2}} ;
$$

and $L_{2} e^{i\langle x-y, \xi\rangle}=e^{i\langle x-y, \xi\rangle}$ for

$$
L_{2}:=\frac{1-\left\langle\xi, D_{y}\right\rangle}{1+|\xi|^{2}} .
$$

We employ these operators and the usual integration by parts argument, to show $a^{\mathrm{w}}(x, D): \mathcal{S} \rightarrow L^{\infty}$.
2. Furthermore,

$$
x_{j} a^{\mathrm{w}}(x, D) u=\frac{1}{(2 \pi)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}}\left(D_{\xi_{j}}+y_{j}\right) e^{i\langle x-y, \xi\rangle} a u d y d \xi
$$

We integrate by parts, to conclude that $x^{\alpha} a^{\mathrm{w}}(x, D): \mathcal{S} \rightarrow L^{\infty}$ for each multinomial $x^{\alpha}$. Also, since

$$
\left(e^{-\frac{i}{2}\left\langle D_{x}, D_{\xi}\right\rangle} a\right)(x, D) u=a^{\mathrm{w}}(x, D) u
$$

we have

$$
\begin{aligned}
& D_{x_{j}} a^{\mathrm{w}}(x, D) u=D_{x_{j}}\left(e^{-\frac{i}{2}\left\langle D_{x}, D_{\xi}\right\rangle} a\right)(x, D) u \\
& \quad=D_{x_{j}}\left(\frac{1}{(2 \pi)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{-\frac{i}{2}\left\langle D_{y}, D_{\xi}\right\rangle} a(y, \xi) e^{i\langle x-y, \xi\rangle} u(y) d y d \xi\right) \\
& \quad=\frac{1}{(2 \pi)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{-\frac{i}{2}\left\langle D_{y}, D_{\xi}\right\rangle} a(y, \xi)\left(-D_{y_{j}} e^{i\langle x-y, \xi\rangle}\right) u(y) d y d \xi .
\end{aligned}
$$

Again integrate by parts, to deduce $D^{\beta} a^{\mathrm{w}}(x, D): \mathcal{S} \rightarrow L^{\infty}$ for each partial derivative $D^{\beta}$.

Consequently, $D^{\beta} x^{\alpha} a^{\mathrm{w}}(x, D): \mathcal{S} \rightarrow L^{\infty}$, for all multiindices $\alpha, \beta$. It follows that $a^{\mathrm{w}}(x, D): \mathcal{S} \rightarrow \mathcal{S}$.
4.3.3 Asymptotic series. Next we consider infinite sums of terms in various symbol classes.
DEFINITION. Let $a \in S_{\delta}^{k_{0}}(m)$ and $a_{j} \in S_{\delta}^{k_{j}}(m)$, where $k_{j+1}<k_{j}$, $k_{j} \rightarrow-\infty$. We say that $a$ is asymptotic to $\sum a_{j}$, and write

$$
a \sim \sum_{j=0}^{\infty} a_{j}
$$

provided for each $N=1,2, \ldots$

$$
\begin{equation*}
a-\sum_{j=0}^{N-1} a_{j} \in S_{\delta}^{k_{N}}(m) . \tag{4.33}
\end{equation*}
$$

INTERPRETATION. Observe that for each $h>0$, the series $\sum_{j=0}^{\infty} a_{j}$ need not converge in any sense. We are requiring rather in (4.33) that for each $N$, the difference $a-\sum_{j=0}^{N-1} a_{j}$, and its derivatives, vanish at appropriate rates as $h \rightarrow 0$.

Perhaps surprisingly, we can always construct such an asymptotic sum of symbols:

## THEOREM 4.11 (Borel's Theorem).

(i) Assume $a_{j} \in S_{\delta}^{k_{j}}(m)$, where $k_{j+1}<k_{j}, k_{j} \rightarrow-\infty$. Then there exists a symbol $a \in S_{\delta}^{k_{0}}(m)$ such that

$$
a \sim \sum_{j=0}^{\infty} a_{j} .
$$

(ii) If also $\hat{a} \sim \sum_{j=0}^{\infty} a_{j}$, then

$$
a-\hat{a} \in S^{-\infty}(m) .
$$

Proof. 1. Choose a $C^{\infty}$ function $\chi$ such that

$$
0 \leq \chi \leq 1, \quad \chi \equiv 1 \text { on }[0,1], \chi \equiv 0 \text { on }[2, \infty)
$$

We define

$$
\begin{equation*}
a:=\sum_{j=0}^{\infty} a_{j} \chi\left(\lambda_{j} h\right) \tag{4.34}
\end{equation*}
$$

where the sequence $\lambda_{j} \rightarrow \infty$ must be selected. Since $\lambda_{j} \rightarrow \infty$, there are for each $h>0$ at most finitely many nonzero terms in the sum (4.34).

Now for each multiindex $\alpha$, with $|\alpha| \leq j$, we have

$$
\begin{align*}
\left|\partial^{\alpha}\left(a_{j} \chi\left(\lambda_{j} h\right)\right)\right| & =\left|\left(\partial^{\alpha} a_{j}\right) \chi\left(\lambda_{j} h\right)\right| \\
& \leq C_{j, \alpha} h^{-k_{j}-\delta|\alpha|} m \chi\left(\lambda_{j} h\right) \\
& =C_{j, \alpha} h^{-k_{j}-\delta|\alpha|} \frac{\lambda_{j} h}{\lambda_{j} h} m \chi\left(\lambda_{j} h\right)  \tag{4.35}\\
& \leq 2 C_{j, \alpha} \frac{h^{-k_{j}-1-\delta|\alpha|}}{\lambda_{j}} m \\
& \leq h^{-k_{j}-1-|\alpha| \delta} 2^{-j} m
\end{align*}
$$

if $\lambda_{j}$ is selected sufficiently large. We can accomplish this for all $j$ and multiindices $\alpha$ with $|\alpha| \leq j$. We may assume also $\lambda_{j+1} \geq \lambda_{j}$, for all $j$.
2. Then

$$
a-\sum_{j=0}^{N} a_{j}=\sum_{j=N+1}^{\infty} a_{j} \chi\left(\lambda_{j} h\right)+\sum_{j=0}^{N} a_{j}\left(\chi\left(\lambda_{j} h\right)-1\right) .
$$

Fix any multiindex $\alpha$. Then taking $N \geq|\alpha|$, we have

$$
\begin{aligned}
\left|\partial^{\alpha}\left(a-\sum_{j=0}^{N} a_{j}\right)\right| \leq & \sum_{j=N+1}^{\infty}\left|\left(\partial^{\alpha} a_{j}\right)\right| \chi\left(\lambda_{j} h\right) \\
& +\sum_{j=0}^{N}\left|\partial^{\alpha} a_{j}\right|\left(1-\chi\left(\lambda_{j} h\right)\right) \\
=: & A+B
\end{aligned}
$$

According to estimate (4.35),

$$
A \leq \sum_{j=N+1}^{\infty} h^{-k_{j}-1-\delta|\alpha|} 2^{-j} m \leq m h^{-k_{N+1}-1-\delta|\alpha|}
$$

Also

$$
B \leq \sum_{j=0}^{N} C_{\alpha, j} h^{-k_{j}-\delta|\alpha|} m\left(1-\chi\left(\lambda_{j} h\right)\right) .
$$

Since $\chi \equiv 1$ on $[0,1], B=0$ if $0<h \leq \lambda_{N}^{-1}$. If $\lambda_{N}^{-1} \leq h \leq 1$, we have $1 \leq \lambda_{N} h$ and hence

$$
\begin{aligned}
B & \leq m \sum_{j=0}^{N} C_{\alpha, j} h^{-|\alpha| \delta} \leq m \sum_{j=0}^{N} C_{\alpha, j} \lambda_{N}^{-k_{N}} h^{-k_{N}-\delta|\alpha|} \\
& =m \widetilde{C}_{\alpha, N} h^{-k_{N}-\delta|\alpha|} .
\end{aligned}
$$

Thus

$$
\left|\partial^{\alpha}\left(a-\sum_{j=0}^{N} a_{j}\right)\right| \leq C_{\alpha, N} h^{-k_{N}-\delta|\alpha|} m
$$

if $N \geq|\alpha|$. Therefore, for any $N$

$$
\left|\partial^{\alpha}\left(a-\sum_{j=0}^{N-1} a_{j}\right)\right| \leq C_{\alpha, N} h^{-k_{N}-\delta|\alpha|} m .
$$

4.3.4 Semiclassical expansions in $\boldsymbol{S}_{\boldsymbol{\delta}}$. Next we need to reexamine some of our earlier asymptotic expansions, deriving improved estimates on the error terms:

THEOREM 4.12 (Semiclassical expansions in $\boldsymbol{S}_{\boldsymbol{\delta}}$.). Let $Q$ be symmetric, nonsingular matrix.
(i) If $0 \leq \delta \leq \frac{1}{2}$, then

$$
e^{\frac{i h}{2}\langle Q D, D\rangle}: S_{\delta}(m) \rightarrow S_{\delta}(m)
$$

(ii) If $0 \leq \delta<\frac{1}{2}$, we furthermore have for each symbol $a \in S_{\delta}(m)$ the expansion

$$
\begin{equation*}
e^{\frac{i h}{2}\langle Q D, D\rangle} a \sim \sum_{k=0}^{\infty} \frac{h^{k}}{k!}\left(i \frac{\langle Q D, D\rangle}{2}\right)^{k} a \quad \text { in } S_{\delta}(m) . \tag{4.36}
\end{equation*}
$$

REMARK. Since we can always rescale to the case $h=1$, there cannot exist an expansion like (4.36) for $\delta=1 / 2$.

Proof. 1. First, let $0 \leq \delta<\frac{1}{2}$ and $a \in S_{\delta}(m)$. Recall from Theorem 4.5,(i) that

$$
e^{\frac{i h}{2}\langle Q D, D\rangle} a(z)=\frac{|\operatorname{det} Q|^{-\frac{1}{2}}}{(2 \pi h)^{n}} e^{\frac{i \pi}{4} \operatorname{sgn} Q} \int_{\mathbb{R}^{2 n}} e^{\varphi(w)} a(z+w) d w
$$

for the quadratic phase

$$
\varphi(w):=-\frac{1}{2}\left\langle Q^{-1} w, w\right\rangle
$$

Let $\chi: \mathbb{R}^{n} \rightarrow \mathbb{R}$ be a smooth function with $\chi \equiv 1$ on $B(0,1), \chi \equiv 0$ on $\mathbb{R}^{n}-B(0,2)$. Then

$$
\begin{aligned}
e^{\frac{i h}{2}\langle Q D, D\rangle} a(z)= & \frac{C}{h^{n}} \int_{\mathbb{R}^{2 n}} e^{\frac{i \varphi(w)}{h}} a(z-w) d w \\
= & \frac{C}{h^{n}} \int_{\mathbb{R}^{2 n}} e^{\frac{i \varphi(w)}{h}} \chi(w) a(z-w) d w \\
& +\frac{C}{h^{n}} \int_{\mathbb{R}^{2 n}} e^{\frac{i \varphi(w)}{h}}(1-\chi(w)) a(z-w) d w \\
= & A+B
\end{aligned}
$$

for the constant

$$
C:=\frac{|\operatorname{det} Q|^{-\frac{1}{2}}}{(2 \pi)^{n}} e^{\frac{i \pi}{4} \operatorname{sgn} Q}
$$

2. Estimate of $A$. Since $\chi(w) a(z-w)$ has compact support, the method of stationary phase, Theorem 4.5, gives

$$
A \sim \sum_{k=0}^{\infty} \frac{h^{k}}{k!}(i\langle Q D, D\rangle)^{k} a(z)
$$

Furthermore, if $|w| \leq 2$, we have $m(z-w) \leq C m(z)$. Consequently, $|A| \leq C m(z)$; and similarly

$$
\left|\partial^{\alpha} A\right| \leq C \sup _{0 \leq \beta \leq \alpha}\left|\partial^{\beta} a\right| \leq C h^{-|\alpha| \delta} m(z) .
$$

Hence $A \in S_{\delta}(m)$.
3. Estimate of $B$. Let

$$
L:=\frac{\langle\partial \varphi, h D\rangle}{|\partial \varphi|^{2}}
$$

then $L e^{i \varphi / h}=e^{i \varphi / h}$ Furthermore, since $|\partial \varphi(w)| \geq \gamma|w|$ for some positive constant $\gamma$, the operator $L$ has smooth coefficients on the support
of $1-\chi$ and

$$
\left|\left(L^{*}\right)^{M}((1-\chi) a)\right| \leq C_{M} \frac{h^{M}}{\langle w\rangle^{M}} \sup _{|\alpha| \leq M}\left|\partial^{\alpha} a(z-w)\right|
$$

Consequently,

$$
\begin{aligned}
|B| & =\frac{C}{h^{n}}\left|\int_{\mathbb{R}^{2 n}}\left(L^{M} e^{i \varphi / h}\right)(1-\chi(w)) a(z-w) d w\right| \\
& =\frac{C}{h^{n}}\left|\int_{\mathbb{R}^{2 n}} e^{i \varphi / h}\left(L^{*}\right)^{M}((1-\chi) a) d w\right| \\
& \leq C h^{M-n} \int_{\mathbb{R}^{2 n}}\langle w\rangle^{-M} \sup _{|\alpha| \leq M}\left|\partial^{\alpha} a(z-w)\right| d w \\
& \leq C h^{M-n-\delta M} \int_{\mathbb{R}^{2 n}}\langle w\rangle^{N-M} m(z) d w \\
& =C h^{M-n-\delta M} m(z),
\end{aligned}
$$

provided $M>2 n+N$. The number $N$ is from the definition (4.28) of the order function $m$.

We similarly check also the higher derivatives, to conclude that $B \in$ $S_{\delta}^{-\infty}(m)$.
4. Now assume $\delta=1 / 2$. In this case we can rescale, by setting

$$
\tilde{w}=w h^{-1 / 2} .
$$

Then

$$
e^{i h\langle Q D, D\rangle} a(z)=C \int_{\mathbb{R}^{n}} e^{i \varphi(\tilde{w})} a\left(z-\tilde{w} h^{1 / 2}\right) d \tilde{w} .
$$

We use $\chi$ to break the integral into two pieces $A$ and $B$, as above.

## THEOREM 4.13 (Symbol class of $a \# b)$.

(i) If $a \in S_{\delta}\left(m_{1}\right)$ and $b \in S_{\delta}\left(m_{2}\right)$, then

$$
\begin{equation*}
a \# b \in S_{\delta}\left(m_{1} m_{2}\right) \tag{4.37}
\end{equation*}
$$

and

$$
a^{\mathrm{w}}(x, h D) b^{\mathrm{w}}(x, h D)=(a \# b)^{\mathrm{w}}(x, h D)
$$

in the sense of operators mapping $\mathcal{S}$ to $\mathcal{S}$.
(ii) Furthermore,

$$
\begin{equation*}
a \# b-a b \in S_{\delta}^{2 \delta-1}\left(m_{1} m_{2}\right) . \tag{4.38}
\end{equation*}
$$

Proof. 1. Clearly

$$
c(w, z):=a(w) b(z) \in S_{\delta}\left(m_{1}(w) m_{2}(z)\right)
$$

in $\mathbb{R}^{4 n}$. If we put $D=\left(D_{x}, D_{\xi}, D_{y}, D_{\eta}\right)$ and

$$
\langle Q D, D\rangle=\sigma\left(D_{x}, D_{\xi} ; D_{y}, D_{\eta}\right),
$$

for $w=(x, \xi)$ and $z=(y, \eta)$, then according to Theorem 4.12, we have

$$
e^{\frac{1}{2 h}\langle Q D, D\rangle} c \in S_{\delta}\left(m_{1}(w) m_{2}(z)\right)
$$

Since (4.16) says

$$
a \# b(z)=e^{\frac{1}{2 h}\langle Q D, D\rangle} c(z, z)
$$

assertion (4.37) follows.
The second statement of assertion (i) follows from the density of $\mathcal{S}$ in $S_{\delta}(m)$.
2. We leave the verification of (4.38) as an exercise.

Finally we describe how to obtain the symbol from the operator, in the particularly nice case of the standard quantization:

THEOREM 4.14 (Constructing the symbol from the operator). Suppose $a \in S_{\delta}(m)$. Then

$$
\begin{equation*}
a(x, \xi)=e^{\frac{i}{h}\langle x, \xi\rangle} a(x, D)\left(e^{\frac{i}{h}\langle\cdot, \xi\rangle}\right) \tag{4.39}
\end{equation*}
$$

Proof. For $a \in \mathcal{S}$ we verify this formula using the inverse Fourier transform:

$$
\begin{aligned}
& \frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} a(x, \eta) e^{\frac{i}{h}\langle x, \eta-\xi\rangle} e^{-\frac{i}{h}\langle y, \eta-\xi\rangle} d y d \xi= \\
& \quad \int_{\mathbb{R}^{n}} a(x, \eta) \delta_{0}(\xi-\eta) e^{\frac{i}{h}\langle x, \eta-\xi\rangle} d \eta=a(x, \xi) .
\end{aligned}
$$

Approximation of $a$ by elements of $\mathcal{S}$ concludes the proof.

### 4.4 OPERATORS ON $L^{2}$.

So far our symbol calculus has built operators acting on either the Schwartz space $\mathcal{S}$ of smooth functions or its dual space $\mathcal{S}^{\prime}$. But for applications we would like to handle functions in more convenient spaces, most notably $L^{2}$. Our next goal is therefore showing that if $a \in S_{\delta}$ for some $0 \leq \delta \leq \frac{1}{2}$, then $a^{\mathrm{w}}(x, h D)$ extends to become a bounded linear operator acting upon $L^{2}$.

For the time being, we take

$$
h=1 .
$$

Preliminaries. We select $\chi \in C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{2 n}\right)$ such that

$$
0 \leq \chi \leq 1, \quad \chi \equiv 0 \quad \text { on } \mathbb{R}^{2 n}-B(0,2)
$$

and

$$
\sum_{\alpha \in \mathbb{Z}^{2 n}} \chi_{\alpha} \equiv 1
$$

where $\chi_{\alpha}:=\chi(\cdot-\alpha)$ denotes $\chi$ shifted by the lattice point $\alpha \in \mathbb{Z}^{2 n}$. Write

$$
a_{\alpha}:=\chi_{\alpha} a ;
$$

then

$$
a=\sum_{\alpha \in \mathbb{Z}^{2 n}} a_{\alpha} .
$$

We also define

$$
\begin{equation*}
b_{\alpha \beta}:=\bar{a}_{\alpha} \# a_{\beta} \quad\left(\alpha, \beta \in \mathbb{Z}^{2 n}\right) . \tag{4.40}
\end{equation*}
$$

## THEOREM 4.15 (Decay of mixed terms).

(i) For each $N$ and each multiindex $\gamma$, we have the estimate

$$
\begin{equation*}
\left|\partial^{\gamma} b_{\alpha \beta}(z)\right| \leq C_{\gamma, N}\langle\alpha-\beta\rangle^{-N}\left\langle z-\frac{\alpha+\beta}{2}\right\rangle^{-N} \tag{4.41}
\end{equation*}
$$

for $z=(x, \xi) \in \mathbb{R}^{2 n}$.
(ii) For each $N$, there exists a constant $C_{N}$ such that

$$
\begin{equation*}
\left\|b_{\alpha \beta}^{\mathrm{w}}(x, D)\right\|_{L^{2} \rightarrow L^{2}} \leq C_{N}\langle\alpha-\beta\rangle^{-N} \tag{4.42}
\end{equation*}
$$

for all $\alpha, \beta \in \mathbb{Z}^{2 n}$.

Proof. 1. We can rewrite formula (4.16) to read

$$
b_{\alpha \beta}(z)=\frac{1}{\pi^{2 n}} \int_{\mathbb{R}^{2 n}} \int_{\mathbb{R}^{2 n}} e^{i \varphi\left(w_{1}, w_{2}\right)} \bar{a}_{\alpha}\left(z-w_{1}\right) a_{\beta}\left(z-w_{2}\right) d w_{1} d w_{2},
$$

for $\varphi\left(w_{1}, w_{2}\right)=-2 \sigma\left(w_{1}, w_{2}\right)$.
Select $\zeta: \mathbb{R}^{4 n} \rightarrow \mathbb{R}$ such that

$$
0 \leq \zeta \leq 1, \zeta \equiv 1 \text { on } B(0,1), \zeta \equiv 0 \text { on } \mathbb{R}^{4 n}-B(0,2)
$$

Then

$$
\begin{aligned}
b_{\alpha \beta}(z)= & \frac{1}{\pi^{2 n}} \int_{\mathbb{R}^{2 n}} \int_{\mathbb{R}^{2 n}} e^{i \varphi} \zeta(w) \bar{a}_{\alpha}\left(z-w_{1}\right) a_{\beta}\left(z-w_{2}\right) d w_{1} d w_{2} \\
& +\frac{1}{\pi^{2 n}} \int_{\mathbb{R}^{2 n}} \int_{\mathbb{R}^{2 n}} e^{i \varphi}(1-\zeta(w)) \bar{a}_{\alpha}\left(z-w_{1}\right) a_{\beta}\left(z-w_{2}\right) d w_{1} d w_{2} \\
= & A+B .
\end{aligned}
$$

2. Estimate of $A$. We have

$$
|A| \leq C \iint_{\{|w| \leq 2\}}\left|\bar{a}_{\alpha}\left(z-w_{1}\right)\right|\left|a_{\beta}\left(z-w_{2}\right)\right| d w_{1} d w_{2}
$$

for $w=\left(w_{1}, w_{2}\right)$. The integrand equals

$$
\chi\left(z-w_{1}-\alpha\right) \chi\left(z-w_{2}-\beta\right)\left|a\left(z-w_{1}\right)\right|\left|a\left(z-w_{2}\right)\right|
$$

and thus vanishes, unless

$$
\left|z-w_{1}-\alpha\right| \leq 2 \text { and }\left|z-w_{2}-\beta\right| \leq 2
$$

But then

$$
|\alpha-\beta| \leq 4+\left|w_{1}\right|+\left|w_{2}\right| \leq 8
$$

and

$$
\left|z-\frac{\alpha+\beta}{2}\right| \leq 4+\left|w_{1}\right|+\left|w_{2}\right| \leq 8 .
$$

Hence

$$
|A| \leq C_{N}\langle\alpha-\beta\rangle^{-N}\left\langle z-\frac{\alpha+\beta}{2}\right\rangle^{-N}
$$

for any $N$. Similarly, for each multiindex $\gamma$ we can estimate

$$
\begin{equation*}
\left|\partial^{\gamma} A\right| \leq C_{N, \gamma}\langle\alpha-\beta\rangle^{-N}\left\langle z-\frac{\alpha+\beta}{2}\right\rangle^{-N} \tag{4.43}
\end{equation*}
$$

3. Estimate of $B$. We have

$$
|\partial \varphi(w)|=2|w|
$$

and $L e^{i \varphi}=e^{i \varphi}$, for

$$
L:=\frac{\langle\partial \varphi, D\rangle}{|\partial \varphi|^{2}} .
$$

Since the integrand of $B$ vanishes unless $|w| \geq 1$, the usual argument based on integration by parts shows that

$$
|B| \leq C_{M} \int_{\mathbb{R}^{2 n}} \int_{\mathbb{R}^{2 n}}\langle w\rangle^{-M} \bar{c}_{\alpha}\left(z-w_{1}\right) c_{\beta}\left(z-w_{2}\right) d w_{1} d w_{2}
$$

for appropriate functions $c_{\alpha}, c_{\beta}$, with $\operatorname{spt} c_{\alpha} \subseteq B(\alpha, 2)$, $\operatorname{spt} c_{\beta} \subseteq B(\beta, 2)$. Thus the integrand vanishes unless

$$
\frac{1}{c}\langle w\rangle \leq\langle\alpha-\beta\rangle,\left\langle z-\frac{\alpha+\beta}{2}\right\rangle \leq C\langle w\rangle
$$

Hence

$$
\begin{aligned}
|B| & \leq C_{M}\langle\alpha-\beta\rangle^{-N}\left\langle z-\frac{\alpha+\beta}{2}\right\rangle^{-N} \int_{\mathbb{R}^{2 n}} \int_{\mathbb{R}^{2 n}}\langle w\rangle^{2 N-M} d w_{1} d w_{2} \\
& \leq C_{M}\langle\alpha-\beta\rangle^{-N}\left\langle z-\frac{\alpha+\beta}{2}\right\rangle^{-N}
\end{aligned}
$$

if $M$ is large enough. Likewise,

$$
\begin{equation*}
\left|\partial^{\gamma} B\right| \leq C_{N, \gamma}\langle\alpha-\beta\rangle^{-N}\left\langle z-\frac{\alpha+\beta}{2}\right\rangle^{-N} \tag{4.44}
\end{equation*}
$$

4. Recall next that

$$
a^{\mathrm{w}}(x, D)=\frac{1}{(2 \pi)^{2 n}} \int_{\mathbb{R}^{2 n}} \hat{a}(l) e^{i l(x, D)} d l
$$

and that, owing to (4.9), $e^{i l(x, D)}$ is a unitary operator on $L^{2}$. Consequently

$$
\left\|a^{\mathrm{w}}(x, D)\right\|_{L^{2} \rightarrow L^{2}} \leq C \int_{\mathbb{R}^{2 n}}|\hat{a}(l)| d l
$$

Therefore we can estimate

$$
\begin{aligned}
\left\|b_{\alpha \beta}^{\mathrm{w}}(x, D)\right\|_{L^{2} \rightarrow L^{2}} & \leq C\left\|\hat{b}_{\alpha \beta}\right\|_{L^{1}} \leq C\left\|\langle\xi\rangle^{2 n+1} \hat{b}_{\alpha \beta}\right\|_{L^{\infty}} \\
& \leq C \sup _{|\gamma| \leq 2 n+1}\left\|\widehat{D^{\gamma} b_{\alpha \beta}}\right\|_{L^{\infty}} \\
& \leq C \sup _{|\gamma| \leq 2 n+1}\left\|D^{\gamma} b_{\alpha \beta}\right\|_{L^{1}} \\
& \leq C \sup _{|\gamma| \leq 2 n+1}\left\|\langle z\rangle^{2 n+1} D^{\gamma} b_{\alpha \beta}\right\|_{L^{1}} \\
& \leq C\langle\alpha-\beta\rangle^{-N}
\end{aligned}
$$

according to (4.43), (4.44).
THEOREM 4.16 (Boundedness on $\boldsymbol{L}^{\mathbf{2}}$ ). If the symbol a belongs to $S_{\delta}$ for some $0 \leq \delta \leq 1 / 2$, then

$$
a^{\mathrm{w}}(x, D): L^{2}\left(\mathbb{R}^{n}\right) \rightarrow L^{2}\left(\mathbb{R}^{n}\right)
$$

is bounded, with the estimate

$$
\begin{equation*}
\left\|a^{\mathrm{w}}(x, D)\right\|_{L^{2} \rightarrow L^{2}} \leq C \sum_{|\alpha| \leq 2 n+1} \sup _{\mathbb{R}^{n}}\left|\partial^{\alpha} a\right| . \tag{4.45}
\end{equation*}
$$

Furthermore,

$$
\begin{equation*}
\left\|a^{\mathrm{w}}(x, h D) u\right\|_{L^{2} \rightarrow L^{2}} \leq C \sum_{|\alpha| \leq 2 n+1} h^{|\alpha|\left(\frac{1}{2}-\delta\right)} \sup _{\mathbb{R}^{n}}\left|\partial^{\alpha} a\right| . \tag{4.46}
\end{equation*}
$$

Proof. 1. We have $b_{\alpha \beta}^{\mathrm{w}}(x, D)=A_{\alpha}^{*} A_{\beta}$, where $A_{\alpha}:=a_{\alpha}^{\mathrm{w}}(x, D)$. Thus Theorem 4.15,(ii) asserts

$$
\left\|A_{\alpha}^{*} A_{\beta}\right\|_{L^{2} \rightarrow L^{2}} \leq C\langle\alpha-\beta\rangle^{-N}
$$

Therefore

$$
\sup _{\alpha} \sum_{\beta}\left\|A_{\alpha} A_{\beta}^{*}\right\|^{1 / 2} \leq C \sum_{\beta}\langle\alpha-\beta\rangle^{-N / 2} \leq C
$$

and similarly

$$
\sup _{\alpha} \sum_{\beta}\left\|A_{\alpha}^{*} A_{\beta}\right\|^{1 / 2} \leq C .
$$

Since

$$
a^{\mathrm{w}}(x, D)=\sum_{\alpha} A_{\alpha}
$$

we can apply the Cotlar-Stein Theorem C.5.
2. Estimate (4.46) follows from a rescaling, the details of which for $\delta=0$ we will later provide in the proof of Theorem 5.1.

As a first application, we record the useful
THEOREM 4.17 (Composition and multiplication). Suppose that $a, b \in S_{\delta}$ for $0 \leq \delta<\frac{1}{2}$.

Then

$$
\begin{equation*}
\left\|a^{\mathrm{w}}(x, h D) b^{\mathrm{w}}(x, h D)-(a b)^{\mathrm{w}}(x, h D)\right\|_{L^{2} \rightarrow L^{2}}=O\left(h^{1-2 \delta}\right) \tag{4.47}
\end{equation*}
$$

as $h \rightarrow 0$.

Proof. 1. In light of (4.38), we have

$$
a \# b-a b \in S_{\delta}^{2 \delta-1}
$$

Hence Theorem 4.16 implies

$$
a^{\mathrm{w}} b^{\mathrm{w}}-(a b)^{\mathrm{w}}=(a \# b-a b)^{\mathrm{w}}=O\left(h^{1-2 \delta}\right) .
$$

For the borderline case $\delta=\frac{1}{2}$, we have this assertion:

THEOREM 4.18 (Disjoint supports). Suppose that $a, b \in S_{\frac{1}{2}}$, and and

$$
\operatorname{dist}(\operatorname{spt}(a), \operatorname{spt}(b)) \geq \gamma>0
$$

for some constant $\gamma$. Assume also that $\operatorname{spt}(a) \subset K$ where the compact set $K$ and the constant $\gamma$ are independent of $h$. Then

$$
\begin{equation*}
\left\|a^{\mathrm{w}}(x, h D) b^{\mathrm{w}}(x, h D)\right\|_{L^{2} \rightarrow L^{2}}=O\left(h^{\infty}\right) \tag{4.48}
\end{equation*}
$$

Proof. Remember from (4.16) that

$$
a \# b(z)=\frac{1}{(h \pi)^{2 n}} \int_{\mathbb{R}^{2 n}} \int_{\mathbb{R}^{2 n}} e^{\frac{i}{h} \varphi\left(w_{1}, w_{2}\right)} a\left(z-w_{1}\right) b\left(z-w_{2}\right) d w_{1} d w_{2}
$$

for for $z=(x, \xi)$ and $\varphi\left(w_{1}, w_{2}\right)=-2 \sigma\left(w_{1}, w_{2}\right)$.
We proceed as in the proof of Theorem 4.15: $|\partial \varphi|=2|w|$ and thus the operator

$$
L:=\frac{\langle\partial \varphi, D\rangle}{|\partial \varphi|^{2}}
$$

has smooth coefficients on the support of $a\left(z-w_{1}\right) b\left(z-w_{2}\right)$. From our assumption that $a, b \in S_{\frac{1}{2}}$, we see that

$$
\left(L^{*}\right)^{M}\left(a\left(z-w_{1}\right) b\left(z-w_{2}\right)\right)=O\left(h^{\frac{M}{2}}\langle w\rangle^{-M}\right) .
$$

The uniform bound on the support shows that $a \# b \in S^{-\infty}$. Its Weyl quantization is therefore bounded on $L^{2}$, with norm of order $O\left(h^{\infty}\right)$.

### 4.5 INVERSES

At this stage we have constructed in appropriate generality the quantizations $a^{\mathrm{w}}(x, h D)$ of various symbols $a$. We turn therefore to the practical problem of understanding how the algebraic and analytic behavior of the function $a$ dictates properties of the corresponding quantized operators.

In this section we suppose that $a: \mathbb{R}^{2 n} \rightarrow \mathbb{C}$ is nonvanishing; so that the function $a$ is pointwise invertible. Can we draw the same conclusion about $a^{\mathrm{w}}(x, h D)$ ?

DEFINITION. We say the symbol $a$ is elliptic if there exists a constant $\gamma>0$ such that

$$
|a| \geq \gamma>0 \quad \text { on } \mathbb{R}^{2 n}
$$

THEOREM 4.19 (Inverses for elliptic symbols). Assume that $a \in S_{\delta}$ for $0 \leq \delta<\frac{1}{2}$ and that $a$ is elliptic.

Then for some constant $h_{0}>0$,

$$
a^{\mathrm{w}}(x, h D)^{-1}
$$

exists as a bounded linear operator on $L^{2}\left(\mathbb{R}^{n}\right)$, provided $0<h \leq h_{0}$.

Proof. Let $b:=\frac{1}{a}, b \in S_{\delta}$. Then (4.38) gives

$$
a \# b=1+r_{1}, \text { with } r_{1} \in S_{\delta}^{2 \delta-1}
$$

Likewise

$$
b \# a=1+r_{2}, \text { with } r_{2} \in S_{\delta}^{2 \delta-1}
$$

Hence if $A:=a^{\mathrm{w}}(x, h D), B:=b^{\mathrm{w}}(x, h D), R_{1}:=r_{1}^{\mathrm{w}}(x, h D)$ and $R_{2}:=$ $r_{2}^{\mathrm{w}}(x, h D)$, we have

$$
\begin{aligned}
& A B=I+R_{1} \\
& B A=I+R_{2},
\end{aligned}
$$

with

$$
\left\|R_{1}\right\|_{L^{2} \rightarrow L^{2}},\left\|R_{2}\right\|_{L^{2} \rightarrow L^{2}}=O\left(h^{1-2 \delta}\right) \leq \frac{1}{2}
$$

if $0<h \leq h_{0}$.
Thus $A=a^{\mathrm{w}}(x, h D)$ has an approximate left inverse and an approximate right inverse. Applying then Theorem C.3, we deduce that $A^{-1}$ exists.

### 4.6 GÅRDING INEQUALITIES

We continue studying how properties of the symbol $a$ translate into properties of the corresponding quantized operators. In this section we suppose that $a$ is real-valued and nonnegative, and ask the consequences for $a^{\mathrm{w}}(x, h D)$.

THEOREM 4.20 (Easy Gårding inequality). Assume $a$ is a realvalued symbol in $S$ and

$$
\begin{equation*}
a \geq \gamma>0 \quad \text { on } \quad \mathbb{R}^{2 n} \tag{4.49}
\end{equation*}
$$

Then for each $\epsilon>0$ there exists $h_{0}=h_{0}(\epsilon)>0$ such that

$$
\begin{equation*}
\left\langle a^{\mathrm{w}}(x, h D) u, u\right\rangle \geq(\gamma-\epsilon)\|u\|_{L^{2}}^{2} \tag{4.50}
\end{equation*}
$$

for all $0<h \leq h_{0}, u \in L^{2}\left(\mathbb{R}^{n}\right)$.

Proof. We will show that

$$
\begin{equation*}
(a-\lambda)^{-1} \in S \quad \text { if } \lambda<\gamma-\epsilon \tag{4.51}
\end{equation*}
$$

Indeed if $b:=(a-\lambda)^{-1}$, then

$$
(a-\lambda) \# b=1+\frac{h}{2 i}\{a-\lambda, b\}+O\left(h^{2}\right)=1+O\left(h^{2}\right),
$$

the bracket term vanishing since $b$ is a function of $a-\lambda$. Therefore

$$
\left(a^{\mathrm{w}}(x, h D)-\lambda\right) b^{\mathrm{w}}(x, h D)=I+O\left(h^{2}\right)_{L^{2} \rightarrow L^{2}},
$$

and so $b^{\mathrm{w}}(x, h D)$ is an approximate right inverse of $a^{\mathrm{w}}(x, h D)-\lambda$. Likewise $b^{\mathrm{w}}(x, h D)$ is an approximate left inverse.

Hence Theorem C. 3 implies $a^{\mathrm{w}}(x, h D)-\lambda$ is invertible for each $\lambda<$ $\gamma-\epsilon$. Consequently,

$$
\operatorname{spec}\left(a^{\mathrm{w}}(x, h D)\right) \subset[\gamma-\epsilon, \infty)
$$

According then to Theorem C.6,

$$
\left\langle a^{\mathrm{w}}(x, h D) u, u\right\rangle \geq(\gamma-\epsilon)\|u\|_{L^{2}}^{2}
$$

for all $u \in L^{2}$.

We next improve the preceding estimate:
THEOREM 4.21 (Sharp Gårding inequality). Let $a=a(x, \xi)$ be a symbol in $S$ and suppose that

$$
\begin{equation*}
a \geq 0 \quad \text { on } \quad \mathbb{R}^{2 n} \tag{4.52}
\end{equation*}
$$

Then there exist constants $h_{0}>0, C \geq 0$ such that

$$
\begin{equation*}
\left\langle a^{\mathrm{w}}(x, h D) u, u\right\rangle \geq-C h\|u\|_{L^{2}}^{2} \tag{4.53}
\end{equation*}
$$

for all $0<h \leq h_{0}$ and $u \in L^{2}\left(\mathbb{R}^{n}\right)$.

REMARK. The estimate (4.53) is in fact true for each quantization $\mathrm{Op}_{t}(a)(0 \leq t \leq 1)$. And for the Weyl quantization, the stronger Fefferman-Phong inequality holds:

$$
\left\langle a^{\mathrm{w}}(x, h D) u, u\right\rangle \geq-C h^{2}\|u\|_{L^{2}}^{2}
$$

for $0<h \leq h_{0}, u \in L^{2}$.
We will need

LEMMA 4.22 (Gradient estimate). Let $f: \mathbb{R}^{n} \rightarrow \mathbb{R}$ be $C^{2}$, with

$$
\left|\partial^{2} f\right| \leq A
$$

Suppose also $f \geq 0$. Then

$$
|\partial f| \leq(2 A f)^{1 / 2}
$$

Proof. By Taylor's Theorem,

$$
f(x+y)=f(x)+\langle\partial f(x), y\rangle+\int_{0}^{1}(1-t)\left\langle\partial^{2} f(x+t y) y, y\right\rangle d t .
$$

Let $y=-\lambda \partial f(x), \lambda>0$ to be selected. Then since $f \geq 0$, we have

$$
\begin{aligned}
\lambda|\partial f(x)|^{2} & \leq f(x)+\lambda^{2} \int_{0}^{1}(1-t)\left\langle\partial^{2} f(x-\lambda t \partial f(x)) \partial f(x), \partial f(x)\right\rangle d t \\
& \leq f(x)+\frac{\lambda^{2}}{2} A|\partial f(x)|^{2}
\end{aligned}
$$

Put $\lambda=\frac{1}{A}$, to conclude $|\partial f(x)|^{2} \leq 2 A f(x)$.
Proof of Theorem 4.21 1. The primary goal is to show that if

$$
\begin{equation*}
\lambda=\mu h \tag{4.54}
\end{equation*}
$$

and $\mu$ is fixed sufficiently large, then

$$
\begin{equation*}
h(a+\lambda)^{-1} \in S_{1 / 2}\left(\frac{1}{\mu}\right) \tag{4.55}
\end{equation*}
$$

with estimates independent of $\mu$.
To begin the proof of (4.55) we consider for any multiindex $\alpha=$ $\left(\alpha_{1}, \ldots, \alpha_{2 n}\right)$ the partial derivative $\partial^{\alpha}$ in the variables $x$ and $\xi$.

We claim that $\partial^{\alpha}(a+\lambda)^{-1}$ has the form

$$
\begin{align*}
& \partial^{\alpha}(a+\lambda)^{-1}=  \tag{4.56}\\
& \quad(a+\lambda)^{-1} \sum_{k=1}^{|\alpha|} \sum_{\substack{\alpha=\beta^{1}+\ldots+\beta^{k} \\
\left|\beta^{j}\right| \geq 1}} C_{\beta^{1}, \ldots, \beta^{k}} \prod_{j=1}^{k}\left((a+\lambda)^{-1} \partial^{\beta^{j}} a\right),
\end{align*}
$$

for appropriate constants $C_{\beta^{1}, \ldots, \beta^{k}}$. To see this, observe that when we compute $\partial^{\alpha}(a+\lambda)^{-1}$ a typical term involves $k$ differentiations of $(a+\lambda)^{-1}$ with the remaining derivatives falling on $a$. In obtaining (4.56) we for each $k \leq|\alpha|$ partition $\alpha$ into multiindices $\beta^{1}, \ldots, \beta^{k}$, each of which corresponds to one derivative falling on $(a+\lambda)^{-1}$ and the remaining derivatives falling on $a$.
2. Now Lemma 4.22 implies for $\left|\beta^{j}\right|=1$ that

$$
\begin{equation*}
\left|\partial^{\beta_{j}} a\right|(a+\lambda)^{-1} \leq C \lambda^{-1 / 2} \tag{4.57}
\end{equation*}
$$

since $\lambda^{1 / 2}|\partial a| \leq C \lambda^{1 / 2} a^{1 / 2} \leq C(\lambda+a)$. Furthermore,

$$
\begin{equation*}
\left|\partial^{\beta_{j}} a\right|(a+\lambda)^{-1} \leq C \lambda^{-1} \tag{4.58}
\end{equation*}
$$

if $\left|\beta_{j}\right| \geq 2$, since $a \in S$.
Consequently, for each partition $\alpha=\beta^{1}+\cdots+\beta^{k}$ and $0<\lambda \leq 1$ :

$$
\left|\prod_{j=1}^{k}(a+\lambda)^{-1} \partial^{\beta_{j}} a\right| \leq C \prod_{\left|\beta_{j}\right| \geq 2} \lambda^{-1} \prod_{\left|\beta_{j}\right|=1} \lambda^{-1 / 2} \leq C \prod_{j=1}^{k} \lambda^{-\frac{\left|\beta_{j}\right|}{2}}=C \lambda^{-\frac{|\alpha|}{2}} .
$$

Therefore

$$
\begin{equation*}
\left|\partial^{\alpha}(a+\lambda)^{-1}\right| \leq C_{\alpha}(a+\lambda)^{-1} \lambda^{-\frac{|\alpha|}{2}} \tag{4.59}
\end{equation*}
$$

But since $\lambda=\mu h$, this implies

$$
(a+\lambda)^{-1} \in S_{1 / 2}\left(\frac{1}{\mu h}\right)
$$

that is,

$$
h(a+\lambda)^{-1} \in S_{1 / 2}\left(\frac{1}{\mu}\right)
$$

with estimates independent of $\mu$.
3. Since $a+\lambda \in S \subseteq S_{\frac{1}{2}}$, we can define $(a+\lambda) \# b$, for $b=(a+\lambda)^{-1}$. Using Taylor's formula, we compute

$$
\begin{aligned}
& (a+\lambda) \# b(z) \\
& =\left.e^{\frac{i h}{2} \sigma\left(D_{z}, D_{w}\right)}(a(z)+\lambda) b(w)\right|_{w=z} \\
& =1+\left.\int_{0}^{1}(1-t) e^{\frac{i t h}{2} \sigma\left(D_{z}, D_{w}\right)}\left(\frac{i h}{2} \sigma\left(D_{z}, D_{w}\right)\right)^{2}(a(z)+\lambda) b(w)\right|_{w=z} d t \\
& =: 1+r(z)
\end{aligned}
$$

where we have noted that $\left\{a+\lambda,(a+\lambda)^{-1}\right\}=0$.
Now according to (4.59), $h b \in S_{1 / 2}(1 / \mu)$ and so $h^{2} \partial^{\alpha} b \in S_{1 / 2}(1 / \mu)$ for $|\alpha|=2$. An application of $e^{\frac{i h}{2} \sigma\left(D_{z}, D_{w}\right)}$ preserves the symbol class $S_{1 / 2}(1 / \mu)$. Consequently,

$$
\left\|r^{\mathrm{w}}(x, h D)\right\|_{L^{2} \rightarrow L^{2}} \leq \frac{C}{\mu} \leq \frac{1}{2}
$$

if $\mu$ is now fixed large enough. Thus $b^{\mathrm{w}}(x, h D)$ is an approximate right inverse of $a^{\mathrm{w}}(x, h D)+\lambda$, and is similarly an approximate left inverse.

So $\left(a^{\mathrm{w}}(x, h D)+\lambda\right)^{-1}$ exists. Likewise $\left(a^{\mathrm{w}}(x, h D)+\gamma+\lambda\right)^{-1}$ exists for all $\gamma \geq 0$. Therefore

$$
\operatorname{spec}\left(a^{\mathrm{w}}(x, h D)\right) \subseteq[-\lambda, \infty) .
$$

According then to Theorem C.6,

$$
\left\langle a^{\mathrm{w}}(x, h D) u, u\right\rangle \geq-\lambda\|u\|_{L^{2}}^{2}
$$

for all $u \in L^{2}$. Since $\lambda=\mu h$, this inequality finishes the proof.

## 5. Semiclassical defect measures

5.1 Construction, examples
5.2 Defect measures and PDE
5.3 Application: damped wave equation

One way to understand limits as $h \rightarrow 0$ of a collection of functions $\{u(h)\}_{0<h \leq h_{0}}$ bounded in $L^{2}$ is to construct corresponding semiclassical defect measures $\mu$, which record the limiting behavior of certain quadratic forms acting on $u(h)$. If in addition these functions solve certain operator equations or PDE, we can deduce various properties of the measure $\mu$ and thereby indirectly recover information about asymptotoics as $h \rightarrow 0$ of the functions $u(h)$.

### 5.1 CONSTRUCTION, EXAMPLES

In the first two sections of this chapter, we consider a collection of functions $\{u(h)\}_{0<h \leq h_{0}}$ that is bounded in $L^{2}\left(\mathbb{R}^{n}\right)$ :

$$
\begin{equation*}
\sup _{0<h \leq h_{0}}\|u(h)\|_{L^{2}}<\infty . \tag{5.1}
\end{equation*}
$$

For the time being, we do not assume that $u(h)$ solves any PDE.

THEOREM 5.1 (An operator norm bound). Suppose $a \in S$. Then

$$
\begin{equation*}
\left\|a^{\mathrm{w}}(x, h D)\right\|_{L^{2} \rightarrow L^{2}} \leq C \sup _{\mathbb{R}^{2 n}}|a|+O\left(h^{\frac{1}{2}}\right) \tag{5.2}
\end{equation*}
$$

as $h \rightarrow 0$.

Proof. We showed earlier in Theorem 4.16 that if $a \in S$ and $h=1$, then

$$
\begin{equation*}
\left\|a^{\mathrm{w}}(x, D)\right\|_{L^{2} \rightarrow L^{2}} \leq C \sup _{|\alpha| \leq 2 n+1}\left|\partial^{\alpha} a\right| . \tag{5.3}
\end{equation*}
$$

Suppose now $a \in S$ and $u \in \mathcal{S}$. We rescale by taking

$$
\tilde{x}:=h^{-\frac{1}{2}} x, \tilde{y}:=h^{-\frac{1}{2}} y, \tilde{\xi}:=h^{-\frac{1}{2}} \xi
$$

and

$$
\tilde{u}(\tilde{x}):=h^{\frac{n}{4}} u(x)=h^{\frac{n}{4}} u\left(h^{\frac{1}{2}} \tilde{x}\right) .
$$

(This is a different rescaling of $u$ from that introduced earlier in (4.31), the advantage being that $u \mapsto \tilde{u}$ is now a unitary transformation of $L^{2}$ : $\|u\|_{L^{2}}=\|\tilde{u}\|_{L^{2}}$.)

Then

$$
\begin{align*}
a^{\mathrm{w}}(x, h D) & u(x) \\
& =\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} a\left(\frac{x+y}{2}, \xi\right) e^{\frac{i}{h}\langle x-y, \xi\rangle} u(y) d y d \xi \\
& =\frac{h^{-\frac{n}{4}}}{(2 \pi)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} a_{h}\left(\frac{\tilde{x}+\tilde{y}}{2}, \tilde{\xi}\right) e^{i\langle\tilde{x}-\tilde{y}, \tilde{\xi}\rangle} \tilde{u}(\tilde{y}) d \tilde{y} d \tilde{\xi}  \tag{5.4}\\
& =h^{-\frac{n}{4}} a_{h}^{\mathrm{w}}(\tilde{x}, D) \tilde{u}(\tilde{x}),
\end{align*}
$$

for

$$
a_{h}(\tilde{x}, \tilde{\xi}):=a(x, \xi)=a\left(h^{\frac{1}{2}} \tilde{x}, h^{\frac{1}{2}} \tilde{\xi}\right) .
$$

Hence, noting that $d x=h^{\frac{n}{2}} d \tilde{x}$, we deduce from (5.4) and (5.3) that

$$
\begin{aligned}
\left\|a^{\mathrm{w}}(x, h D) u\right\|_{L^{2}} & =\left\|a_{h}^{\mathrm{w}}(\tilde{x}, D) \tilde{u}\right\|_{L^{2}} \\
& \leq\left\|a_{h}^{\mathrm{w}}\right\|_{L^{2} \rightarrow L^{2}}\|\tilde{u}\|_{L^{2}} \\
& \leq C \sup _{|\alpha| \leq 2 n+1}\left|\partial^{\alpha} a_{h}\right|\|u\|_{L^{2}} \\
& \leq C \sup _{|\alpha| \leq 2 n+1} h^{\left.\frac{|\alpha|}{2} \right\rvert\,}\left|\partial^{\alpha} a\right|\|u\|_{L^{2}} .
\end{aligned}
$$

This implies (5.2).

THEOREM 5.2 (Existence of defect measure). There exists a Radon measure $\mu$ on $\mathbb{R}^{2 n}$ and a sequence $h_{j} \rightarrow 0$ such that

$$
\begin{equation*}
\left\langle a^{\mathrm{w}}\left(x, h_{j} D\right) u\left(h_{j}\right), u\left(h_{j}\right)\right\rangle \rightarrow \int_{\mathbb{R}^{2 n}} a(x, \xi) d \mu \tag{5.5}
\end{equation*}
$$

for each symbol $a \in C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{2 n}\right)$.

DEFINITION. We call $\mu$ a microlocal defect measure associated with the family $\{u(h)\}_{0<h \leq h_{0}}$.

Proof. 1. Choose $\left\{a_{k}\right\}_{k=0}^{\infty} \subset C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{2 n}\right)$ to be dense in $C_{\mathrm{c}}\left(\mathbb{R}^{2 n}\right)$. Select a sequence $h_{j}^{1} \rightarrow 0$ such that

$$
\left\langle a_{1}^{\mathrm{w}}\left(x, h_{j}^{1} D\right) u\left(h_{j}^{1}\right), u\left(h_{j}^{1}\right)\right\rangle \rightarrow \alpha_{1} .
$$

Choose next a further subsequence $\left\{h_{j}^{2}\right\} \subseteq\left\{h_{j}^{1}\right\}$ such that

$$
\left\langle a_{2}^{\mathrm{w}}\left(x, h_{j}^{2} D\right) u\left(h_{j}^{2}\right), u\left(h_{j}^{2}\right)\right\rangle \rightarrow \alpha_{2}
$$

Continue, at the $k^{t h}$ step extracting a subsequence $\left\{h_{j}^{k}\right\} \subseteq\left\{h_{j}^{k-1}\right\}$ such that

$$
\left\langle a_{k}^{\mathrm{w}}\left(x, h_{j}^{k} D\right) u\left(h_{j}^{k}\right), u\left(h_{j}^{k}\right)\right\rangle \rightarrow \alpha_{k} .
$$

By a standard diagonal argument, we see that the sequence $h_{j}:=h_{j}^{j}$ converges to 0 , with

$$
\left\langle a_{k}^{\mathrm{w}}\left(x, h_{j} D\right) u\left(h_{j}\right), u\left(h_{j}\right)\right\rangle \rightarrow \alpha_{k}
$$

for all $k=1, \ldots$.
2. Define $\Phi\left(a_{k}\right):=\alpha_{k}$. Owing to Theorem 5.1, we see for each $k$ that

$$
\begin{aligned}
\left|\Phi\left(a_{k}\right)\right|=\left|\alpha_{k}\right|=\lim _{h_{j} \rightarrow \infty} & \left|\left\langle a_{k}^{\mathrm{w}} u\left(h_{j}\right), u\left(h_{j}\right)\right\rangle\right| \\
& \leq C \limsup _{h_{j} \rightarrow \infty}\left\|a_{k}^{\mathrm{w}}\right\|_{L^{2} \rightarrow L^{2}} \leq C \sup _{\mathbb{R}^{2 n}}\left|a_{k}\right| .
\end{aligned}
$$

The mapping $\Phi$ is bounded, linear and densely defined, and therefore uniquely extends to a bounded linear functional on $C_{\mathrm{c}}\left(\mathbb{R}^{2 n}\right)$, with the estimate

$$
|\Phi(a)| \leq C \sup _{\mathbb{R}^{2 n}}|a|
$$

for all $a \in C_{c}\left(\mathbb{R}^{2 n}\right)$. The Riesz Representation Theorem therefore implies the existence of a (possibly complex-valued) Radon measure on $\mathbb{R}^{2 n}$ such that

$$
\Phi(a)=\int_{\mathbb{R}^{2 n}} a(x, \xi) d \mu
$$

REMARK. Theorem 5.2 is also valid if we replace the Weyl quantization $a^{\mathrm{w}}=\mathrm{Op}_{1 / 2}(a)$ by $\mathrm{Op}_{t}(a)$ for any $0 \leq t \leq 1$, since the error is then $O(h)$.

THEOREM 5.3 (Positivity). The measure $\mu$ is real and nonnegative:

$$
\begin{equation*}
\mu \geq 0 \tag{5.6}
\end{equation*}
$$

Proof. We must show that $a \geq 0$ implies

$$
\int_{\mathbb{R}^{2 n}} a d \mu \geq 0
$$

Now when $a \geq 0$, the sharp Gårding inequality, Theorem 4.21, implies

$$
a^{\mathrm{w}}(x, h D) \geq-C h ;
$$

that is,

$$
\left\langle a^{\mathrm{w}}(x, h D) u(h), u(h)\right\rangle \geq-C h\|u(h)\|_{L^{2}}^{2}
$$

for sufficiently small $h>0$. Let $h=h_{j} \rightarrow 0$, to deduce

$$
\int_{\mathbb{R}^{2 n}} a d \mu=\lim _{h_{j} \rightarrow \infty}\left\langle a^{\mathrm{w}}\left(x, h_{j} D\right) u\left(h_{j}\right), u\left(h_{j}\right)\right\rangle \geq 0
$$

EXAMPLE 1: Coherent states. Fix a point $\left(x_{0}, \xi_{0}\right)$ and define the corresponding coherent state

$$
u(h)(x):=(\pi h)^{-\frac{n}{4}} e^{\frac{i}{h}\left\langle x-x_{0}, \xi_{0}\right\rangle-\frac{1}{2 h}\left|x-x_{0}\right|^{2}},
$$

where we have normalized so that $\|u(h)\|_{L^{2}}=1$. Then there exists precisely one associated semiclassical defect measure, namely

$$
\mu:=\delta_{\left(x_{0}, \xi_{0}\right)}
$$

To confirm this statement, take $t=1$ in the quantization and calculate

$$
\begin{aligned}
& \langle a(x, h D) u(h), u(h)\rangle \\
& \quad=\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} a(x, \xi) e^{\frac{i}{h}\langle x-y, \xi\rangle} u(h)(y) \overline{u(h)}(x) d y d \xi d x \\
& =\frac{2^{\frac{n}{2}}}{(2 \pi h)^{\frac{3 n}{2}}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} a(x, \xi) e^{\frac{i}{h}\left(\langle x-y, \xi\rangle+\left\langle y-x_{0}, \xi_{0}\right\rangle-\left\langle x-x_{0}, \xi_{0}\right\rangle\right)} \\
& =\frac{2^{\frac{n}{2}}}{(2 \pi h)^{\frac{3 n}{2}}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} a(x, \xi) e^{-\frac{1}{2 h}\left(\left|y-x_{0}\right|^{2}+\left|x-x_{0}\right|^{2}\right)} d y d \xi d x \\
& e^{-\frac{1}{2 h}\left(\left|y-y, \xi-x_{0}\right|^{2}+\left|x-x_{0}\right|^{2}\right)} d y d \xi d x .
\end{aligned}
$$

For each fixed $x$ and $\xi$, the integral in $y$ is

$$
\begin{aligned}
\int_{\mathbb{R}^{n}} e^{\frac{i}{h}\left\langle x-y, \xi-\xi_{0}\right\rangle} e^{-\frac{1}{2 h}\left|y-x_{0}\right|^{2}} d y & =e^{\frac{i}{h}\left\langle x-x_{0}, \xi-\xi_{0}\right\rangle} \int_{\mathbb{R}^{n}} e^{-\frac{i}{h}\left\langle y, \xi-\xi_{0}\right\rangle} e^{-\frac{1}{2 h}|y|^{2}} d y \\
& =e^{\frac{i}{h}\left\langle x-x_{0}, \xi-\xi_{0}\right\rangle} \mathcal{F}\left(e^{-\frac{1}{2 h}|y|^{2}}\right)\left(\frac{\xi-\xi_{0}}{h}\right) \\
& =(2 \pi h)^{\frac{n}{2}} e^{\frac{i}{h}\left\langle x-x_{0}, \xi-\xi_{0}\right\rangle} e^{-\frac{1}{2 h}\left|\xi-\xi_{0}\right|^{2}}
\end{aligned}
$$

where we used formula (3.2) for the last equality. Therefore

$$
\begin{aligned}
& \langle a(x, h D) u(h), u(h)\rangle \\
& \quad=\frac{2^{\frac{n}{2}}}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} a(x, \xi) e^{\frac{i}{h}\left\langle x-x_{0}, \xi-\xi_{0}\right\rangle} e^{-\frac{1}{2 h}\left(\left|x-x_{0}\right|^{2}+\left|\xi-\xi_{0}\right|^{2}\right)} d x d \xi \\
& \quad=a\left(x_{0}, \xi_{0}\right) \frac{2^{\frac{n}{2}}}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}\left\langle x-x_{0}, \xi-\xi_{0}\right\rangle} e^{-\frac{1}{2 h}\left(\left|x-x_{0}\right|^{2}+\left|\xi-\xi_{0}\right|^{2}\right)} d x d \xi+o(1) \\
& \quad=C a\left(x_{0}, \xi_{0}\right)+o(1),
\end{aligned}
$$

for the constant

$$
C:=\frac{2^{\frac{n}{2}}}{(2 \pi)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{i\langle x, \xi\rangle} e^{-\frac{1}{2}\left(|x|^{2}+|\xi|^{2}\right)} d x d \xi
$$

Taking $a \equiv 1$ and recalling that $\|u(h)\|_{L^{2}}=1$, we deduce that $C=$ 1.

EXAMPLE 2: Stationary phase and defect measures. For our next example, take

$$
u(h)(x):=e^{\frac{i \varphi(x)}{h}} b(x)
$$

where $\varphi, b \in C^{\infty}$ and $\|b\|_{L^{2}}=1$. Then

$$
\begin{aligned}
& \langle a(x, h D) u(h), u(h)\rangle= \\
& \quad \frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} a(x, \xi) e^{\frac{i}{h}(\langle x-y, \xi\rangle+\varphi(y)-\varphi(x))} b(y) \overline{b(x)} d y d \xi d x .
\end{aligned}
$$

We assume $a \in C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{2 n}\right)$ and apply stationary phase. For a given value of $x$, define

$$
\phi(y, \xi):=\langle x-y, \xi\rangle+\varphi(y)-\varphi(x)
$$

Then

$$
\partial_{y} \phi=\partial \varphi(y)-\xi, \partial_{\xi} \phi=x-y
$$

and the Hessian matrix of $\phi$ is

$$
\partial^{2} \phi=\left(\begin{array}{cc}
\partial^{2} \varphi & -I \\
-I & O
\end{array}\right)
$$

The signature of a matrix is integer valued, and consequently is invariant if we move along a curve of nonsingular matrices. Since

$$
\operatorname{sgn}\left(\begin{array}{cc}
O & -I \\
-I & O
\end{array}\right)=0
$$

it follows that

$$
\operatorname{sgn}\left(\begin{array}{cc}
t \partial^{2} \varphi & -I \\
-I & O
\end{array}\right)=0
$$

for $0 \leq t \leq 1$; and therefore

$$
\operatorname{sgn}\left(\partial^{2} \phi\right)=0
$$

In addition, $\left|\operatorname{det} \partial^{2} \phi\right|=1$. Thus as $h \rightarrow 0$ the stationary phase asymptotic expression (3.43) implies

$$
\left\langle a^{\mathrm{w}}(x, h D) u(h), u(h)\right\rangle \rightarrow \int_{\mathbb{R}^{n}} a(x, \partial \varphi(x))|b(x)|^{2} d x=\int_{\mathbb{R}^{2 n}} a(x, \xi) d \mu
$$

for the semiclassical defect measure

$$
\mu:=|b(x)|^{2} \delta_{\{\xi=\partial \varphi(x)\}} \mathcal{L}^{n},
$$

$\mathcal{L}^{n}$ denoting $n$-dimensional Lebesgue measure in the $x$-variables.

### 5.2 DEFECT MEASURES AND PDE

We now assume more about the family $\{u(h)\}_{0<h \leq h_{0}}$, namely that each function $u(h)$ is an approximate solution of a equation involving the operator $P(h)=p^{\mathrm{w}}(x, h D)$ for some real symbol $p \in S\left(\langle\xi\rangle^{m}\right)$ satisfying

$$
\begin{equation*}
|p| \geq \gamma\langle\xi\rangle^{m} \quad \text { if } \quad|\xi| \geq C \tag{5.7}
\end{equation*}
$$

for constants $C, \gamma>0$.
First, let us suppose $P(h) u(h)$ vanishes up to an $o(1)$ error term and see what we can conclude about a corresponding semiclassical defect measure $\mu$.

THEOREM 5.4 (Support of defect measure). Suppose that $u(h)$

$$
\left\{\begin{array}{l}
\|P(h) u(h)\|_{L^{2}}=o(1) \quad \text { as } h \rightarrow 0,  \tag{5.8}\\
\|u(h)\|_{L^{2}}=1 .
\end{array}\right.
$$

Then if $\mu$ is any microlocal defect measure associated with $\{u(h)\}_{0<h \leq h_{0}}$,

$$
\begin{equation*}
\operatorname{spt} \mu \subseteq p^{-1}(0) \tag{5.9}
\end{equation*}
$$

Interpretation. We sometimes call $p^{-1}(0)$ the characteristic variety or zero energy surface of the symbol $p$. We understand (5.9) as saying that in the semiclassical limit $h \rightarrow 0$, all of the "mass" of the approximate solution $u(h)$ coalesces in phase space onto this set.

Proof. Select $a \in C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{2 n}\right)$ such that $\operatorname{spt}(a) \cap p^{-1}(0)=0$. We must show

$$
\int_{\mathbb{R}^{2 n}} a d \mu=0
$$

To do so, first select $q \in S\left(\langle\xi\rangle^{m}\right)$ such that $\operatorname{spt}(a) \cap \operatorname{spt}(q)=\emptyset$ and

$$
|p+i q| \geq \delta\langle\xi\rangle^{m}>0 \quad \text { on } \mathbb{R}^{2 n}
$$

for some $\delta>0$. We can for instance choose a function $q \in C^{\infty}$ that is equal to one on $p^{-1}(0)$, and then modify it near the compact support of $a$.

Write $Q(h):=q^{\mathrm{w}}(x, h D)$. Then Theorem 4.19 ensures us that for small enough $h$ the operator $\langle h D\rangle^{-m}(P(h)+i Q(h))$ is invertible on $L^{2}$.

Next, put $A(h):=a^{\mathrm{w}}(x, h D)$. We observe that

$$
\frac{a p}{p+i q}-a=-i \frac{a q}{p+i q}
$$

Since $a$ and $q$ have disjoint support, Theorems 4.17 and 4.18 imply

$$
\left\|A(h)(P(h)+i Q(h))^{-1} P(h)-A(h)\right\|_{L^{2} \rightarrow L^{2}}=O(h) .
$$

Therefore (5.8) implies

$$
\|A(h) u(h)\|_{L^{2}}=o(1)
$$

and thus

$$
\langle A(h) u(h), u(h)\rangle \rightarrow 0 .
$$

But also

$$
\left\langle A\left(h_{j}\right) u\left(h_{j}\right), u\left(h_{j}\right)\right\rangle=\left\langle a^{\mathrm{w}}\left(x, h_{j} D\right) u\left(h_{j}\right), u\left(h_{j}\right)\right\rangle \rightarrow \int_{\mathbb{R}^{2 n}} a d \mu .
$$

Now we make the stronger assumption that the error term in (5.8) is $o(h)$.

THEOREM 5.5 (Flow invariance). Assume

$$
\left\{\begin{array}{l}
\|P(h) u(h)\|_{L^{2}}=o(h) \quad \text { as } h \rightarrow 0  \tag{5.10}\\
\|u(h)\|_{L^{2}}=1
\end{array}\right.
$$

Then

$$
\begin{equation*}
\int_{\mathbb{R}^{2 n}}\{p, a\} d \mu=0 \tag{5.11}
\end{equation*}
$$

for all $a \in C_{c}^{\infty}\left(\mathbb{R}^{2 n}\right)$.

Interpretation. Let $\varphi_{t}$ be the flow generated by the Hamiltonian vector field $H_{p}$. Then

$$
\frac{d}{d t} \int_{\mathbb{R}^{2 n}} \varphi_{t}^{*} a d \mu=\int_{\mathbb{R}^{2 n}}\left(H_{p} a\right)\left(\varphi_{t}\right) d \mu=\int_{\mathbb{R}^{2 n}}\{p, a\} d \mu .
$$

Conseqently (5.11) asserts that the semiclassical defect measure $\mu$ is flow-invariant.

The proof below illustrates one of the basic principles mentioned in Chapter 1, that an assertion about Hamiltonian dynamics involving the Poisson bracket corresponds to a commutator argument at the quantum level.

Proof. Since $p$ is real, $P(h)=p^{\mathrm{w}}(x, h D)$ is self-adjoint on $L^{2}$. Select $a$ as above and write $A(h)=a^{\mathrm{w}}(x, h D)$. Recall that $A(h)=A(h)^{*}$. Then

$$
\begin{aligned}
\langle[P(h), A(h)] u(h), u(h)\rangle= & \langle(P(h) A(h)-A(h) P(h)) u(h), u(h)\rangle \\
& =\langle A(h) u(h), P(h) u(h)\rangle \\
& \quad-\langle P(h) u(h), A(h) u(h)\rangle \\
& =o(h) \quad
\end{aligned}
$$

as $h \rightarrow 0$. On the other hand,

$$
[P(h), A(h)]=\frac{h}{i}\{p, a\}^{\mathrm{w}}(x, h D)+O\left(h^{2}\right)_{L^{2} \rightarrow L^{2}} .
$$

Hence

$$
\langle[P(h), A(h)] u(h), u(h)\rangle=\frac{h}{i}\left\langle\{p, a\}^{\mathrm{w}} u(h), u(h)\right\rangle+o(h) .
$$

Divide by $h>0$ and let $h=h_{j} \rightarrow 0$ :

$$
\int_{\mathbb{R}^{2 n}}\{p, a\} d \mu=0
$$

Note that even though $p$ may not have compact support, $\{p, a\}$ does.

We have similar statements if we replace $\mathbb{R}^{n} \times \mathbb{R}^{n}$ by the torus $\mathbb{T}^{n} \times \mathbb{R}^{n}$. We will need this observation for the following application.

### 5.3 APPLICATION: DAMPED WAVE EQUATION

A damped wave equation. In this section $\mathbb{T}^{n}$ denotes the flat $\mathrm{n}-$ dimensional torus. We consider now the initial-value problem

$$
\left\{\begin{align*}
\left(\partial_{t}^{2}+a \partial_{t}-\Delta\right) u=0 & \text { on } \mathbb{T}^{n} \times\{t>0\}  \tag{5.12}\\
u=0, u_{t}=f & \text { on } \mathbb{T}^{n} \times\{t=0\}
\end{align*}\right.
$$

in which the smooth function $a=a(x)$ is nonnegative, and thus represents a damping mechanism, as we will see.
DEFINITION. The energy at time $t$ is

$$
E(t):=\frac{1}{2} \int_{\mathbb{T}^{n}}\left(\partial_{t} u\right)^{2}+\left|\partial_{x} u\right|^{2} d x .
$$

## LEMMA 5.6 (Elementary energy estimates).

(i) If $a \equiv 0, t \mapsto E(t)$ is constant.
(ii) If $a \geq 0, t \mapsto E(t)$ is nonincreasing.

Proof. These assertions follow easily from the calculation

$$
\begin{aligned}
E^{\prime}(t)=\int_{\mathbb{T}^{n}} \partial_{t} u \partial_{t}^{2} u & +\left\langle\partial_{x} u, \partial_{x t}^{2} u\right\rangle d x \\
& =\int_{\mathbb{T}^{n}} \partial_{t} u\left(\partial_{t}^{2} u-\Delta u\right) d x=-\int_{\mathbb{T}^{n}} a\left(\partial_{t} u\right)^{2} d x \leq 0
\end{aligned}
$$

Our eventual goal is showing that if the support of the damping term $a$ is large enough, then we have exponential energy decay for our solution of the wave equation (5.12). Here is the key assumption:

## DYNAMICAL HYPOTHESIS.

$$
\left\{\begin{array}{l}
\text { There exists a time } T>0 \text { such that any }  \tag{5.13}\\
\text { trajectory of the Hamiltonian vector field of } \\
p(x, \xi)=|\xi|^{2}, \text { starting at time } 0 \text { with }|\xi|=1, \\
\text { intersects the set }\{a>0\} \text { by the time } T
\end{array}\right.
$$

Equivalently, for each initial point $z=(x, \xi) \in \mathbb{T}^{n} \times \mathbb{R}^{n}$, with $|\xi|=1$, we have

$$
\langle a\rangle_{T}:=\frac{1}{T} \int_{0}^{T} a(x+t \xi) d t>0
$$

Motivation. Since the damping term $a$ in general depends upon $x$, we cannot use Fourier transform (or Fourier series) in $x$ to solve (5.12). Instead we define $u \equiv 0$ for $t<0$ and take the Fourier transform in $t$ :

$$
\hat{u}(x, \tau):=\int_{0}^{\infty} e^{-i t \tau} u(x, t) d t \quad(\tau \in \mathbb{R})
$$

Then

$$
\begin{aligned}
\Delta \hat{u} & =\int_{0}^{\infty} e^{-i t \tau} \Delta u d t=\int_{0}^{\infty} e^{-i t \tau}\left(\partial_{t}^{2} u+a \partial_{t} u\right) d t \\
& =\int_{0}^{\infty}\left((i \tau)^{2}+a i \tau\right) e^{-i t \tau} u d t-f=\left(-\tau^{2}+a i \tau\right) \hat{u}-f
\end{aligned}
$$

Consequently,

$$
\begin{equation*}
P(\tau) \hat{u}:=\left(-\Delta+i \tau a-\tau^{2}\right) \hat{u}=f . \tag{5.14}
\end{equation*}
$$

Now take $\tau$ to be complex, with $\operatorname{Re} \tau \geq 0$, and define

$$
\begin{equation*}
P(z, h):=-h^{2} \Delta+i \sqrt{z} h a-z \tag{5.15}
\end{equation*}
$$

for the rescaled variable

$$
\begin{equation*}
z=\tau^{2} h^{2} \tag{5.16}
\end{equation*}
$$

Then (5.14) reads

$$
P(z, h) \hat{u}=h^{2} f ;
$$

and so, if $P(z, h)$ is invertible,

$$
\begin{equation*}
\hat{u}=h^{2} P(z, h)^{-1} f . \tag{5.17}
\end{equation*}
$$

We therefore need to study the inverse of $P(z, h)$.

THEOREM 5.7 (Resolvent bounds). Under the dynamical assumption (5.13), there exist constants $\alpha, C, h_{0}>0$ such that

$$
\begin{equation*}
\left\|P(z, h)^{-1}\right\|_{L^{2} \rightarrow L^{2}} \leq \frac{C}{h} \tag{5.18}
\end{equation*}
$$

for

$$
\begin{equation*}
|\operatorname{Im} z| \leq \alpha h,|z-1| \leq \alpha, 0<h \leq h_{0} \tag{5.19}
\end{equation*}
$$

Proof. 1. It is enough to show that there exists a constant $C$ such that

$$
\|u\|_{L^{2}} \leq \frac{C}{h}\|P(z, h) u\|_{L^{2}}
$$

for all $u \in L^{2}$, provided $z$ and $h$ satisfy (5.19).
We argue by contradiction. If the assertion were false, then for $m=$ $1,2, \ldots$ there would exist $z_{m} \in \mathbb{C}, 0<h_{m} \leq 1 / m$ and functions $u_{m}$ in $L^{2}$ such that

$$
\left\|P\left(z_{m}, h_{m}\right) u_{m}\right\|_{L^{2}} \leq \frac{h_{m}}{m}\left\|u_{m}\right\|_{L^{2}}, \quad\left|\operatorname{Im} z_{m}\right| \leq \frac{h_{m}}{m}, \quad\left|z_{m}-1\right| \leq \frac{1}{m} .
$$

We may assume $\left\|u_{m}\right\|_{L^{2}}=1$. Then

$$
\begin{equation*}
P\left(z_{m}, h_{m}\right) u_{m}=o\left(h_{m}\right) . \tag{5.20}
\end{equation*}
$$

Also,

$$
\begin{equation*}
z_{m} \rightarrow 1, \operatorname{Im}\left(z_{m}\right)=o\left(h_{m}\right) \tag{5.21}
\end{equation*}
$$

2. Let $\mu$ be a microlocal defect measure associated with $\left\{u_{m}\right\}_{m=1}^{\infty}$. Then Theorem 5.4 implies for the symbol $p:=|\xi|^{2}-1$ that

$$
\operatorname{spt}(\mu) \subseteq p^{-1}(0)=\left\{|\xi|^{2}=1\right\}
$$

$\operatorname{But}\left\langle u_{m}, u_{m}\right\rangle=1$, and so

$$
\begin{equation*}
\int_{\mathbb{T}^{n} \times \mathbb{R}^{n}} d \mu=1 \tag{5.22}
\end{equation*}
$$

We will derive a contradiction to this.
3. Hereafter write $P_{m}:=P\left(z_{m}, h_{m}\right)$. Then

$$
\begin{aligned}
& P_{m}=-h_{m}^{2} \Delta+i \sqrt{z_{m}} h_{m} a-z_{m}, \\
& P_{m}^{*}=-h_{m}^{2} \Delta-i \sqrt{\overline{z_{m}}} h_{m} a-\bar{z}_{m} ;
\end{aligned}
$$

and therefore

$$
\begin{equation*}
P_{m}-P_{m}^{*}=i\left(\sqrt{z_{m}}+\sqrt{\bar{z}_{m}}\right) h_{m} a-z_{m}+\bar{z}_{m}=2 i h_{m} a+o\left(h_{m}\right), \tag{5.23}
\end{equation*}
$$

since (5.21) implies that $\sqrt{z_{m}}+\sqrt{\overline{z_{m}}}=2+o(1)$ and that $-z_{m}+\bar{z}_{m}=$ $-2 i \operatorname{Im}\left(z_{m}\right)=o\left(h_{m}\right)$.

Now select $b \in C_{\mathrm{c}}^{\infty}\left(\mathbb{T}^{n} \times \mathbb{R}^{n}\right)$ and set $B_{m}:=b^{\mathrm{w}}\left(x, h_{m} D\right)$. Then $B_{m}=B_{m}^{*}$. Using (5.20) and (5.23), we calculate that

$$
\begin{aligned}
o\left(h_{m}\right)=2 i \operatorname{Im}\left\langle B_{m} P_{m} u_{m}, u_{m}\right\rangle= & \left\langle B_{m} P_{m} u_{m}, u_{m}\right\rangle-\left\langle u, B_{m} P_{m} u_{m}\right\rangle \\
= & \left\langle\left(B_{m} P_{m}-P_{m}^{*} B_{m}\right) u_{m}, u_{m}\right\rangle \\
= & \left\langle\left[B_{m}, P_{m}\right] u_{m}, u_{m}\right\rangle \\
& +\left\langle\left(P_{m}-P_{m}^{*}\right) B_{m} u_{m}, u_{m}\right\rangle \\
= & \frac{h_{m}}{i}\left\langle\{b, p\}^{\mathrm{w}} u_{m}, u_{m}\right\rangle \\
& +2 h_{m} i\left\langle(a b)^{\mathrm{w}} u_{m}, u_{m}\right\rangle+o\left(h_{m}\right) .
\end{aligned}
$$

Divide by $h_{m}$ and let $h_{m} \rightarrow 0$, through a subsequence if necessary, to discover that

$$
\begin{equation*}
\int_{\mathbb{T}^{n} \times \mathbb{R}^{n}}\{p, b\}+2 a b d \mu=0 . \tag{5.24}
\end{equation*}
$$

We will build a function $b$ so that $\{p, b\}+2 a b>0$ on $\operatorname{spt}(\mu)$. This will imply $\int_{\mathbb{T}^{n} \times \mathbb{R}^{n}} d \mu=0$, a contradiction to (5.22).
4. For $(x, \xi) \in \mathbb{T}^{n} \times \mathbb{R}^{n}$, with $|\xi|=1$, define

$$
c(x, \xi):=\frac{1}{T} \int_{0}^{T}(T-t) a(x+\xi t) d t
$$

where $T$ is the time from the dynamical hypothesis (5.13). Hence

$$
\begin{aligned}
\left\langle\xi, \partial_{x} c\right\rangle & =\frac{1}{T} \int_{0}^{T}(T-t)\langle\xi, \partial a(x+\xi t)\rangle d t \\
& =\frac{1}{T} \int_{0}^{T}(T-t) \frac{d}{d t} a(x+\xi t) d t \\
& =\frac{1}{T} \int_{0}^{T} a(x+\xi t) d t-a(x) \\
& =\langle a\rangle_{T}-a .
\end{aligned}
$$

Let

$$
b:=e^{c} \chi(p),
$$

where $\chi \in C_{\mathrm{c}}^{\infty}(\mathbb{R})$ is equal to 1 near 0 . Then

$$
\left\langle\xi, \partial_{x} b\right\rangle=e^{c}\left\langle\xi, \partial_{x} c\right\rangle \chi(p)=e^{c}\langle a\rangle_{T} \chi(p)-a e^{c} \chi(p)
$$

since $H_{p}(\chi(p))=0$. Consequently

$$
\{p, b\}+2 a b=2\left\langle\xi, \partial_{x} b\right\rangle+2 a b=2 e^{c}\langle a\rangle_{T} \chi(p)>0 \quad \text { on } p^{-1}(0),
$$

as desired.

THEOREM 5.8 (Exponential energy decay). Assume the $d y$ namic hypothesis (5.13) and suppose $u$ solves the wave equation with damping (5.12).

Then there exists constants $C, \beta>0$ such that

$$
\begin{equation*}
E(t) \leq C e^{-\beta t}\|f\|_{L^{2}} \quad \text { for all times } t>0 . \tag{5.25}
\end{equation*}
$$

Motivation. The following calculations are based upon this idea: to get decay estimates of $g$ on the positive real axis, we estimate $\hat{g}$ in a complex strip $|\operatorname{Im} z| \leq \alpha$. Then if $\beta<\alpha$,

$$
\widehat{e^{\beta t}} g(\tau)=\int_{-\infty}^{\infty} e^{\beta t} g(t) e^{-i t \tau} d t=\int_{-\infty}^{\infty} g(t) e^{-i t(\tau+i \beta)} d t=\hat{g}(\tau+i \beta)
$$

Hence our $L^{2}$ estimate of $\hat{g}(\cdot+i \beta)$ will imply exponential decay of $g(t)$ for $t \rightarrow \infty$.

Proof. 1. Recall from (5.15), (5.16) that

$$
P(\tau)=h^{-2} P(z, h) \quad \text { for } \tau^{2}=h^{-2} z .
$$

First we assert that there exists $\gamma>0$ such that

$$
\begin{equation*}
\left\|P(\tau)^{-1}\right\|_{L^{2} \rightarrow H^{1}} \leq C \quad \text { for } \quad|\operatorname{Im} \tau| \leq \gamma,|\tau|>1 / \gamma \tag{5.26}
\end{equation*}
$$

To prove (5.26), we note that provided the inequalities (5.19) hold, then

$$
\left\|h^{2} \Delta P(z, h)^{-1} u\right\|_{L^{2}}=\left\|(i \sqrt{z} h a-z) P(z, h)^{-1} u-u\right\|_{L^{2}} \leq \frac{C}{h}\|u\|_{L^{2}}
$$

the last inequality holding according to Theorem 5.7. Thus

$$
\begin{equation*}
\left\|h^{2} P(z, h)^{-1} u\right\|_{H^{2}} \leq \frac{C}{h}\|u\|_{L^{2}} . \tag{5.27}
\end{equation*}
$$

Recall next that $z=h^{2} \tau^{2}$. Write $\tau=\lambda+i \mu$, for $\lambda>0$, and set $h=\lambda^{-1}$; so that

$$
z=h^{2}\left(\lambda^{2}-\mu^{2}\right)+i\left(h^{2} 2 \lambda \mu\right) .
$$

Thus $|\operatorname{Im} z| \leq \alpha h$ and $|z-1| \leq \alpha$ provided if $|\mu| \leq \gamma$ and $|\lambda|>1 / \gamma$ for some sufficiently small $\gamma$, and so the inequalities (5.19) hold. Hence (5.27) implies

$$
\begin{equation*}
\left\|P(\tau)^{-1} u\right\|_{H^{2}} \leq \frac{C}{|\tau|}\|u\|_{L^{2}} \tag{5.28}
\end{equation*}
$$

for $|\mu| \leq \gamma$ and $|\lambda|>1 / \gamma$.
Also

$$
\left\|P(\tau)^{-1} u\right\|_{L^{2}} \leq \frac{C}{|\tau|}\|u\|_{L^{2}} .
$$

Interpolating between the last two inequalities demonstrates that

$$
\left\|P(\tau)^{-1} u\right\|_{H^{1}} \leq C\|u\|_{L^{2}}
$$

for $|\operatorname{Im} \tau| \leq \gamma$ and $|\tau|>1 / \gamma$.
This proves (5.26) except for a bounded range of $\tau$ 's. To see the estimate for all $\tau$, outside of a neighborhood of 0 , we simply need to exclude the possibility of a real non-zero $\tau$ satisfying

$$
\begin{equation*}
\left(-\Delta-\tau^{2}+i \tau a\right) u=0 \tag{5.29}
\end{equation*}
$$

for some $u \neq 0$. Multiplying by $\bar{u}$, integrating, and taking the imaginary part shows that

$$
\int_{\mathbb{T}^{n}} a|u|^{2} d x=0 .
$$

Since $a \geq 0$, this implies that $u \equiv 0$ on spt $a$. Hence $\left(-\Delta-\tau^{2}\right) u=0$. But this is impossible owing to unique continuation results which we will prove in Section 7.2, since spt $a$ has a nonempty interior. The Fredholm alternative now guarantees that $P(\tau)^{-1}$ has no pole on the real axis.
2. Since $P(\tau)^{-1}$ is meromorphic, we conclude that in $|\operatorname{Im} \tau| \leq \alpha$ the only possible pole occurs at $\tau=0$. In particular for $0<\beta<\alpha$ we have

$$
\begin{equation*}
\sup _{\tau \in \mathbb{R}}\left\|P(\tau+i \beta)^{-1}\right\|_{L^{2} \rightarrow H^{1}} \leq C_{\beta} . \tag{5.30}
\end{equation*}
$$

3. Next select $\chi: \mathbb{R} \rightarrow \mathbb{R}, \chi=\chi(t)$, such that

$$
0 \leq \chi \leq 1, \chi \equiv 1 \text { on }[1, \infty), \chi \equiv 0 \text { on }(-\infty, 0)
$$

Then if $u_{1}:=\chi u$, we have

$$
\begin{equation*}
\left(\partial_{t}^{2}+a \partial_{t}-\Delta\right) u_{1}=g_{1}, \tag{5.31}
\end{equation*}
$$

for

$$
\begin{equation*}
g_{1}:=\chi^{\prime \prime} u+2 \chi^{\prime} \partial_{t} u+a(x) \chi^{\prime} u . \tag{5.32}
\end{equation*}
$$

Note that $u_{1}(t)=0$ for $t \leq 0$, and observe also that the support of $g_{1}$ lies within $\mathbb{T}^{n} \times[0,1]$. Furthermore, using energy estimates in Lemma 5.6 , we see that

$$
\begin{align*}
& \left\|g_{1}\right\|_{L^{2}\left(\mathbb{R}_{+} ; L^{2}\right)} \\
& \quad \leq C\left(\|u\|_{L^{2}\left((0,1) ; L^{2}\right)}+\left\|\partial_{t} u\right\|_{L^{2}\left((0,1) ; L^{2}\right)}\right) \leq C\|f\|_{L^{2}} . \tag{5.33}
\end{align*}
$$

Now take the Fourier transform of (5.31) in time:

$$
P(\tau) \hat{u}_{1}(\tau)=\hat{g}_{1}(\tau) .
$$

Then

$$
\begin{equation*}
\hat{u}_{1}(\tau)=P(\tau)^{-1} \hat{g}_{1}(\tau) \tag{5.34}
\end{equation*}
$$

where, in principle, we allow the left hand side to have a pole at $\tau=0$.
4. We now deduce exponential decay. Noting that $u_{1}$ is supported in $t>0$, we use Plancherel's theorem to compute

$$
\begin{aligned}
\left\|e^{\beta t} u_{1}\right\|_{L^{2}\left(\mathbb{R}_{+} ; H^{1}\right)} & =(2 \pi)^{-\frac{1}{2}}\left\|\widehat{e^{\beta t} u_{1}}\right\|_{L^{2}\left(\mathbb{R} ; H^{1}\right)} \\
& =(2 \pi)^{-\frac{1}{2}}\left\|\hat{u}_{1}(\cdot+i \beta)\right\|_{L^{2}\left(\mathbb{R} ; H^{1}\right)} \\
& =(2 \pi)^{-\frac{1}{2}}\left\|P(\cdot+i \beta)^{-1} \hat{g}_{1}(\cdot+i \beta)\right\|_{L^{2}\left(\mathbb{R} ; H^{1}\right)} \\
& \leq C\left\|\hat{g}_{1}(\cdot+i \beta)\right\|_{L^{2}\left(\mathbb{R} ; H^{1}\right)}
\end{aligned}
$$

Since $g_{1}$ is compactly supported in $t$ we also see that

$$
\hat{g}_{1}(\cdot+i \beta)=\widehat{e^{\beta t} g_{1}}(\cdot) ;
$$

and hence

$$
\begin{aligned}
\left\|e^{\beta t} u_{1}\right\|_{L^{2}\left(\mathbb{R}_{+} ; H^{1}\right)} & \leq C\left\|\hat{g}_{1}\right\|_{L^{2}\left(\mathbb{R} ; L^{2}\right)} \\
& \leq C\left\|g_{1}\right\|_{L^{2}\left(\mathbb{R}_{+} ; L^{2}\right)} \leq C\|f\|_{L^{2}} .
\end{aligned}
$$

Since $u_{1}=\chi u$, it follows that

$$
\begin{equation*}
\left\|e^{\beta t} u\right\|_{L^{2}\left((1, \infty) ; H^{1}\right)} \leq C\|f\|_{L^{2}} \tag{5.35}
\end{equation*}
$$

5. Finally, fix $T>2$ and

$$
\chi_{T}:=\chi(t-T+1),
$$

where $\chi$ is as in Step 2. Let $u_{2}=\chi_{T} u$. Then

$$
\begin{equation*}
\left(\partial_{t}^{2}+a \partial_{t}-\Delta\right) u_{2}=g_{2}, \tag{5.36}
\end{equation*}
$$

for

$$
\begin{equation*}
g_{2}:=\chi_{T}^{\prime \prime} u+2 \chi_{T}^{\prime} \partial_{t} u+a \chi_{T}^{\prime} u . \tag{5.37}
\end{equation*}
$$

Therefore $\operatorname{spt} g \subseteq \mathbb{T}^{n} \times(T-1, T)$.

Define

$$
E_{2}(t):=\frac{1}{2} \int_{\mathbb{T}^{n}}\left(\partial_{t} u_{2}\right)^{2}+\left|\partial_{x} u_{2}\right|^{2} d x
$$

Modifying the calculations in the proof of Lemma 5.6, we use (5.36) and (5.37) to compute

$$
\begin{aligned}
E_{2}^{\prime}(t) & =\int_{\mathbb{T}^{n}} \partial_{t} u_{2} \partial_{t}^{2} u_{2}+\left\langle\partial_{x} u_{2}, \partial_{x t}^{2} u_{2}\right\rangle d x \\
& =\int_{\mathbb{T}^{n}} \partial_{t} u_{2}\left(\partial_{t}^{2} u_{2}-\Delta u_{2}\right) d x \\
& =-\int_{\mathbb{T}^{n}} a\left(\partial_{t} u_{2}\right)^{2} d x+\int_{\mathbb{T}^{n}} \partial_{t} u_{2} g_{2} d x \\
& \leq C \int_{\mathbb{T}^{n}}\left|\partial_{t} u_{2}\right|\left(\left|\partial_{t} u\right|+|u|\right) d x \\
& \leq C E_{2}(t)+C \int_{\mathbb{T}^{n}} u^{2}+\left(\partial_{t} u\right)^{2} d x
\end{aligned}
$$

Since $E_{2}(T-1)=0$ and $E_{2}(T)=E(T)$, Gronwall's inequality implies that

$$
\begin{equation*}
E(T) \leq C\left(\|u\|_{L^{2}\left((T-1, T) ; L^{2}\right)}^{2}+\left\|\partial_{t} u\right\|_{L^{2}\left((T-1, T) ; L^{2}\right)}^{2}\right) . \tag{5.38}
\end{equation*}
$$

6. We need to control the right hand term in (5.38). For this, select $\chi: \mathbb{R} \rightarrow \mathbb{R}$, such that

$$
\left\{\begin{array}{l}
0 \leq \chi \leq 1 \\
\chi \equiv 0 \text { for } t \leq T-2 \text { and } t \geq T+1 \\
\chi \equiv 1 \text { for } T-1 \leq t \leq T
\end{array}\right.
$$

We multiply the wave equation (5.12) by $\chi^{2} u$ and integrate by parts, to find

$$
\begin{aligned}
0 & =\int_{T-2}^{T+1} \int_{\mathbb{T}^{n}} \chi^{2} u\left(\partial_{t}^{2} u+a \partial_{t} u-\Delta u\right) d x d t \\
& =\int_{T-2}^{T+1} \int_{\mathbb{T}^{n}}-\chi^{2}\left(\partial_{t} u\right)^{2}-2 \chi \chi^{\prime} u \partial_{t} u+\chi^{2} a u \partial_{t} u+\chi^{2}\left|\partial_{x} u\right|^{2} d x d t
\end{aligned}
$$

From this identity we derive the estimate

$$
\left\|\partial_{t} u\right\|_{L^{2}\left((T-1, T) ; L^{2}\right)} \leq C\|u\|_{L^{2}\left((T-2, T+1) ; H^{1}\right)} .
$$

This, (5.38) and (5.35) therefore imply

$$
E(T) \leq C\|u\|_{L^{2}\left((T-2, T+1) ; H^{1}\right)}^{2} \leq C e^{-\beta T}\|f\|_{L^{2}}
$$

as asserted.

Our methods extend with no difficulty if $\mathbb{T}^{n}$ is replaced by a general compact Riemannian manifold: see Appendix E.

## 6. Eigenvalues and eigenfunctions

6.1 The harmonic oscillator
6.2 Symbols and eigenfunctions
6.3 Spectrum and resolvents
6.4 Weyl's Law

In this chapter we are given the potential $V: \mathbb{R}^{n} \rightarrow \mathbb{R}$, and investigate how the symbol

$$
\begin{equation*}
p(x, \xi)=|\xi|^{2}+V(x) \tag{6.1}
\end{equation*}
$$

provides interesting information about the corresponding operator

$$
\begin{equation*}
P(h):=P(x, h D)=-h^{2} \Delta+V . \tag{6.2}
\end{equation*}
$$

We will focus mostly upon learning how $p$ controls the aysmptotic distribution of the eigenvalues of $P(h)$ in the semiclassical limit $h \rightarrow 0$.

### 6.1 THE HARMONIC OSCILLATOR

Our plan is to consider first the simplest case, when the potential is quadratic; and to simplify even more, we begin in one dimension. So suppose that $n=1, h=1$ and $V(x)=x^{2}$. Thus we start with the one-dimensional quantum harmonic oscillator, meaning the operator

$$
P_{0}:=-\partial^{2}+x^{2}
$$

6.1.1 Eigenvalues and eigenfunctions of $\boldsymbol{P}_{\mathbf{0}}$. We can as follows employ certain auxiliary first-order differential operators to compute explicitly the eigenvalues and eigenfunctions for $P_{0}$.

NOTATION. Let us write

$$
A_{+}:=D_{x}+i x, A_{-}:=D_{x}-i x,
$$

where $D_{x}=\frac{1}{i} \partial_{x}$, and call $A_{+}$the creation operator and $A_{-}$the annihilation operator. (This terminology is from particle physics.)

LEMMA 6.1 (Properties of $\boldsymbol{A}_{ \pm}$). The creation and annihilation operators satisfy these identities:

$$
\begin{gathered}
A_{+}^{*}=A_{-}, A_{-}^{*}=A_{+} \\
P_{0}=A_{+} A_{-}+1=A_{-} A_{+}-1 .
\end{gathered}
$$

Proof..It is easy to check that $D_{x}^{*}=D_{x}$ and $(i x)^{*}=-i x$. Furthermore,

$$
\begin{aligned}
A_{+} A_{-} u & =\left(D_{x}+i x\right)\left(D_{x}-i x\right) u \\
& =\left(\frac{1}{i} \partial_{x}+i x\right)\left(\frac{1}{i} u_{x}-i x u\right) \\
& =-u_{x x}-(x u)_{x}+x u_{x}+x^{2} u \\
& =-u_{x x}-u-x u_{x}+x u_{x}+x^{2} u \\
& =P_{0} u-u
\end{aligned}
$$

and similarly,

$$
\begin{aligned}
A_{-} A_{+} u & =\left(D_{x}-i x\right)\left(D_{x}+i x\right) u \\
& =\left(\frac{1}{i} \partial_{x}-i x\right)\left(\frac{1}{i} u_{x}+i x u\right) \\
& =-u_{x x}+(x u)_{x}-x u_{x}+x^{2} u \\
& =P_{0} u+u
\end{aligned}
$$

We can now use $A_{ \pm}$to find all the eigenvalues and eigenfunctions of $P_{0}$ :

## THEOREM 6.2 (Eigenvalues and eigenfunctions).

(i) We have

$$
\left\langle P_{0} u, u\right\rangle \geq\|u\|_{L^{2}}^{2}
$$

for all $u \in C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{n}\right)$. That is,

$$
P_{0} \geq 1
$$

(ii) The function

$$
v_{0}=: e^{-\frac{x^{2}}{2}}
$$

is an eigenfunction corresponding to the smallest eigenvalue 1.
(iii) Set

$$
v_{n}:=A_{+}^{n} v_{0}
$$

for $n=1,2, \ldots$ Then

$$
\begin{equation*}
P_{0} v_{n}=(2 n+1) v_{n} . \tag{6.3}
\end{equation*}
$$

(iv) Define the normalized eigenfunctions

$$
u_{n}:=\frac{v_{n}}{\left\|v_{n}\right\|_{L^{2}}}
$$

Then

$$
\begin{equation*}
u_{n}(x)=H_{n}(x) e^{-\frac{x^{2}}{2}} \tag{6.4}
\end{equation*}
$$

where $H_{n}(x)=c_{n} x^{n}+\cdots+c_{0} \quad\left(c_{n} \neq 0\right)$ is a polynomial of degree $n$.
(v) We have

$$
\left\langle u_{n}, u_{m}\right\rangle=\delta_{n m}
$$

and furthermore, the collection of eigenfunctions $\left\{u_{n}\right\}_{n=0}^{\infty}$ is complete in $L^{2}\left(\mathbb{R}^{n}\right)$.

The functions $H_{n}$ mentioned in assertion (iv) are the Hermite polynomials.

Proof. 1. We note that

$$
\left[D_{x}, x\right] u=\frac{1}{i}(x u)_{x}-\frac{x}{i} u_{x}=\frac{u}{i},
$$

and consequently $i\left[D_{x}, x\right]=1$. Therefore

$$
\begin{aligned}
\|u\|_{L^{2}}^{2} & =\left\langle i\left[D_{x}, x\right] u, u\right\rangle \leq 2\|x u\|_{L^{2}}\left\|D_{x} u\right\|_{L^{2}} \\
& \leq\|x u\|_{L^{2}}^{2}+\left\|D_{x} u\right\|_{L^{2}}^{2}=\left\langle P_{0} u, u\right\rangle .
\end{aligned}
$$

Next, observe

$$
A_{-} v_{0}=\frac{1}{i}\left(e^{-\frac{x^{2}}{2}}\right)_{x}-i x e^{-\frac{x^{2}}{2}}=0
$$

so that $P_{0} v_{0}=\left(A_{+} A_{-}+1\right) v_{0}=v_{0}$.
2. We can further calculate that

$$
\begin{aligned}
P_{0} v_{n} & =\left(A_{+} A_{-}+1\right) A_{+} v_{n-1} \\
& =A_{+}\left(A_{-} A_{+}-1\right) v_{n-1}+2 A_{+} v_{n-1} \\
& =A_{+} P_{0} v_{n-1}+2 A_{+} v_{n-1} \\
& =(2 n-1) A_{+} v_{n-1}+2 A_{+} v_{n-1} \quad \text { (by induction) } \\
& =(2 n+1) v_{n} .
\end{aligned}
$$

The form (6.4) of $v_{n}, u_{n}$ follows by induction.
3. Note also that

$$
\begin{aligned}
{\left[A_{-}, A_{+}\right] } & =A_{-} A_{+}-A_{+} A_{-} \\
& =\left(P_{0}+1\right)-\left(P_{0}-1\right)=2 .
\end{aligned}
$$

Hence if $m>n$,

$$
\begin{aligned}
\left\langle v_{n}, v_{m}\right\rangle & =\left\langle A_{+}^{n} v_{0}, A_{+}^{m} v_{0}\right\rangle \\
& =\left\langle A_{-}^{m} A_{+}^{n} v_{0}, v_{0}\right\rangle \quad\left(\text { since } A_{-}=A_{+}^{*}\right) \\
& =\left\langle A_{-}^{m-1}\left(A_{+} A_{-}+2\right) A_{+}^{n-1} v_{0}, v_{0}\right\rangle
\end{aligned}
$$

After finitely many steps, the foregoing equals

$$
\left\langle(\ldots) A_{-} v_{0}, v_{0}\right\rangle=0,
$$

since $A_{-} v_{0}=0$.
4. Lastly, we demonstrate that the collection of eigenfunctions we have found spans $L^{2}$. Suppose $\left\langle u_{n}, g\right\rangle=0$ for $n=0,1,2, \ldots$; we must show $g \equiv 0$.

Now since $H_{n}(x)=c_{n} x^{n}+\ldots$, with $c_{n} \neq 0$, we have

$$
\int_{-\infty}^{\infty} g(x) e^{-\frac{x^{2}}{2}} p(x) d x=0
$$

for each polynomial $p$. Hence

$$
\int_{-\infty}^{\infty} g(x) e^{-\frac{x^{2}}{2}} e^{-i x \xi} d x=\int_{-\infty}^{\infty} g(x) e^{-\frac{x^{2}}{2}} \sum_{k=0}^{\infty} \frac{(-i x \xi)^{k}}{k!} d x
$$

and so $\mathcal{F}\left(g e^{-\frac{x^{2}}{2}}\right) \equiv 0$. This implies $g e^{-\frac{x^{2}}{2}} \equiv 0$ and consequently $g \equiv 0$.
6.1.2 Higher dimensions, rescaling. Suppose now $n>1$, and write

$$
P_{0}:=-\Delta+|x|^{2} ;
$$

this is the $n$-dimensional quantum harmonic oscillator. We define also

$$
u_{\alpha}(x):=\prod_{j=1}^{n} u_{\alpha_{j}}\left(x_{j}\right)=\prod_{j=1}^{n} H_{\alpha_{j}}\left(x_{j}\right) e^{-\frac{|x|^{2}}{2}}
$$

for each multiindex $\alpha=\left(\alpha_{1}, \ldots, \alpha_{n}\right)$. Then

$$
P_{0} u_{\alpha}=\left(-\Delta+|x|^{2}\right) u_{\alpha}=(2|\alpha|+n) u_{\alpha},
$$

for $|\alpha|=\alpha_{1}+\cdots+\alpha_{n}$. Hence $u_{\alpha}$ is an eigenfunction of $P_{0}$ corresponding to the eigenvalue $2|\alpha|+n$.

We next restore the parameter $h>0$ by setting

$$
\begin{gather*}
P_{0}(h):=-h^{2} \Delta+|x|^{2},  \tag{6.5}\\
u_{\alpha}(h)(x):=h^{-\frac{n}{4}} \prod_{j=1}^{n} H_{\alpha_{j}}\left(\frac{x_{j}}{\sqrt{h}}\right) e^{-\frac{|x|^{2}}{2 h}}, \tag{6.6}
\end{gather*}
$$

and

$$
\begin{equation*}
E_{\alpha}(h):=(2|\alpha|+n) h . \tag{6.7}
\end{equation*}
$$

Then

$$
P_{0}(h) u_{\alpha}(h)=E_{\alpha}(h) u_{\alpha}(h) ;
$$

and upon reindexing, we can write these eigenfunction equations as

$$
\begin{equation*}
P_{0}(h) u_{j}(h)=E_{j}(h) u_{j}(h) \quad(j=1, \ldots) . \tag{6.8}
\end{equation*}
$$

6.1.3 Asymptotic distribution of eigenvalues. With these explicit formulas in hand, we can study the behavior in the semiclassical limit of the eigenvalues $E(h)$ of the harmonic oscillator:

THEOREM 6.3 (Weyl's law for harmonic oscillator). Assume that $0 \leq a<b<\infty$. Then

$$
\begin{align*}
\#\{E(h) \mid a \leq E(h) & \leq b\}  \tag{6.9}\\
& =\frac{1}{(2 \pi h)^{n}}\left(\left|\left\{a \leq|\xi|^{2}+|x|^{2} \leq b\right\}\right|+o(1)\right) .
\end{align*}
$$

as $h \rightarrow 0$.

Proof. We may assume that $a=0$. Since $E(h)=(2|\alpha|+n) h$ for some multiindex $\alpha$ according to (6.7), we have

$$
\begin{aligned}
\#\{E(h) \mid 0 \leq E(h) \leq b\} & =\#\left\{\alpha|0 \leq 2| \alpha \left\lvert\,+n \leq \frac{b}{h}\right.\right\} \\
& =\#\left\{\alpha \mid \alpha_{1}+\cdots+\alpha_{n} \leq R\right\}
\end{aligned}
$$

for $R:=\frac{b-n h}{2 h}$. Therefore

$$
\begin{aligned}
\#\{E(h) \mid & 0 \leq E(h) \leq b\} \\
& =\left|\left\{x \mid x_{i} \geq 0, x_{1}+\cdots+x_{n} \leq R\right\}\right|+o\left(R^{n}\right) \\
& =\frac{1}{n!} R^{n}+o\left(R^{n}\right) \quad \text { as } R \rightarrow \infty \\
& =\frac{1}{n!}\left(\frac{b}{2 h}\right)^{n}+o\left(h^{-n}\right) \quad \text { as } h \rightarrow 0 .
\end{aligned}
$$

We used in this calculation the fact that the volume of the simplex $\left\{x \mid x_{i} \geq 0, x_{1}+\cdots+x_{n} \leq 1\right\}$ is $(n!)^{-1}$. Next we note that $\mid\left\{|\xi|^{2}+\right.$ $\left.|x|^{2} \leq b\right\} \mid=\alpha(2 n) b^{n}$, where $\alpha(k):=\pi^{\frac{k}{2}}\left(\Gamma\left(\frac{k}{2}+1\right)\right)^{-1}$ is the volume of the unit ball in $\mathbb{R}^{k}$. Setting $k=2 n$, we compute that $\alpha(2 n)=\pi^{n}(n!)^{-1}$. Hence

$$
\begin{aligned}
\#\{E(h) \mid 0 \leq E(h) \leq b\} & =\frac{1}{n!}\left(\frac{b}{2 h}\right)^{n}+o\left(h^{-n}\right) \\
& =\frac{1}{(2 \pi h)^{n}}\left|\left\{|\xi|^{2}+|x|^{2} \leq b\right\}\right|+o\left(h^{-n}\right) .
\end{aligned}
$$

### 6.2 SYMBOLS AND EIGENFUNCTIONS

For this section, we return to the general symbol (6.1) and the quantized operator (6.2). We assume that the potential $V: \mathbb{R}^{n} \rightarrow \mathbb{R}$ is smooth, and satisfies the growth conditions:

$$
\begin{equation*}
\left|\partial^{\alpha} V(x)\right| \leq C_{\alpha}\langle x\rangle^{k}, \quad V(x) \geq C\langle x\rangle^{k} \quad \text { for }|x| \geq R, \tag{6.10}
\end{equation*}
$$

for appropriate constants $k, C, C_{\alpha}, R>0$.
Our plan in the next section is to employ our detailed knowledge about the eigenvalues of the harmonic oscillator (6.5) to estimate the asymptotics of the eigenvalues of $P(h)$. This section develops some useful techniques that will aid us in this task.
6.2.1 Concentration in phase space. First, we make the important observation that in the semiclassical limit the eigenfunctions $u(h)$ "are concentrated in phase space" on the energy surface $\left\{|\xi|^{2}+V(x)=E\right\}$.

THEOREM $6.4\left(\boldsymbol{h}^{\infty}\right.$ estimates). Suppose that $u(h) \in L^{2}\left(\mathbb{R}^{n}\right)$ solves

$$
\begin{equation*}
P(h) u(h)=E(h) u(h) . \tag{6.11}
\end{equation*}
$$

Assume as well that $a \in S$ is a symbol satisfying

$$
\left\{|\xi|^{2}+V(x)=E\right\} \cap \operatorname{spt}(a)=\emptyset .
$$

Then if

$$
|E(h)-E|<\delta
$$

for some sufficiently small $\delta>0$, we have the estimate

$$
\begin{equation*}
\left\|a^{\mathrm{w}}(x, h D) u(h)\right\|_{L^{2}}=O\left(h^{\infty}\right)\|u(h)\|_{L^{2}} . \tag{6.12}
\end{equation*}
$$

Proof. 1. The set $K:=\left\{|\xi|^{2}+V(x)=E\right\} \subset \mathbb{R}^{2 n}$ has compact. Hence there exists $\chi \in C_{c}^{\infty}\left(\mathbb{R}^{2 n}\right)$ such that

$$
0 \leq \chi \leq 1, \chi \equiv 1 \text { on } K, \chi \equiv 0 \text { on } \operatorname{spt}(a)
$$

Define the symbol

$$
b:=|\xi|^{2}+V(x)-E(h)+i \chi=p-E(h)+i \chi
$$

and the order function

$$
m:=\langle\xi\rangle^{2}+\langle x\rangle^{k} .
$$

Therefore if $|E(h)-E|$ is small enough,

$$
|b| \geq \gamma m \quad \text { on } \mathbb{R}^{2 n}
$$

for some constant $\gamma>0$. Consequently $b \in S(m)$, with $b^{-1} \in S\left(m^{-1}\right)$.
2. Thus there exist $c \in S\left(m^{-1}\right), r_{1}, r_{2} \in S$ such that

$$
\left\{\begin{array}{l}
b^{\mathrm{w}}(x, h D) c^{\mathrm{w}}(x, h D)=I+r_{1}^{\mathrm{w}}(x, h D) \\
c^{\mathrm{w}}(x, h D) b^{\mathrm{w}}(x, h D)=I+r_{2}^{\mathrm{w}}(x, h D)
\end{array}\right.
$$

where $r_{1}^{\mathrm{w}}(x, h D), r_{2}^{\mathrm{w}}(x, h D)$ are $O\left(h^{\infty}\right)$. Then

$$
\begin{equation*}
a^{\mathrm{w}}(x, h D) c^{\mathrm{w}}(x, h D) b^{\mathrm{w}}(x, h D)=a^{\mathrm{w}}(x, h D)+O\left(h^{\infty}\right), \tag{6.13}
\end{equation*}
$$

and

$$
\begin{equation*}
b^{\mathrm{w}}(x, h D)=P(h)-E(h)+i \chi^{\mathrm{w}}(x, h D) . \tag{6.14}
\end{equation*}
$$

Furthermore

$$
a^{\mathrm{w}}(x, h D) c^{\mathrm{w}}(x, h D) \chi^{\mathrm{w}}(x, h D)=O\left(h^{\infty}\right)
$$

since $\operatorname{spt}(a) \cap \operatorname{spt}(\chi)=\emptyset$. Since $P(h) u(h)=E(h) u(h)$, (6.13) and (6.14) imply that

$$
\begin{array}{r}
a^{\mathrm{w}}(x, h D) u(h)=a^{\mathrm{w}}(x, h D) c^{\mathrm{w}}(x, h D)\left(P(h)-E(h)+i \chi^{\mathrm{w}}\right) u(h)+O\left(h^{\infty}\right) \\
=O\left(h^{\infty}\right)
\end{array}
$$

For the next result, we temporarily return to the case of the quantum harmonic oscillator, developing some sharper estimates:

THEOREM 6.5 (Improved estimates for the harmonic oscillator). Suppose that $u(h) \in L^{2}\left(\mathbb{R}^{n}\right)$ is an eigenfuction of the harmonic oscillator:

$$
\begin{equation*}
P_{0}(h) u(h)=E(h) u(h) . \tag{6.15}
\end{equation*}
$$

Assume also that $a \in C_{\mathrm{c}}^{\infty}$.
Then there exists $R>0$, depending only on the support of $a$, such that for $E(h)>R$,

$$
\begin{equation*}
\left\|a^{\mathrm{w}}(x, h D) u(h)\right\|_{L^{2}}=O\left(\left(\frac{h}{E(h)}\right)^{\infty}\right)\|u(h)\|_{L^{2}} . \tag{6.16}
\end{equation*}
$$

The precise form of the right hand side of (6.16) will later let us handle eigenvalues $E(h) \rightarrow \infty$.

Proof. 1. We rescale the harmonic oscillator so that we can work near a fixed energy level $E$. Set

$$
y:=\frac{x}{E^{\frac{1}{2}}}, \quad \tilde{h}:=\frac{h}{E}, \quad E(\tilde{h}):=\frac{E(h)}{E},
$$

where we choose $E$ so that $|E(h)-E| \leq E / 4$. Then put

$$
P_{0}(h):=-h^{2} \Delta_{x}+|x|^{2}, \quad P_{0}(\tilde{h}):=-\tilde{h}^{2} \Delta_{y}+|y|^{2} ;
$$

whence

$$
P_{0}(h)-E(h)=E(P(\tilde{h})-\widetilde{E}(\tilde{h})) .
$$

We next introduce the unitary transformation

$$
U u(y):=E^{\frac{n}{2}} u\left(E^{\frac{1}{2}} y\right)
$$

Then

$$
U P_{0}(h) U^{-1}=E P_{0}(\tilde{h}) ;
$$

and more generally

$$
U b^{\mathrm{w}}(x, h D) U^{-1}=\widetilde{b}^{\mathrm{w}}(y, \tilde{h} D), \quad \widetilde{b}(y, \eta):=b\left(E^{\frac{1}{2}} y, E^{\frac{1}{2}} \eta\right) .
$$

We will denote the symbol classes defined using $\tilde{h}$ by the symbol $\widetilde{S}_{\delta}$.
2. We now apply Theorem 6.4, to eigenfuctions of $P_{0}(\tilde{h})$. If

$$
\left(P_{0}(\tilde{h})-E(\tilde{h})\right) \widetilde{u}(\tilde{h})=0, \quad|E(\tilde{h})-1|<\delta,
$$

and $\widetilde{b}(y, \eta) \in \widetilde{S}$ has its support contained in

$$
\left\{|y|^{2}+|\eta|^{2} \leq 1 / 2\right\}
$$

then

$$
\left\|\widetilde{b}^{\mathrm{w}}(y, \tilde{h} D) \widetilde{u}(\tilde{h})\right\|_{L^{2}}=O\left(\tilde{h}^{\infty}\right)\|\widetilde{u}(\tilde{h})\|_{L^{2}} .
$$

Translated to the original $h$ and $x$ as above, this assertion provides us with the bound

$$
\begin{equation*}
\left\|b^{\mathrm{w}}(x, h D) u(h)\right\|_{L^{2}}=O\left((h / E)^{\infty}\right)\|u(h)\|_{L^{2}}, \tag{6.17}
\end{equation*}
$$

for

$$
b(x, \xi)=\widetilde{b}\left(E^{-1 / 2} x, E^{-1 / 2} \xi\right) \in S
$$

Note that $\operatorname{spt}(b) \subset\left\{|x|^{2}+|\xi|^{2} \leq E / 2\right\}$.
3. In view of (6.17), we only need to show that for

$$
a \in C^{\infty}\left(\mathbb{R}^{2 n}\right), \quad \operatorname{spt}(a) \subset\left\{|x|^{2}+|\xi|^{2} \leq 1 / 4\right\}
$$

we have

$$
\|\left(a^{\mathrm{w}}(x, h D)\left(1-b^{\mathrm{w}}(x, h D)\right) \|_{L^{2} \rightarrow L^{2}}=O\left((h / E)^{\infty}\right)\right.
$$

for $E$ large enough, where $b$ is as in (6.17). That is the same as showing

$$
\begin{equation*}
\left\|\widetilde{a}^{\mathrm{w}}(y, \tilde{h} D)\left(1-\widetilde{b}^{\mathrm{w}}(y, \tilde{h} D)\right)\right\|_{L^{2} \rightarrow L^{2}}=O\left(\tilde{h}^{\infty}\right) \tag{6.18}
\end{equation*}
$$

for

$$
\widetilde{a}(y, \eta)=a\left(E^{\frac{1}{2}} y, E^{\frac{1}{2}} \eta\right)
$$

We first observe that $E=h / \tilde{h}<1 / \tilde{h}$ and hence

$$
\widetilde{a} \in \widetilde{S}_{\frac{1}{2}} .
$$

Since the support of $a$ is compact, we see that for $E$ large enough,

$$
\operatorname{dist}(\operatorname{spt}(\widetilde{a}), \operatorname{spt}(1-\widetilde{b})) \geq 1 / C>0
$$

uniformly in $\tilde{h}$. The estimate (6.18) is now a consequence of Theorem 4.18
6.2.2 Projections. We next study how projections onto the span of various eigenfunctions of the harmonic oscillator $P_{0}(h)$ are related to our symbol calculus.

THEOREM 6.6 (Projections and symbols). Suppose for the symbol $a \in S$ that

$$
\operatorname{spt}(a) \subset\left\{|\xi|^{2}+|x|^{2}<R\right\}
$$

Let

$$
\begin{aligned}
& \Pi:=\text { projection in } L^{2} \text { onto } \\
& \qquad \operatorname{span}\left\{u(h) \mid P_{0}(h) u(h)=E(h) u(h) \text { for } E(h) \leq R\right\} .
\end{aligned}
$$

Then

$$
\begin{equation*}
\left\|a^{\mathrm{w}}(x, h D)(I-\Pi)\right\|_{L^{2} \rightarrow L^{2}}=O\left(h^{\infty}\right) \tag{6.19}
\end{equation*}
$$

and

$$
\begin{equation*}
\left\|(I-\Pi) a^{\mathrm{w}}(x, h D)\right\|_{L^{2} \rightarrow L^{2}}=O\left(h^{\infty}\right) \tag{6.20}
\end{equation*}
$$

Proof. First of all, observe

$$
(I-\Pi)=\sum_{E_{j}(h)>R} u_{j}(h) \otimes u_{j}(h),
$$

meaning that

$$
(I-\Pi) u=\sum_{E_{j}(h)>R}\left\langle u_{j}(h), u\right\rangle u_{j}(h) .
$$

Therefore

$$
a^{\mathrm{w}}(x, h D)(I-\Pi)=\sum_{E_{j}(h)>R}\left(a^{\mathrm{w}}(x, h D) u_{j}(h)\right) \otimes u_{j}(h) ;
$$

and so

$$
\begin{equation*}
\left\|a^{\mathrm{w}}(x, h D)(I-\Pi)\right\|_{L^{2} \rightarrow L^{2}} \leq\left(\sum_{E_{j}(h)>R}\left\|a^{\mathrm{w}}(x, h D) u_{j}(h)\right\|_{L^{2}}^{2}\right)^{\frac{1}{2}} . \tag{6.21}
\end{equation*}
$$

Next, observe that Weyl's Law for the harmonic oscillator, Theorem 6.3, implies that

$$
E_{j}(h) \geq \gamma j^{\frac{1}{n}} h
$$

for some constant $\gamma>0$. According then to Theorem 6.5, for each $M<N$ we have

$$
\begin{aligned}
\left\|a^{\mathrm{w}}(x, h D) u_{j}(h)\right\|_{L^{2}} & \leq C_{N}\left(\frac{h}{E_{j}(h)}\right)^{N} \\
& \leq C h^{M}\left(\frac{h}{E_{j}(h)}\right)^{N-M} \\
& \leq C h^{M} j^{-\frac{N-M}{n}} .
\end{aligned}
$$

Consequently, if we fix $N-M>n$, the sum on the right hand side of (6.21) is less than or equal to $C h^{M}$. This proves (6.19), and the proof of (6.20) is similar.

### 6.3 SPECTRUM AND RESOLVENTS

We next show that the spectrum of $P(h)$ consists entirely of eigenvalues.

THEOREM 6.7 (Resolvents and spectrum). There exists a constant $h_{0}>0$ such that if $0<h \leq h_{0}$, then the resolvent

$$
(P(h)-z)^{-1}: L^{2}\left(\mathbb{R}^{n}\right) \rightarrow L^{2}\left(\mathbb{R}^{n}\right)
$$

is a meromorphic function of $z$ with only simple, real poles.
In particular, the spectrum of $P(h)$ is discrete.

Proof. 1. Let $|z| \leq E$, where $E$ is fixed; and as before let $P_{0}(h)=$ $-h^{2} \Delta+|x|^{2}$ be the harmonic oscillator. As in Theorem 6.6 define
$\Pi:=$ projection in $L^{2}$ onto

$$
\operatorname{span}\left\{u \mid P_{0}(h) u=E(h) u \text { for } E(h) \leq R+1\right\} .
$$

Suppose now $\operatorname{spt}(a) \subset\left\{|x|^{2}+|\xi|^{2} \leq R\right\}$. Owing to Theorem 6.6, we have

$$
\left\|a^{\mathrm{w}}(x, h D)-a^{\mathrm{w}}(x, h D) \Pi\right\|_{L^{2} \rightarrow L^{2}}=O\left(h^{\infty}\right)
$$

and

$$
\left\|a^{\mathrm{w}}(x, h D)-\Pi a^{\mathrm{w}}(x, h D)\right\|_{L^{2} \rightarrow L^{2}}=O\left(h^{\infty}\right)
$$

2. Fix $R>0$ so large that

$$
\left\{|\xi|^{2}+V(x) \leq E\right\} \subset\left\{|x|^{2}+|\xi|^{2}<R\right\}
$$

Select $\chi \in C^{\infty}\left(\mathbb{R}^{2 n}\right)$ with $\operatorname{spt}(\chi) \subset\left\{|x|^{2}+|\xi|^{2} \leq R\right\}$ so that

$$
|\xi|^{2}+V(x)-z+\chi \geq \gamma m
$$

for $m=\langle\xi\rangle^{2}+\langle x\rangle^{k}$ and all $|z| \leq E$. Then $\chi=\Pi \chi \Pi+O\left(h^{\infty}\right)$. Recall that the symbolic calculus guarantees that $P(h)-z+\chi$ is invertible, if $h$ is small enough. Consequently, so is $P(h)-z+\Pi \chi \Pi$, since the two operators differ by an $O\left(h^{\infty}\right)$ term.
3. Now write

$$
P(h)-z=P(h)-z+\Pi \chi \Pi-\Pi \chi \Pi
$$

Consequently

$$
P(h)-z=(P(h)-z+\Pi \chi \Pi)\left(I-(P(h)-z+\Pi \chi \Pi)^{-1} \Pi \chi \Pi\right)
$$

Note that $\Pi \chi \Pi$ is an operator of finite rank. So Theorem D. 4 asserts that the family of operators

$$
\left(I-(P(h)-z+\Pi \chi \Pi)^{-1} \Pi \chi \Pi\right)^{-1}
$$

is meromorphic in $z$. It follows that $(P(h)-z)^{-1}$ is meromorphic on $L^{2}$. The poles are the eigenvalues, and the self-adjointness of $P(h)$ implies these eigenvalues are real and simple.
REMARK: Theorem 6.7 can be obtained more directly by using the Spectral Theorem and compactness of $(P(h)+i)^{-1}$.

### 6.4 WEYL'S LAW

We are now ready for the main result of this chapter:
THEOREM 6.8 (Weyl's Law). Suppose that $V$ satisfies the conditions (6.10) and that $E(h)$ are the eigenvalues of $P(h)=-h^{2} \Delta+V(x)$.

Then for each $a<b$, we have

$$
\begin{align*}
\#\{E(h) \mid a \leq E(h) & \leq b\}  \tag{6.22}\\
& =\frac{1}{(2 \pi h)^{n}}\left(\left|\left\{a \leq|\xi|^{2}+V(x) \leq b\right\}\right|+o(1)\right)
\end{align*}
$$

as $h \rightarrow 0$.

Proof. 1. Let

$$
N(\lambda)=\#\{E(h) \mid E(h) \leq \lambda\} .
$$

Select $\chi \in C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{2 n}\right)$ so that

$$
\chi \equiv 1 \text { on }\{p \leq \lambda+\epsilon\}, \chi \equiv 0 \text { on }\{p \geq \lambda+2 \epsilon\} .
$$

Then

$$
a:=p+(\lambda+\epsilon) \chi-\lambda \geq \gamma_{\epsilon} m
$$

for $m=\langle\xi\rangle^{2}+\langle x\rangle^{m}$ and some constant $\gamma_{\epsilon}>0$. Hence $a$ is elliptic; and so for small $h>0, a^{\mathrm{w}}(x, h D)$ is invertible.
2. Claim \#1: We have

$$
\begin{equation*}
\left\langle\left(P(h)+(\lambda+\epsilon) \chi^{\mathrm{w}}(x, h D)-\lambda\right) u, u\right\rangle \geq \gamma\|u\|_{L^{2}}^{2} \tag{6.23}
\end{equation*}
$$

for some $\gamma>0$.
To see this, take $b \in S\left(m^{1 / 2}\right)$ so that $b^{2}=a$. Then $b^{2}=b \# b+r_{0}$, where $r_{0} \in S^{-1}(m)$. We also recall from Theorem 4.19, or rather its proof, that $b^{\mathrm{w}}(x, h D)^{-1}$ exists and

$$
b^{\mathrm{w}}(x, h D)^{-1} r_{0}^{\mathrm{w}}(x, h D) b^{\mathrm{w}}(x, h D)^{-1}=O(1)_{L^{2} \rightarrow L^{2}}
$$

Thus

$$
\begin{aligned}
a^{\mathrm{w}}(x, h D)= & b^{\mathrm{w}}(x, h D) b^{\mathrm{w}}(x, h D)+r_{0}^{\mathrm{w}}(x, h D) \\
= & b^{\mathrm{w}}(x, h D) \\
& \left(1+b^{\mathrm{w}}(x, h D)^{-1} r_{0}^{\mathrm{w}}(x, h D) b^{\mathrm{w}}(x, h D)^{-1}\right) b^{\mathrm{w}}(x, h D) \\
= & b^{\mathrm{w}}(x, h D)\left(1+O(h)_{L^{2} \rightarrow L^{2}}\right) b^{\mathrm{w}}(x, h D) .
\end{aligned}
$$

Hence for sufficiently small $h>0$,

$$
\begin{aligned}
\left\langle\left(P(h)+(\lambda+\epsilon) \chi^{\mathrm{w}}-\lambda\right) u, u\right\rangle & =\left\langle a^{\mathrm{w}}(x, h D) u, u\right\rangle \\
& \geq\left\|b^{\mathrm{w}}(x, h D) u\right\|_{L^{2}}^{2}(1-O(h)) \\
& \geq \gamma\|u\|_{L^{2}}^{2}
\end{aligned}
$$

for some $\gamma>0$, since $b^{\mathrm{w}}(x, h D)^{-1}$ exists. This proves (6.23).
3. Claim \#2: For each $\delta>0$, there exists a bounded linear operator $Q$ such that

$$
\begin{equation*}
\chi^{\mathrm{w}}(x, h D)=Q+O\left(h^{\infty}\right)_{L^{2} \rightarrow L^{2}} \tag{6.24}
\end{equation*}
$$

and

$$
\begin{equation*}
\operatorname{rank}(Q) \leq \frac{1}{(2 \pi h)^{n}}(|\{p \leq \lambda+2 \epsilon\}|+\delta) \tag{6.25}
\end{equation*}
$$

To prove this, cover the set $\{p \leq \lambda+2 \epsilon\}$ with balls

$$
B_{j}:=B\left(\left(x_{j}, \xi_{j}\right), r_{j}\right) \quad(j=1, \cdots, N)
$$

such that

$$
\sum_{j=1}^{N}\left|B_{j}\right| \leq|\{p \leq \lambda+2 \epsilon\}|+\frac{\delta}{2}
$$

Define the "shifted" harmonic oscillator

$$
P_{j}(h):=\left|h D_{x}-\xi_{j}\right|^{2}+\left|x-x_{j}\right|^{2} ;
$$

and set

$$
\begin{aligned}
\Pi:= & \text { orthogonal projection in } L^{2} \text { onto } V \text {, the span of } \\
& \left\{u \mid P_{j}(h) u=E_{j}(h) u, E_{j}(h) \leq r_{j}, j=1, \ldots, N\right\} .
\end{aligned}
$$

We now claim that

$$
\begin{equation*}
(I-\Pi) \chi^{\mathrm{w}}(x, h D)=O\left(h^{\infty}\right)_{L^{2} \rightarrow L^{2}} . \tag{6.26}
\end{equation*}
$$

To see this, let $\chi=\sum_{j=1}^{N} \chi_{j}$, where spt $\chi_{j} \subset \subset B\left(\left(x_{j}, \xi_{j}\right), r_{j}\right)$, and put

$$
\begin{aligned}
\Pi_{j}:= & \text { orthogonal projection in } L^{2} \text { onto the span of } \\
& \left\{u \mid P_{j}(h) u=E_{j}(h) u, E_{j}(h) \leq r_{j}\right\} .
\end{aligned}
$$

Theorem 6.6 shows that $\left(I-\Pi_{j}\right) \chi_{j}^{\mathrm{w}}(x, h D)=O\left(h^{\infty}\right)$. We note that $\Pi \Pi_{j}=\Pi_{j}$ and hence

$$
\begin{aligned}
(I-\Pi) \chi^{\mathrm{w}}(x, h D) & =\sum_{j=1}^{N}(I-\Pi) \chi_{j}^{\mathrm{w}}(x, h D) \\
& =\sum_{j=1}^{N}(I-\Pi)\left(I-\Pi_{j}\right) \chi_{j}^{\mathrm{w}}(x, h D) \\
& =O\left(h^{\infty}\right)_{L^{2} \rightarrow L^{2}}
\end{aligned}
$$

This proves (6.26).
It now follows that

$$
\chi^{\mathrm{w}}(x, h D)=\Pi \chi^{\mathrm{w}}(x, h D)+(I-\Pi) \chi^{\mathrm{w}}(x, h D)=Q+O\left(h^{\infty}\right)
$$

for

$$
Q:=\Pi \chi^{\mathrm{w}}(x, h D)
$$

Clearly $Q$ has finite rank, since

$$
\begin{aligned}
\operatorname{rank} Q & =\operatorname{dim}(\text { image of } Q) \leq \operatorname{dim}(\text { image of } \Pi) \\
& \leq \sum_{j=1}^{N} \#\left\{E_{j}(h) \mid E_{j}(h) \leq r_{j}\right\} \\
& =\frac{1}{(2 \pi h)^{n}}\left(\sum_{j=1}^{N}\left|B_{j}\right|+o(1)\right),
\end{aligned}
$$

according to Weyl's law for the harmonic oscillator, Theorem 6.3. Consequently

$$
\begin{equation*}
\operatorname{rank} Q \leq \frac{1}{(2 \pi h)^{n}}\left(|\{p \leq \lambda+2 \epsilon\}|+\frac{\delta}{2}+o(1)\right) . \tag{6.27}
\end{equation*}
$$

This proves Claim \#2.
4. We next employ Claims \#1,2 and Theorem C.8. We have

$$
\begin{aligned}
\langle P(h) u, u\rangle & \geq(\lambda+\gamma)\|u\|_{L^{2}}^{2}-(\lambda+\epsilon)\langle Q u, u\rangle+\left\langle O\left(h^{\infty}\right) u, u\right\rangle \\
& \geq \lambda\|u\|_{L^{2}}^{2}-(\lambda+\epsilon)\langle Q u, u\rangle,
\end{aligned}
$$

where the rank of $Q$ is bound by (6.27). Theorem C.8,(i) implies then that

$$
N(\lambda) \leq \frac{1}{(2 \pi h)^{n}}(|\{p \leq \lambda+2 \epsilon\}|+\delta+o(1))
$$

This holds for all $\epsilon, \delta>0$, and so

$$
\begin{equation*}
N(\lambda) \leq \frac{1}{(2 \pi h)^{n}}(|\{p \leq \lambda\}|+o(1)) \tag{6.28}
\end{equation*}
$$

as $h \rightarrow 0$.
5. We must prove the opposite inequality.

Claim \#3: Suppose $B_{j}=B\left(\left(x_{j}, \xi_{j}\right), r_{j}\right) \subset\{p<\lambda\}$. Then if

$$
P_{j}(h) u=E_{j}(h) u
$$

and $E(h) \leq r_{j}$, we have

$$
\begin{equation*}
\langle P(h) u, u\rangle \leq\left(\lambda+\epsilon+O\left(h^{\infty}\right)\right)\|u\|_{L^{2}}^{2} \tag{6.29}
\end{equation*}
$$

To prove this claim, select a symbol $a \in C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{2 n}\right)$, with

$$
a \equiv 1 \text { on }\{p \leq \lambda\}, \operatorname{spt}(a) \subset\{p \leq \lambda+\epsilon\} .
$$

Let $c:=1-a$. Then $u-a^{\mathrm{w}}(x, h D) u=c^{\mathrm{w}}(x, h D) u=O\left(h^{\infty}\right)$ according to Theorem 6.6, since $\operatorname{spt}(1-a) \cap B_{j}=\emptyset$.

Define $b^{\mathrm{w}}:=P(h) a^{\mathrm{w}}(x, h D)$. Now $p \in S(m)$ and $a \in S\left(m^{-1}\right)$. Thus $b=p a+O(h) \in S$ and so $b^{\mathrm{w}}$ is bounded in $L^{2}$. Observe also that $b \leq \lambda+\frac{\epsilon}{2}$, and so

$$
b^{\mathrm{w}}(x, h D) \leq \lambda+\frac{3 \epsilon}{4} .
$$

Therefore

$$
\left\langle P(h) a^{\mathrm{w}}(x, h D) u, u\right\rangle=\left\langle b^{\mathrm{w}}(x, h D) u, u\right\rangle \leq\left(\lambda+\frac{3 \epsilon}{4}\right)\|u\|_{L^{2}}^{2} .
$$

Since $a^{\mathrm{w}}(x, h D) u=u+O\left(h^{\infty}\right)$, we deduce

$$
\langle P(h) u, u\rangle \leq\left(\lambda+\epsilon+O\left(h^{\infty}\right)\right)\|u\|_{L^{2}}^{2} .
$$

This proves Claim \#3.
6. Now find disjoint balls $B_{j} \subset\{p<\lambda\}$ such that

$$
|\{p<\lambda\}| \leq \sum_{j=1}^{N}\left|B_{j}\right|+\delta .
$$

Let $V:=\operatorname{span}\left\{u \mid P_{j}(h) u=E_{j}(h) u, E_{j}(h) \leq r_{j}, j=1, \ldots, N\right\}$. Owing to Claim \#3,

$$
\langle P u, u\rangle \leq(\lambda+\delta)\|u\|_{L^{2}}^{2}
$$

for all $u \in V$. Also, Theorem 6.3 implies

$$
\begin{aligned}
\operatorname{dim} V & \geq \sum_{j=1}^{N} \#\left\{E_{j}(h) \leq r_{j}\right\} \\
& =\frac{1}{(2 \pi h)^{n}}\left(\sum_{j=1}^{N}\left|B_{j}\right|+o(1)\right) \\
& \geq \frac{1}{(2 \pi h)^{n}}(|\{p<\lambda\}|-\delta+o(1)) .
\end{aligned}
$$

According then to Theorem C.8,(ii),

$$
N(\lambda) \geq \frac{1}{(2 \pi h)^{n}}(|\{p<\lambda\}|-\delta+o(1))
$$

## 7. Exponential estimates for eigenfunctions

7.1 Classically forbidden regions
7.2 Tunneling
7.3 Order of vanishing

This chapter continues our study of semiclassical behavior of eigenfunctions:

$$
\begin{equation*}
P(h) u(h)=E(h) u(h) \tag{7.1}
\end{equation*}
$$

for the operator

$$
P(h)=-h^{2} \Delta+V(x)
$$

and corresponding symbol

$$
p(x, \xi)=|\xi|^{2}+V(x)
$$

We first demonstrate that if $E(h)$ is close to the energy level $E$, then $u(h)$ exponentially small within the classically forbidden region

$$
V^{-1}(E, \infty)=\left\{x \in \mathbb{R}^{n} \mid V(x)>E\right\}
$$

Then we show, conversely, that in any open set the $L^{2}$ norm of $u(h)$ is bounded from below by a quantity exponentially small in $h$. We conclude with a discussion of the order of vanishing of eigenfunctions in the semiclassical limit.

### 7.1 CLASSICALLY FORBIDDEN REGIONS

We begin with some definitions and general facts.
DEFINITION. Let $U \subset \mathbb{R}^{n}$ be an open set. The semiclassical Sobolev norms are defined as

$$
\|u\|_{H_{h}^{k}(U)}:=\left(\sum_{|\alpha| \leq k} \int_{U}\left|(h D)^{\alpha} u\right|^{2} d x\right)^{1 / 2}
$$

for $u \in C^{\infty}(U), k=0,1, \ldots$
These differ from the standard Sobolev norms by the introduction of appropriate powers of $h$.

LEMMA 7.1 (Semiclassical elliptic estimates). Let $W \subset \subset U$ be open sets. Then there exists a constant $C$ such that

$$
\begin{equation*}
\|u\|_{H_{h}^{2}(W)} \leq C\left(\|P(h) u\|_{L^{2}(U)}+\|u\|_{L^{2}(U)}\right) \tag{7.2}
\end{equation*}
$$

for all $u \in C^{\infty}(U)$.

Proof. 1. Let $\chi \in C_{\mathrm{c}}^{\infty}(U), \chi \equiv 1$ on $W$. We multiply $P(h) u$ by $\chi^{2} \bar{u}$ and integrate by parts:

$$
\int_{U} h^{2}\left\langle\partial\left(\chi^{2} \bar{u}\right), \partial u\right\rangle+(V-E)|u|^{2} \chi^{2} d x=\int_{U} P(h) u \bar{u} \chi^{2} d x
$$

Therefore

$$
h^{2} \int_{U} \chi^{2}|\partial u|^{2} d x \leq C \int_{U}|P(h) u|^{2}+|u|^{2} d x
$$

and so

$$
h^{2} \int_{W}|\partial u|^{2} d x \leq C \int_{U}|P(h) u|^{2}+|u|^{2} d x
$$

2. Similarly, multiply $P(h) u$ by $\chi^{2} \Delta \bar{u}$ and integrate by parts, to deduce

$$
h^{4} \int_{W}\left|\partial^{2} u\right|^{2} d x \leq C \int_{U}|P(h) u|^{2}+|u|^{2} d x .
$$

Before turning again to eigenfunctions, we present the following general estimates. Our primary tool will be properly designed conjugations of the operator $P(h)$.
DEFINITION Given $\varphi \in C^{\infty}\left(\mathbb{R}^{n}\right)$, we define the conjugation of $P(h)$ by $e^{\varphi / h}$ :

$$
\begin{equation*}
P_{\varphi}(h):=e^{\varphi / h} P e^{-\varphi / h} . \tag{7.3}
\end{equation*}
$$

LEMMA 7.2 (Symbol of conjugation). We have

$$
\begin{equation*}
P_{\varphi}(h)=\left(p_{\varphi}\right)^{\mathrm{w}}+O(h) \tag{7.4}
\end{equation*}
$$

for the symbol

$$
\begin{equation*}
p_{\varphi}(x, \xi):=|\xi+i \partial \varphi(x)|^{2}+V(x) . \tag{7.5}
\end{equation*}
$$

Proof. We calculate for functions $u \in C^{\infty}\left(\mathbb{R}^{n}\right)$ that

$$
\begin{aligned}
& P_{\varphi}(h) u=e^{\varphi / h}\left(-h^{2} \Delta+V\right)\left(e^{-\varphi / h} u\right) \\
&=-h^{2} \Delta u+2 h\langle\partial \varphi, \partial u\rangle-|\partial \varphi|^{2} u+V u+h \Delta \varphi u
\end{aligned}
$$

The expression on the right is $\left(p_{\varphi}\right)^{\mathrm{w}} u+O(h) u$.
THEOREM 7.3 (Exponential estimate from above). Suppose that $U$ is an open set such that

$$
U \subset \subset V^{-1}(E, \infty)
$$

For each open set $W \supset \supset U$ and for each $\lambda$ near $E$, there exist constants $h_{0}, \delta, C>0$, such that

$$
\begin{equation*}
\|u\|_{L^{2}(U)} \leq C e^{-\delta / h}\|u\|_{L^{2}(W)}+C\|(P(h)-\lambda) u\|_{L^{2}(W)} \tag{7.6}
\end{equation*}
$$

for $u \in C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{n}\right)$ and $0<h \leq h_{0}$.
We call (7.6) an Agmon-type estimate.
Proof. 1. Select $\psi, \varphi \in C_{\mathrm{c}}^{\infty}(W)$ such that $0 \leq \psi, \varphi \leq 1, \psi \equiv 1$ on $U$, and $\varphi \equiv 1$ on spt $\psi$. We may assume as well that $W \subset \subset V^{-1}(E, \infty)$.

As in Lemma 7.2, we observe that the symbol of

$$
A(h):=e^{\delta \psi / h}(P(h)-\lambda) e^{-\delta \psi / h}
$$

is

$$
|\xi+i \delta \partial \psi|^{2}+V-\lambda+O(h) .
$$

Now for $\lambda$ close to $E, x \in W$ and $\delta$ sufficiently small, we have

$$
\left(|\xi+i \delta \partial \psi|^{2}+V-\lambda\right)^{2} \geq \gamma>0
$$

for some positive constant $\gamma$. Then according to the sharp Gårding inequality, Theorem 4.21, we see that provided $\delta>0$ is sufficiently small,

$$
\varphi A(h)^{*} A(h) \varphi \geq \sigma^{2} \varphi^{2}-O_{L^{2} \rightarrow L^{2}}(h),
$$

for some constant $\sigma>0$. Hence if $h$ is small enough, we have

$$
\varphi A(h)^{*} A(h) \varphi \geq \frac{1}{2} \sigma^{2} \varphi^{2}
$$

in the sense of operators.
2. This implies that

$$
\begin{aligned}
\left\|e^{\delta \psi / h} \varphi u\right\|_{L^{2}} & \leq C\left\|A(h)\left(e^{\delta \psi / h} \varphi u\right)\right\|_{L^{2}}=C\left\|e^{\delta \psi / h}(P(h)-\lambda) \varphi u\right\|_{L^{2}} \\
& \leq C\left\|e^{\delta \psi / h} \varphi(P(h)-\lambda) u\right\|_{L^{2}}+C\left\|e^{\delta \psi / h}[P(h), \varphi] u\right\|_{L^{2}}
\end{aligned}
$$

for $u \in C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{n}\right)$.
Next is the key observation that since $\varphi \equiv 1$ on spt $\psi$, we have $\psi \equiv 0$ on spt $[P(h), \varphi] u$. Thus Lemma 7.1 implies

$$
\begin{aligned}
\left\|e^{\delta \psi / h}[P(h), \varphi] u\right\|_{L^{2}} & =\|[P(h), \varphi] u\|_{L^{2}} \\
& \leq C\left(\left\|h D_{x} u\right\|_{L^{2}(W)}+\|u\|_{L^{2}(W)}\right) \\
& \leq C\|u\|_{L^{2}(W)}+C\|(P(h)-\lambda) u\|_{L^{2}(W)} .
\end{aligned}
$$

Combining these estimates, we conclude that

$$
\begin{aligned}
& e^{\delta / h}\|u\|_{L^{2}(U)} \leq\left\|e^{\delta \psi / h} u\right\|_{L^{2}} \\
& \quad \leq C\|u\|_{L^{2}(W)}+C\left(e^{\delta / h}+1\right)\|(P(h)-\lambda) u\|_{L^{2}(W)} .
\end{aligned}
$$

Specializing to eigenfunctions, we deduce
THEOREM 7.4 (Exponential decay estimates). Suppose that $U \subset \subset V^{-1}(E, \infty)$, and that $u(h) \in L^{2}\left(\mathbb{R}^{n}\right)$ solves

$$
P(h) u(h)=E(h) u(h),
$$

where

$$
E(h) \rightarrow E \quad \text { as } h \rightarrow 0
$$

Then there exists a constant $\delta>0$ such that

$$
\begin{equation*}
\|u(h)\|_{L^{2}(U)} \leq e^{-\delta / h}\|u(h)\|_{L^{2}\left(\mathbb{R}^{n}\right)} . \tag{7.7}
\end{equation*}
$$

### 7.2 TUNNELING

We continue to assume in this section $u=u(h)$ solves the eigenvalue problem (7.1).

In the previous section we showed that $u(h)$ is exponentially small in the physically forbidden region. In this section we will show that it can never be smaller than this: for small $h>0$ and any bounded, open subset $U$ of $\mathbb{R}^{n}$, we have the lower bound

$$
\|u\|_{L^{2}(U)} \geq e^{-\frac{C}{h}}\|u\|_{L^{2}\left(\mathbb{R}^{n}\right)}
$$

This is a mathematical version of quantum mechanical "tunneling into the physically forbidden region".

DEFINITION. Hörmander's hypoellipticity condition is the requirement for the symbol $p_{\varphi}$, defined by (7.5), that

$$
\begin{equation*}
\text { if } p_{\varphi}=0, \quad \text { then } i\left\{p_{\varphi}, \overline{p_{\varphi}}\right\}>0 \tag{7.8}
\end{equation*}
$$

Observe that for any complex function $q=q(x, \xi)$,

$$
i\{q, \bar{q}\}=i\{\operatorname{Re} q+i \operatorname{Im} q, \operatorname{Re} q-i \operatorname{Im} q\}=2\{\operatorname{Re} q, \operatorname{Im} q\}
$$

Hence the expression $i\left\{p_{\varphi}, \overline{p_{\varphi}}\right\}$ is real.

THEOREM $7.5\left(\boldsymbol{L}^{2}\right.$-estimate for $\boldsymbol{P}_{\boldsymbol{\varphi}}(\boldsymbol{h})$ ). If Hörmander's condition (7.8) is valid within $\bar{W} \subset \subset \mathbb{R}^{n}$, then

$$
\begin{equation*}
h^{1 / 2}\|u\|_{L^{2}(W)} \leq C\left\|P_{\varphi}(h) u\right\|_{L^{2}(W)} \tag{7.9}
\end{equation*}
$$

for all $u \in C_{\mathrm{c}}^{\infty}(W)$, provided $0<h \leq h_{0}$ for $h_{0}>0$ sufficiently small.

Proof. We calculate

$$
\begin{aligned}
\left\|P_{\varphi}(h) u\right\|_{L^{2}}^{2} & =\left\langle P_{\varphi}(h) u, P_{\varphi}(h) u\right\rangle=\left\langle P_{\varphi}^{*}(h) P_{\varphi} u, u\right\rangle \\
& =\left\langle P_{\varphi}(h) P_{\varphi}^{*}(h) u, u\right\rangle+\left\langle\left[P_{\varphi}^{*}(h), P_{\varphi}(h)\right] u, u\right\rangle \\
& =\left\|P_{\varphi}^{*}(h) u\right\|_{L^{2}}+\left\langle\left[P_{\varphi}^{*}(h), P_{\varphi}(h)\right] u, u\right\rangle .
\end{aligned}
$$

The idea will be to use the positivity of the second term on the right hand side wherever $P_{\varphi}^{*}(h)$ fails to be elliptic. More precisely, for any $M>1$ and $h$ small enough the calculation above gives

$$
\begin{aligned}
\left\|P_{\varphi}(h) u\right\|_{L^{2}}^{2} & \geq M h\left\|P_{\varphi}^{*}(h) u\right\|_{L^{2}}+\left\langle\left[P_{\varphi}^{*}(h), P_{\varphi}(h)\right] u, u\right\rangle \\
& =h\left\langle\left(M\left|p_{\varphi}\right|^{2}+i\left\{p_{\varphi}, \bar{p}_{\varphi}\right\}\right)^{\mathrm{w}} u, u\right\rangle-O\left(h^{2}\right)\|u\|_{H_{h}^{2}}^{2},
\end{aligned}
$$

the last term resulting from estimates of the lower order terms in $\bar{p}_{\varphi} \# p_{\varphi}$ and the commutator. Hörmander's hypoellipticity condition (7.8) implies for $M$ large enough that

$$
M\left|p_{\varphi}(x, \xi)\right|^{2}+i\left\{p_{\varphi}, \bar{p}_{\varphi}\right\}(x, \xi) \geq \gamma>0 .
$$

for $x \in \bar{W}$. Then Theorem 4.20, the easy Gårding inequality, and Lemma 7.1 show us that

$$
\left\|P_{\varphi}(h) u\right\|_{L^{2}}^{2} \geq C h\|u\|_{L^{2}}^{2}-O\left(h^{2}\right)\left(\left\|P_{\varphi}(h) u\right\|_{L^{2}}^{2}+\|u\|_{L^{2}}^{2}\right) .
$$

Next we carefully design a weight $\varphi$, to ensure that $P_{\varphi}(h)$ satisfies the hypothesis of Theorem 7.5.

LEMMA 7.6 (Constructing a weight). Let $0<r<R$. There exists a nonincreasing radial function $\varphi \in C^{\infty}\left(\mathbb{R}^{n}\right)$ such that Hörmander's hypoellipticity condition (7.8) holds on $B(0, R)-B(0, r)$.

Proof. 1. Recall that

$$
p_{\varphi}=|\xi+i \partial \varphi|^{2}+V-E=|\xi|^{2}+2 i\langle\xi, \partial \varphi\rangle-|\partial \varphi|^{2}+V-E .
$$

So $p_{\varphi}=0$ implies both

$$
\begin{equation*}
|\xi|^{2}-|\partial \varphi|^{2}+V-E=0 \tag{7.10}
\end{equation*}
$$

and

$$
\begin{equation*}
\langle\xi, \partial \varphi\rangle=0 . \tag{7.11}
\end{equation*}
$$

Furthermore,

$$
\begin{align*}
\frac{i}{2}\left\{p_{\varphi}, \overline{p_{\varphi}}\right\}= & \left\{\operatorname{Re} p_{\varphi}, \operatorname{Im} p_{\varphi}\right\} \\
= & \left\langle\partial_{\xi}\left(|\xi|^{2}-|\partial \varphi|^{2}+V-E\right), 2 \partial_{x}\langle\xi, \partial \varphi\rangle\right\rangle \\
& -\left\langle\partial_{x}\left(|\xi|^{2}-|\partial \varphi|^{2}+V-E\right), 2 \partial_{\xi}\langle\xi, \partial \varphi\rangle\right\rangle  \tag{7.12}\\
= & 4\left\langle\partial^{2} \varphi \xi, \xi\right\rangle+4\left\langle\partial^{2} \varphi \partial \varphi, \partial \varphi\right\rangle-2\langle\partial V, \partial \varphi\rangle .
\end{align*}
$$

2. Assume now

$$
\varphi=e^{\lambda \psi}
$$

where $\lambda>0$ will be selected and $\psi: \mathbb{R}^{n} \rightarrow \mathbb{R}$ is positive and radial, $\psi=\psi(|x|)$. Then

$$
\partial \varphi=\lambda \partial \psi e^{\lambda \psi}
$$

and

$$
\partial^{2} \varphi=\left(\lambda^{2} \partial \psi \otimes \partial \psi+\lambda \partial^{2} \psi\right) e^{\lambda \psi}
$$

Hence

$$
\left\langle\partial^{2} \varphi \xi, \xi\right\rangle=\left(\lambda^{2}\langle\partial \psi, \xi\rangle^{2}+\lambda\left\langle\partial^{2} \psi \xi, \xi\right\rangle\right) e^{\lambda \psi}=\lambda\left\langle\partial^{2} \psi \xi, \xi\right\rangle e^{\lambda \psi}
$$

since (7.11) implies $\langle\partial \psi, \xi\rangle=0$. Also

$$
\left\langle\partial^{2} \varphi \partial \varphi, \partial \varphi\right\rangle=\lambda^{4}|\partial \psi|^{4} e^{3 \lambda \psi}+\lambda^{3}\left\langle\partial^{2} \psi \partial \psi, \partial \psi\right\rangle e^{3 \lambda \psi}
$$

and

$$
\langle\partial V, \partial \varphi\rangle=\lambda\langle\partial V, \partial \psi\rangle e^{\lambda \psi}
$$

According to (7.12), we have

$$
\begin{align*}
\frac{i}{2}\left\{p_{\varphi}, \overline{p_{\varphi}}\right\}= & 4 \lambda\left\langle\partial^{2} \psi \xi, \xi\right\rangle e^{\lambda \psi}+4 \lambda^{4}|\partial \psi|^{4} e^{3 \lambda \psi}  \tag{7.13}\\
& +4 \lambda^{3}\left\langle\partial^{2} \psi \partial \psi, \partial \psi\right\rangle e^{3 \lambda \psi}-2 \lambda\langle\partial V, \partial \psi\rangle e^{\lambda \psi}
\end{align*}
$$

3. Now take

$$
\psi:=\mu-|x|,
$$

for a constant $\mu$ so large that $\psi \geq 1$ on $B(0, R)$. Then $\varphi$ is radial and nonincreasing. Furthermore

$$
|\partial \psi|=1,\left|\partial^{2} \psi\right| \leq C \quad \text { on } B(0, R)-B(0, r)
$$

Owing to (7.10) we have

$$
|\xi|^{2} \leq C+|\partial \varphi|^{2} \leq C+C \lambda^{2} e^{2 \lambda \psi} \quad \text { on } B(0, R)-B(0, r)
$$

Plugging these estimates into (7.13), we compute

$$
\frac{i}{2}\left\{p_{\varphi}, \overline{p_{\varphi}}\right\} \geq 2 \lambda^{4} e^{3 \lambda \psi}-C \lambda^{3} e^{3 \lambda \psi}-C \geq 1
$$

in $B(0, R)-B(0, r)$, if $\lambda$ is selected large enough.

Lastly, we modify $\psi$ within $B(0, r)$ to obtain a smooth fuction on $B(0, R)$.

THEOREM 7.7 (Exponential estimate from below). Let $a<b$ and suppose $U \subset \mathbb{R}^{n}$ is an open set. There exist constants $C, h_{0}>0$ such that if $u(h)$ solves

$$
P(h) u=E(h) u(h) \quad \text { in } \mathbb{R}^{n}
$$

for $E(h) \in[a, b]$ and $0<h \leq h_{0}$, then

$$
\begin{equation*}
\|u(h)\|_{L^{2}(U)} \geq e^{-\frac{C}{h}}\|u(h)\|_{L^{2}\left(\mathbb{R}^{n}\right)} \tag{7.14}
\end{equation*}
$$

We call (7.14) a Carleman-type estimate.

Proof. 1. We may assume without loss that $U=B(0,3 r)$ for some $0<r<\frac{1}{3}$. Select $R>1$ so large that

$$
p(x, \xi)-\lambda=|\xi|^{2}+V(x)-\lambda \geq|\xi|^{2}+\langle x\rangle^{k}
$$

for $|x| \geq R$ and $a \leq \lambda \leq b$. Since $p-E(h)$ is therefore elliptic on $\mathbb{R}^{n}-B(0, R)$, we have the estimate

$$
\begin{equation*}
\|v\|_{L^{2}\left(\mathbb{R}^{n}-B(0, R)\right)} \leq C\|(P(h)-E(h)) v\|_{L^{2}\left(\mathbb{R}^{n}-B(0, R)\right)} \tag{7.15}
\end{equation*}
$$

for all $v \in C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{n}-B(0, R)\right)$.
2. Select two radial functions $\chi_{1}, \chi_{2}: \mathbb{R}^{n} \rightarrow \mathbb{R}$ such that $0 \leq \chi_{1} \leq 1$ and

$$
\begin{cases}\chi_{1} \equiv 0 & \text { on } B(0, r) \\ \chi_{1} \equiv 1 & \text { on } B(0, R+2)-B(0,2 r) \\ \chi_{1} \equiv 0 & \text { on } \mathbb{R}^{2 n}-B(0, R+3)\end{cases}
$$

and $0 \leq \chi_{2} \leq 1$,

$$
\begin{cases}\chi_{2} \equiv 0 & \text { on } B(0, R) \\ \chi_{2} \equiv 1 & \text { on } \mathbb{R}^{2 n}-B(0, R+1)\end{cases}
$$

Applying (7.15) to $v=\chi_{2} u$ gives

$$
\left\|\chi_{2} u\right\|_{L^{2}} \leq C\left\|(P(h)-E(h))\left(\chi_{2} u\right)\right\|_{L^{2}}=C\left\|\left[P(h), \chi_{2}\right] u\right\|_{L^{2}} .
$$

Now $\left[P(h), \chi_{2}\right] u=-h^{2} u \Delta \chi_{2}-2 h^{2}\left\langle\partial \chi_{2}, \partial u\right\rangle$, and consequently $\left[P(h), \chi_{2}\right] u$ is supported within $B(0, R+1)-B(0, R)$. Hence Lemma 7.1 implies

$$
\begin{aligned}
&\left\|\left[P(h), \chi_{2}\right] u\right\|_{L^{2}} \leq C h\|u\|_{H_{h}^{1}(B(0, R+1)-B(0, R))} \\
& \leq C h\left(\|P(h) u\|_{L^{2}(B(0, R+2)-B(0, R-1))}\right. \\
&\left.\quad \quad+\|u\|_{L^{2}(B(0, R+2)-B(0, R-1))}\right) \\
& \leq C h\|u\|_{L^{2}(B(0, R+2)-B(0, R-1))} \\
& \leq C h\left\|\chi_{1} u\right\|_{L^{2}}
\end{aligned}
$$

Therefore

$$
\begin{equation*}
\left\|\chi_{2} u\right\|_{L^{2}} \leq C h\left\|\chi_{1} u\right\|_{L^{2}} . \tag{7.16}
\end{equation*}
$$

3. Next apply Theorem 7.5:
$h^{1 / 2}\left\|e^{\frac{\varphi}{h}} \chi_{1} u\right\|_{L^{2}} \leq C\left\|e^{\frac{\varphi}{h}}(P(h)-E(h))\left(\chi_{1} u\right)\right\|_{L^{2}}=C\left\|e^{\frac{\varphi}{h}}\left[P(h), \chi_{1}\right] u\right\|_{L^{2}}$
Now $\left[P(h), \chi_{1}\right]$ is supported within the union of $B(0,2 r)-B(0, r)$ and $B(0, R+3)-B(0, R+2)$. Since $\varphi$ is nonincreasing, we therefore have

$$
\begin{aligned}
&\left\|e^{\frac{\varphi}{h}}\left[P(h), \chi_{1}\right] u\right\|_{L^{2}} \leq C h e^{\frac{\varphi(R+2)}{h}}\left\|\chi_{2} u\right\|_{\left.\left.H_{h}^{1}(B(0, R+3)-B(0, R+2))\right)\right)} \\
&+C h e^{\frac{\varphi(0)}{h}}\|u\|_{H_{h}^{1}(B(0,2 r)} .
\end{aligned}
$$

The right hand sides can be estimated by Lemma 7.1. This gives

$$
\begin{equation*}
\left\|e^{\frac{\varphi}{h}} \chi_{1} u\right\|_{L^{2}} \leq C h^{1 / 2} e^{\frac{\varphi(R+2)}{h}}\left\|\chi_{2} u\right\|_{L^{2}}+C h^{1 / 2} e^{\frac{\varphi(0)}{h}}\|u\|_{L^{2}(U)} \tag{7.17}
\end{equation*}
$$

4. Select a constant $A>0$ so that

$$
\begin{cases}\varphi>A & \text { on } B(0, R+1)-B(0, R) \\ \varphi<A & \text { on } B(0, R+3)-B(0, R+2)\end{cases}
$$

Multiply (7.16) by $e^{A / h}$ and add to (7.17):

$$
\begin{aligned}
& \left\|e^{\frac{A}{h}} \chi_{2} u\right\|_{L^{2}}+\left\|e^{\frac{\varphi}{h}} \chi_{1} u\right\|_{L^{2}} \\
& \quad \leq C h\left\|e^{\frac{\varphi}{h}} \chi_{1} u\right\|_{L^{2}}+C h^{1 / 2}\left\|e^{\frac{A}{h}} \chi_{2} u\right\|_{L^{2}}+C h^{1 / 2} e^{\frac{\varphi(0)}{h}}\|u\|_{L^{2}(U)}
\end{aligned}
$$

Take $0<h \leq h_{0}$, for $h_{0}$ sufficiently small, to deduce

$$
\left\|e^{\frac{A}{h}} \chi_{2} u\right\|_{L^{2}}+\left\|e^{\frac{\varphi}{h}} \chi_{1} u\right\|_{L^{2}} \leq C h^{1 / 2} e^{\frac{\varphi(0)}{h}}\|u\|_{L^{2}(U)} .
$$

Since $\chi_{1}+\chi_{2} \geq 1$ on $\mathbb{R}^{n}-B(0,2 r)=\mathbb{R}^{n}-U$, the Theorem follows.

### 7.3 ORDER OF VANISHING

Assume, as usual, that

$$
\begin{equation*}
P(h) u(h)=E(h) u(h), \tag{7.18}
\end{equation*}
$$

where $E(h) \in[a, b]$. To simplify notation, we will in this subsection write $u$ for $u(h)$.

We propose now to give an estimate for the order of vanishing of $u$, following a suggestion of N. Burq.
DEFINITION. We say $u$ vanishes to order $N$ at the point $x_{0}$ if

$$
u(x)=O\left(\left|x-x_{0}\right|^{N}\right) \quad \text { as } x \rightarrow x_{0} .
$$

We will consider potentials which are analytic in $x$ and, to avoid technical difficulties, make a strong assumption on the growth of derivatives:

$$
\begin{equation*}
V(x) \geq\langle x\rangle^{m} / C_{0}-C_{0}, \quad\left|\partial^{\alpha} V(x)\right| \leq C_{0}^{1+|\alpha|}|\alpha|^{|\alpha|}\langle x\rangle^{m} \tag{7.19}
\end{equation*}
$$

for some $m>0$ and all multiindices $\alpha$.
We note that the second condition holds when $V$ has a holomorphic extension bounded by $|z|^{m}$ into a conic neighbourhood of $\mathbb{R}^{n}$ in $\mathbb{C}^{n}$.

THEOREM 7.8 (Semiclassical estimate on vanishing order). Suppose that $u \in L^{2}$ solves (7.18) for $a \leq E(h) \leq b$ and that $V$ a real analytic potential satisfying (7.19). Let $K$ be compact subset of $\mathbb{R}^{n}$.

Then there exists a constant $C$ such that if $u$ vanishes to order $N$ at a point $x_{0} \in K$, we have the estimate

$$
\begin{equation*}
N \leq C h^{-1} \tag{7.20}
\end{equation*}
$$

We need the following lemma to establish analyticity of the solution in a semiclassically quantitative way:

LEMMA 7.9 (Semiclassical derivative estimates). If $u$ satisfies the assumptions of Theorem 7.8, then there exists a constant $C_{1}$ such that for any positive integer $k$ :

$$
\begin{equation*}
\|u\|_{H_{h}^{k}\left(\mathbb{R}^{n}\right)} \leq C_{1}^{k}(1+k h)^{k}\|u\|_{L^{2}\left(\mathbb{R}^{n}\right)} . \tag{7.21}
\end{equation*}
$$

Proof. 1. By adding $C_{0}$ to $V$ we can assume without loss that $V(x) \geq$ $\langle x\rangle^{m} / C_{0}$. The Lemma will follow from the following stronger estimate, which we will prove by induction:

$$
\begin{align*}
\left\|\langle x\rangle^{m / 2}(h D)^{\alpha} u\right\|_{L^{2}} & +\left\|(h \partial)(h D)^{\alpha} u\right\|_{L^{2}} \\
& \leq C_{2}^{k+2}(1+k h)^{k+1}\|u\|_{L^{2}} . \tag{7.22}
\end{align*}
$$

for $|\alpha|=k$.
2. To prove this inequality, we observe first that our multiplying (7.18) by $\bar{u}$ and integrating by parts shows that estimate (7.22) holds for $|\alpha|=0$

Next, note that

$$
\begin{aligned}
\| V^{\frac{1}{2}} & (h D)^{\alpha} u\left\|_{L^{2}}^{2}+\right\|(h \partial)(h D)^{\alpha} u \|_{L^{2}}^{2} \\
& =\left\langle\left(-h^{2} \Delta+V-E(h)\right)(h D)^{\alpha} u,(h D)^{\alpha} u\right\rangle+E(h)\left\|(h D)^{\alpha} u\right\|_{L^{2}}^{2} \\
& =\left\langle V^{-\frac{1}{2}}\left[V,(h D)^{\alpha}\right] u, V^{\frac{1}{2}}(h D)^{\alpha} u\right\rangle+E(h)\left\|(h D)^{\alpha} u\right\|_{L^{2}}^{2} \\
& \leq 2\left\|V^{-\frac{1}{2}}\left[V,(h D)^{\alpha}\right] u\right\|_{L^{2}}^{2}+\frac{1}{2}\left\|V^{\frac{1}{2}}(h D)^{\alpha} u\right\|_{L^{2}}^{2}+E(h)\left\|(h D)^{\alpha} u\right\|_{L^{2}}^{2} .
\end{aligned}
$$

Hence

$$
\begin{align*}
\frac{1}{2}\left\|V^{\frac{1}{2}}(h D)^{\alpha} u\right\|_{L^{2}}^{2}+ & \left\|(h \partial)(h D)^{\alpha} u\right\|_{L^{2}}^{2}  \tag{7.23}\\
& \leq 2\left\|V^{-\frac{1}{2}}\left[V,(h D)^{\alpha}\right] u\right\|_{L^{2}}^{2}+E(h)\left\|(h D)^{\alpha} u\right\|_{L^{2}}^{2} .
\end{align*}
$$

3. We can now expand the commutator, to deduce from (7.19) (with $V$ replaced by $V+C_{0}$ ) the inequality:

$$
\begin{align*}
& \left\|V^{-\frac{1}{2}}\left[V,(h D)^{\alpha}\right] u\right\|_{L^{2}} \leq \\
& \quad \sum_{l=0}^{k-1}\binom{k}{l} C_{0}^{k-l+1}(h(k-l))^{k-l} \sup _{|\beta|=l}\left\|\langle x\rangle^{m / 2}(h D)^{\beta} u\right\|_{L^{2}} . \tag{7.24}
\end{align*}
$$

We proceed by induction, and so now assume that (7.22) is valid for $|\alpha|<k$. Now Stirling's formula implies

$$
\binom{k}{l} \leq C \frac{k^{k}}{l^{l}(k-l)^{k-l}}
$$

Hence, in view of (7.23) and (7.24), it is enough to show that there exists a constant $C_{2}$ such that

$$
\sum_{l=0}^{k-1} h^{k-l} \frac{k^{k}}{l^{l}} C_{0}^{k-l+1} C_{2}^{l+2}(1+l h)^{l+1}+C_{2}^{k+1}(1+h k)^{k} \leq C_{2}^{k+2}(1+h k)^{k+1}
$$

This estimate we can rewrite as

$$
\begin{gathered}
C_{0} \sum_{l=1}^{k-1}\left(\frac{C_{0}}{C_{2}}\right)^{k-l}(h l)^{-l}(1+h l)^{l}(1+h l)+C_{2}^{-1}(h k)^{-k}(1+h k)^{k} \\
\leq(h k)^{-k}(1+h k)^{k}(1+h k)
\end{gathered}
$$

Since we can choose $C_{2}$ to be large and since we can estimate the $(1+h l)$ factor in the sum by $(1+h k)$, this will follow once we show that for $\epsilon$ small enough,

$$
\sum_{l=0}^{k-1} \epsilon^{k-l} a_{l} \leq a_{k} \quad \text { for } \quad a_{l}:=\left(1+(h l)^{-1}\right)^{l}
$$

This is true by induction if $a_{k-1} / a_{k}$ is bounded:

$$
\sum_{l=0}^{k-1} \epsilon^{k-l} a_{l} \leq 2 \epsilon a_{k-1}
$$

In our case,

$$
\begin{aligned}
\frac{a_{k-1}}{a_{k}} & =\left(\frac{1+(h(k-1))^{-1}}{1+(h k)^{-1}}\right)^{k-1}\left(1+(h k)^{-1}\right)^{-1} \\
& =\left(1+\frac{1}{(k-1)(1+h k)}\right)^{k-1}\left(1+(h k)^{-1}\right)^{-1} \\
& \leq \exp \left(\frac{1}{1+h k}\right) \frac{h k}{1+h k}<1
\end{aligned}
$$

Proof of Theorem 7.8: Assume now that $\|u\|_{L^{2}}=1$ and that $u$ vanishes to order $N$ at a point $x_{0} \in K$.

Then $D^{\alpha} u\left(x_{0}\right)=0$ for $|\alpha|<N$ and Taylor's formula shows that

$$
\begin{equation*}
|u(x)| \leq \frac{\epsilon^{N}}{N!} \sup _{|\alpha|=N} \sup _{y \in \mathbb{R}^{n}}\left|D^{\alpha} u(y)\right| \quad \text { for }\left|x-x_{0}\right|<\epsilon \tag{7.25}
\end{equation*}
$$

The Sobolev inequality (Lemma 3.5) and Lemma 7.9 allow us to estimate the derivatives. If say $M=N+n$ and $|\alpha|=N$, then

$$
\sup _{y \in \mathbb{R}^{n}}\left|D^{\alpha} u(y)\right| \leq\|u\|_{H^{M}} \leq h^{-M}\|u\|_{H_{h}^{M}} \leq h^{-M} C_{1}^{M}(1+h M)^{M}
$$

Inserting this into (7.25) and using Stirling's formula, we deduce that if for $\left|x-x_{0}\right|<\epsilon$, then

$$
|u(x)| \leq\left(\frac{e \epsilon}{N}\right)^{N}\left(\frac{C}{h}\right)^{M}(1+h M)^{M} \leq\left(\frac{N}{e \epsilon}\right)^{n}\left(\frac{e \epsilon C}{N h}\right)^{M}(1+h M)^{M}
$$

If we put $A:=M h$, then for $\epsilon$ small enough we have, with $K=C \epsilon^{-1}$ large,

$$
\begin{aligned}
|u(x)| & \leq\left(K A h^{-1}\right)^{n}\left(\frac{1}{K A}\right)^{A h^{-1}}(1+A)^{A h^{-1}} \\
& =\left(K A h^{-1}\right)^{n}(1+1 / A)^{A h^{-1}} \exp \left(-A h^{-1} \log K\right)
\end{aligned}
$$

We can assume that $A$ is large, as otherwise there is nothing to prove. Hence

$$
|u(x)| \leq \exp \left(-\alpha A h^{-1}\right)
$$

for $\alpha>0$ and $\left|x-x_{0}\right|<\epsilon$. It follows that

$$
\int_{\left\{\left|x-x_{0}\right|<\epsilon\right\}}|u(x)|^{2} d x \leq C_{1} e^{-2 \alpha A / h}
$$

uniformly in $h$. But according to Theorem 7.7,

$$
\int_{\left\{\left|x-x_{0}\right|<\epsilon\right\}}|u(x)|^{2} d x>e^{-C_{2} / h}
$$

Consequently $A=M h=(N+n) h$ is bounded, and this means that $N \leq C h^{-1}$, as claimed.

EXAMPLE : Optimal order of vanishing. Theorem 7.8 is optimal in the semiclassical limit, meaning as regards the dependence on $h$ in estimate (7.20).

We can see this by considering the harmonic oscillator in dimension $n=2$. In polar coordinates $(r, \theta)$ the harmonic oscillator for $h=1$ takes the form

$$
P_{0}=r^{-2}\left(\left(r D_{r}\right)^{2}+D_{\theta}^{2}+r^{4}\right) .
$$

The eigenspace corresponding to the eigenvalue $2 k+2$ has dimension $k+1$, corresponding to the number of multiindicies $\alpha=\left(\alpha_{1}, \alpha_{2}\right)$, with $|\alpha|=\alpha_{1}+\alpha_{2}=k$. Separating variables, we look for eigenfunctions of the form

$$
u=u_{k m}(r) e^{i m \theta}
$$

Then

$$
r^{-2}\left(\left(r D_{r}\right)^{2}+m^{2}+r^{4}-(2 n+1) r^{2}\right) u_{k m}(r)=0
$$

Since the number of linearly independent eigenfunctions is $k+1$, there must be solution for some integer $m>k / 2$. Near $r=0$, we have the asymptotics $u_{k m} \simeq r^{ \pm m}$, and the case $u_{k m} \simeq r^{-m}$ is impossible since $u \in L^{2}$. Therefore $u \simeq r^{m}$ has to vanish to order $m$.

Rescaling to the semiclassical case, we see that for the eigenvalue $E(h)=(2 k+1) h \simeq 1$ we have an eigenfunction vanishing to order $\simeq 1 / h$.

## 8. Quantum ergodicity

8.1 Classical ergodicity
8.2 Egorov's Theorem
8.3 Weyl's Theorem generalized
8.4 A quantum ergodic theorem

In this chapter we are given a smooth potential $V$ on a compact Riemannian manifold $(M, g)$ and write

$$
\begin{equation*}
p(x, \xi)=|\xi|_{g}^{2}+V(x) \tag{8.1}
\end{equation*}
$$

for $(x, \xi) \in T^{*} M$, the cotangent space of $M$. As explained in Appendix E , the associated quantum operator is

$$
\begin{equation*}
P(h)=-h^{2} \Delta_{g}+V, \tag{8.2}
\end{equation*}
$$

and the Hamiltonian flow generated by $p$ is denoted

$$
\varphi_{t}=\exp \left(t H_{p}\right) \quad(t \in \mathbb{R})
$$

We address in this chapter quantum implications of ergodicity for the classical evolution $\left\{\varphi_{t}\right\}_{t \in \mathbb{R}}$. The proofs will rely on various advanced material presented in Appendix E.

### 8.1 CLASSICAL ERGODICITY

We hereafter select $a<b$, and assume that

$$
\begin{equation*}
|\partial p| \geq \gamma>0 \text { on }\{a \leq p \leq b\} . \tag{8.3}
\end{equation*}
$$

According then to the Implicit Function Theorem, for each $a \leq c \leq b$, the set

$$
\Sigma_{c}:=p^{-1}(c)
$$

is a smooth, $2 n-1$ dimensional hypersurface in the cotangent space $T^{*} M$. We can interpret $\Sigma_{c}$ as an energy surface.

NOTATION. For each $c \in[a, b]$, we denote by $\mu$ Liouville measure on the hypersurface $\Sigma_{c}=p^{-1}(c)$ corresponding to $p$. This measure is characterized by the formula

$$
\iint_{p^{-1}[a, b]} f d x d \xi=\int_{a}^{b} \int_{\Sigma_{c}} f d \mu d c
$$

for all $a<b$ and each continuous function $f: T^{*} M \rightarrow \mathbb{R}^{n}$.
DEFINITION. Let $m \in \Sigma_{c}$ and $f: T^{*} M \rightarrow \mathbb{C}$. For $T>0$ we define the time average

$$
\begin{equation*}
\langle f\rangle_{T}:=\frac{1}{T} \int_{0}^{T} f \circ \varphi_{t}(m) d t=f_{0}^{T} f \circ \varphi_{t}(m) d t \tag{8.4}
\end{equation*}
$$

the slash through the second integral denoting an average. Note carefully that $\langle f\rangle_{T}=\langle f\rangle_{T}(m)$ depends upon the starting point $m$.

DEFINITION. We say the flow $\varphi_{t}$ is ergodic on $p^{-1}[a, b]$ if for each $c \in[a, b]$,

$$
\left\{\begin{array}{l}
\text { if } E \subseteq \Sigma_{c} \text { is flow invariant, then }  \tag{8.5}\\
\text { either } \mu(E)=0 \text { or else } \mu(E)=\mu\left(\Sigma_{c}\right)
\end{array}\right.
$$

In other words, we are requiring that each flow invariant subset of the energy level $\Sigma_{c}$ have either zero measure or full measure.

THEOREM 8.1 (Mean Ergodic Theorem). Suppose the flow is ergodic on $\Sigma_{c}:=p^{-1}(c)$. Then for each $f \in L^{2}\left(\Sigma_{c}, \mu\right)$ we have

$$
\begin{equation*}
\lim _{T \rightarrow \infty} \int_{\Sigma_{c}}\left(\langle f\rangle_{T}-\int_{\Sigma_{c}} f d \mu\right)^{2} d \mu=0 \tag{8.6}
\end{equation*}
$$

REMARK. According to Birkhoff's Ergodic Theorem,

$$
\langle f\rangle_{T} \rightarrow \int_{\Sigma_{c}} f d \mu \quad \text { as } T \rightarrow \infty
$$

for $\mu$-a.e. point $m$ belonging to $\Sigma_{c}$, But we will only need the weaker statement of Theorem 8.1.

Proof. 1. Define

$$
\begin{gathered}
A:=\left\{f \in L^{2}\left(\Sigma_{c}, \mu\right) \mid \varphi_{t}^{*} f=f \text { for all times } t\right\}, \\
B_{0}:=\left\{H_{p} g \mid g \in C^{\infty}\left(\Sigma_{c}\right)\right\}, \quad B:=\bar{B}_{0} .
\end{gathered}
$$

We claim that

$$
\begin{equation*}
h \in B_{0}^{\perp} \text { if and only if } h \in A \tag{8.7}
\end{equation*}
$$

To see this, first let $h \in A$ and $f=H_{p} g \in B_{0}$. Then

$$
\begin{aligned}
\int_{\Sigma_{c}} h \bar{f} d \mu & =\int_{\Sigma_{c}} h \overline{H_{p} g} d \mu=\left.\frac{d}{d t} \int_{\Sigma_{c}} h \overline{\varphi_{t}^{*} g} d \mu\right|_{t=0} \\
& =\left.\frac{d}{d t} \int_{\Sigma_{c}} \varphi_{-t}^{*} h \bar{g} d \mu\right|_{t=0}=\left.\frac{d}{d t} \int_{\Sigma_{c}} h \bar{g} d \mu\right|_{t=0}=0
\end{aligned}
$$

and consequently $h \in B_{0}^{\perp}$.
Conversely, suppose $h \in B_{0}^{\perp}$. Then for any $g \in C^{\infty}$, we have

$$
0=\int_{\Sigma_{c}} h \overline{H_{p} \varphi_{-t}^{*} g} d \mu=\frac{d}{d t} \int_{\Sigma_{c}} h \overline{\varphi_{-t}^{*} g} d \mu=\frac{d}{d t} \int_{\Sigma_{c}} \varphi_{t}^{*} h \bar{g} d \mu .
$$

Therefore for all times $t$ and all functions $g$,

$$
\int_{\Sigma_{c}} \varphi_{t}^{*} h \bar{g} d \mu=\int_{\Sigma_{c}} h \bar{g} d \mu
$$

Hence $\varphi_{t}^{*} h \equiv h$, and so $h \in A$.
2. It follows from (8.7) that we have the orthogonal decomposition

$$
L^{2}\left(\Sigma_{c}, \mu\right)=A \oplus B
$$

Thus if we write $f=f_{A}+f_{B}$, for $f_{A} \in A, f_{B} \in B$, then

$$
\left\langle f_{A}\right\rangle_{T} \equiv f_{A}
$$

for all $T$.
Now suppose $f_{B}=H_{p} g \in B_{0}$. We can then compute

$$
\begin{aligned}
\int_{\Sigma_{c}}\left|\left\langle f_{B}\right\rangle_{T}\right|^{2} d \mu & =\frac{1}{T^{2}} \int_{\Sigma_{c}}\left|\int_{0}^{T}(d / d t) \varphi_{t}^{*} g d t\right|^{2} d \mu \\
& =\frac{1}{T^{2}} \int_{\Sigma_{c}}\left|\varphi_{T}^{*} g-g\right|^{2} d \mu \\
& \leq \frac{4}{T^{2}} \int_{\Sigma_{c}}|g|^{2} d \mu \longrightarrow 0
\end{aligned}
$$

as $T \rightarrow \infty$. Since $f_{B} \in B:=\bar{B}_{0}$, we have $\left\langle f_{B}\right\rangle_{T} \rightarrow 0$ in $L^{2}\left(\Sigma_{c}, d \mu\right)$.
3. The ergodicity hypothesis is equivalent to saying that $A$ consists of constant functions. Indeed, for any $h \in A$, the set $h^{-1}([\alpha, \infty))$ is invariant under the flow, and hence has either full measure or measure zero. Since the functions in $L^{2}\left(\Sigma_{c}, d \mu\right)$ are defined up to sets of measure zero, $h$ is equivalent to a constant function.

Lastly, observe that the orthogonal projection $f \mapsto f_{A}$ is just the space average with respect to $\mu$. This proves (8.6).

### 8.2 EGOROV'S THEOREM

We next estimate the difference between the classical and quantum evolutions governed by our symbol $p(x, \xi)=|\xi|^{2}+V(x)$.

NOTATION. (i) We write

$$
\begin{equation*}
e^{-\frac{i t P(h)}{h}} \quad(t \in \mathbb{R}) \tag{8.8}
\end{equation*}
$$

for the unitary group on $L^{2}(M)$ generated by the self-adjoint operator $P(h)$.

Note that since $P(h) u_{j}(h)=E_{j}(h) u_{j}(h)$, we have

$$
\begin{equation*}
e^{-\frac{i t P(h)}{h}} u_{j}(h)=e^{-\frac{i t E_{j}}{h}} u_{j}(h) \quad(t \in \mathbb{R}) . \tag{8.9}
\end{equation*}
$$

(ii) If $A$ is a symbol in $\Psi^{-\infty}$, we also write

$$
\begin{equation*}
A_{t}:=e^{\frac{i t P(h)}{h}} A e^{-\frac{i t P(h)}{h}} \quad(t \in \mathbb{R}) . \tag{8.10}
\end{equation*}
$$

THEOREM 8.2 (Weak form of Egorov's Theorem). Fix a time $T>0$ and define for $0 \leq t \leq T$

$$
\begin{equation*}
\tilde{A}_{t}:=\operatorname{Op}\left(a_{t}\right) \tag{8.11}
\end{equation*}
$$

where

$$
\begin{equation*}
a_{t}(x, \xi):=a\left(\varphi_{t}(x, \xi)\right) . \tag{8.12}
\end{equation*}
$$

Then

$$
\begin{equation*}
\left\|A_{t}-\tilde{A}_{t}\right\|_{L^{2} \rightarrow L^{2}}=O(h) \quad \text { uniformly for } 0 \leq t \leq T . \tag{8.13}
\end{equation*}
$$

Proof. We have

$$
\frac{d}{d t} a_{t}=\left\{p, a_{t}\right\} .
$$

Recall from Appendix E that $\sigma$ denotes the symbol of an operator. Then, since $\sigma\left(\frac{i}{h}[P(h), B]\right)=\{p, \sigma(B)\}$, it follows that

$$
\begin{equation*}
\frac{d}{d t} \tilde{A}_{t}=\frac{i}{h}\left[P(h), \tilde{A}_{t}\right]+E_{t} \tag{8.14}
\end{equation*}
$$

with an error term $\left\|E_{t}\right\|_{L^{2} \rightarrow L^{2}}=O(h)$. Hence

$$
\begin{aligned}
& \frac{d}{d t}\left(e^{-\frac{i t P(h)}{h}} \tilde{A}_{t} e^{\frac{i t P(h)}{h}}\right) \\
& \quad=e^{-\frac{i t P(h)}{h}}\left(\frac{d}{d t} \tilde{A}_{t}-\frac{i}{h}\left[P(h), \tilde{A}_{t}\right]\right) e^{\frac{i t P(h)}{h}} \\
& \quad=e^{-\frac{i t P(h)}{h}}\left(\frac{i}{h}\left[P(h), \tilde{A}_{t}\right]+E_{t}-\frac{i}{h}\left[P(h), \tilde{A}_{t}\right]\right) e^{\frac{i t P(h)}{h}} \\
& =e^{-\frac{i t P(h)}{h}} E_{t} e^{\frac{i t P(h)}{h}}=O(h) .
\end{aligned}
$$

Integrating, we deduce

$$
\left\|e^{-\frac{i t P(h)}{h}} \tilde{A}_{t} e^{i t P(h)} h\right\|_{L^{2} \rightarrow L^{2}}=O(h) ;
$$

and so

$$
\left\|\tilde{A}_{t}-A_{t}\right\|_{L^{2} \rightarrow L^{2}}=\left\|\tilde{A}_{t}-e^{\frac{i t P(h)}{h}} A e^{-\frac{i t P(h)}{h}}\right\|_{L^{2} \rightarrow L^{2}}=O(h),
$$

uniformly for $0 \leq t \leq T$.

### 8.3 WEYL'S THEOREM GENERALIZED

NOTATION. We hereafter consider the eigenvalue problems

$$
P(h) u_{j}(h)=E_{j}(h) u_{j}(h) \quad(j=1, \ldots) .
$$

To simplify notation a bit, we write $u_{j}=u_{j}(h)$ and $E_{j}=E_{j}(h)$. We assume as well the normalization

$$
\begin{equation*}
\left\|u_{j}\right\|_{L^{2}(M)}=1 \tag{8.15}
\end{equation*}
$$

The following result generalizes Theorem 6.8, showing that we can localize the asymptotics using a quantum observable:

THEOREM 8.3 (Weyl's Theorem generalized). Let $B \in \Psi(M)$. Then

$$
\begin{equation*}
(2 \pi h)^{n} \sum_{a \leq E_{j} \leq b}\left\langle B u_{j}, u_{j}\right\rangle \rightarrow \iint_{\{a \leq p \leq b\}} \sigma(B) d x d \xi \tag{8.16}
\end{equation*}
$$

REMARK. If $B=I$, whence $\sigma(B) \equiv 1$, (8.16) reads

$$
(2 \pi h)^{n} \#\left\{a \leq E_{j} \leq b\right\} \rightarrow \operatorname{Vol}(\{a \leq p \leq b\})
$$

This is the usual form of Weyl's Law, Theorem E.7.
Proof. 1. We first assume that $B \in \Psi^{-\infty}$; so that the operator $B$ : $L^{2}(M) \rightarrow L^{2}(M)$ is of trace class. According to Lidskii's Theorem C. 9 from Appendix B, we have

$$
\begin{equation*}
\operatorname{tr}(B)=\frac{1}{(2 \pi h)^{n}}\left(\int_{M} \int_{\mathbb{R}^{n}} \sigma(B) d x d \xi+O(h)\right) \tag{8.17}
\end{equation*}
$$

2. Fix a small munber $\epsilon>0$ and write $\Omega_{\epsilon}:=p^{-1}(a-\epsilon, a+\epsilon) \cup$ $p^{-1}(b-\epsilon, b+\epsilon)$. Select $\psi_{\epsilon} \in C_{\mathrm{c}}^{\infty}, \varphi_{\epsilon} \in C^{\infty}$ so that

$$
\left\{\begin{array}{l}
\mathrm{WF}_{h}\left(\psi_{\epsilon}\right) \subset\{a<p<b\} \\
\mathrm{WF}_{h}\left(\varphi_{\epsilon}\right) \subset\{p<a\} \cup\{p>b\} \\
\mathrm{WF}_{h}\left(I-\varphi_{\epsilon}+\psi_{\epsilon}\right) \subset \Omega_{\epsilon}
\end{array}\right.
$$

Define

$$
\Pi:=\text { projection onto the span of }\left\{u_{j} \mid a \leq E_{j} \leq b\right\} .
$$

We claim that

$$
\left\{\begin{array}{l}
\psi_{\epsilon} \Pi=\psi_{\epsilon}+O\left(h^{\infty}\right)  \tag{8.18}\\
\varphi_{\epsilon} \Pi=O\left(h^{\infty}\right)
\end{array}\right.
$$

The second assertion follows by an adaptation of the proof of Theorem E.7.

To establish the first part, we need to show that $\psi_{\epsilon}(I-\Pi)=O\left(h^{\infty}\right)$. We can find $f$ satisfying the assumptions of Theorem 7.6 and such that $\psi_{\epsilon}(x, \xi) / f(p(x, \xi))$ is smooth. Using a symbolic construction we can find $T_{\epsilon} \in \Psi^{-\infty}$ with $\mathrm{WF}_{h}\left(T_{\epsilon}\right)=\mathrm{WF}_{h}\left(\psi_{\epsilon}\right)$, for which

$$
\psi_{\epsilon}(I-\Pi)=T_{\epsilon} f(P)(I-\Pi)+O(h) .
$$

The first term on the right hand side can be rewritten as

$$
\sum_{E_{j}(h)<a, E_{j}(h)>b} f\left(E_{j}(h)\right) T_{\epsilon} u_{j} \otimes \bar{u}_{j} .
$$

The rough estimate (E.31) and the rapid decay of $f$ show that for all $M$ we have the bound

$$
f\left(E_{j}(h)\right) \leq C_{M}(h j)^{-M} .
$$

The proof of Theorem 6.4 shows also that $T_{\epsilon} u_{j}=O\left(h^{\infty}\right)$, uniformly in $j$. Hence

$$
\left\|T_{\epsilon} f(P)(I-\Pi)\right\|_{L^{2} \rightarrow L^{2}}=O\left(h^{\infty}\right)
$$

3. We now write

$$
\begin{aligned}
\sum_{a \leq E_{j} \leq b}\left\langle B u_{j}, u_{j}\right\rangle & =\operatorname{tr}(\Pi B \Pi) \\
& =\operatorname{tr}\left(\Pi B\left(\psi_{\epsilon}+\varphi_{\epsilon}+\left(1-\varphi_{\epsilon}-\psi_{\epsilon}\right) \Pi\right)\right.
\end{aligned}
$$

Using (8.18) we see that

$$
(2 \pi h)^{n} \operatorname{tr}\left(\Pi B \varphi_{\epsilon} \Pi\right)=O(h) .
$$

The Weyl Law given in Theorem E. 7 implies

$$
(2 \pi h)^{n} \operatorname{tr}\left(\Pi B\left(1-\varphi_{\epsilon}-\psi_{\epsilon}\right) \Pi\right)=O(\epsilon)
$$

since $1-\varphi_{\epsilon}-\psi_{\epsilon} \neq 0$ only on $\Omega_{\epsilon}$. Furthermore,

$$
\begin{aligned}
(2 \pi h)^{n} \operatorname{tr} & \left(\Pi B \psi_{\epsilon} \Pi\right) \\
= & (2 \pi h)^{n} \operatorname{tr}\left(\Pi B \psi_{\epsilon}\right)+O\left(h^{\infty}\right) \\
= & (2 \pi h)^{n} \operatorname{tr}\left(\left(\left(\psi_{\epsilon}+\varphi_{\epsilon}\right.\right.\right. \\
& \left.\quad+\left(1-\varphi_{\epsilon}-\psi_{\epsilon}\right) \Pi B \psi_{\epsilon}\right)+O\left(h^{\infty}\right) \\
& =(2 \pi h)^{n} \operatorname{tr}\left(\psi_{\epsilon} B \psi_{\epsilon}\right)+O\left(h^{\infty}\right)+O(\epsilon) .
\end{aligned}
$$

Combining these calculations gives

$$
\begin{aligned}
(2 \pi h)^{n} \sum_{a \leq E_{j} \leq b}\left\langle B u_{j}, u_{j}\right\rangle & =(2 \pi h)^{n} \operatorname{tr}\left(\psi_{\epsilon} B \psi_{\epsilon}\right)+O(h)+O(\epsilon) \\
& =\iint \sigma\left(\psi_{\epsilon}\right)^{2} \sigma(B) d x d \xi+O(h)+O(\epsilon) \\
& \rightarrow \iint_{\{a \leq p \leq b\}} \sigma(B) d x d \xi
\end{aligned}
$$

as $h \rightarrow 0, \epsilon \rightarrow 0$.
4. Finally, to pass from $B \in \Psi^{-\infty}$ to an arbitrary $B \in \Psi$, we decompose the latter as

$$
B=B_{0}+B_{1},
$$

with $B_{0} \in \Psi^{-\infty}$ and

$$
\begin{aligned}
& \mathrm{WF}_{h}\left(B_{0}\right) \subset\{a-2<p<b+2\}, \\
& \mathrm{WF}_{h}\left(B_{1}\right) \cap\{a-1<p<b+1\}=\emptyset .
\end{aligned}
$$

We have $B_{1} u_{j}=O\left(h^{\infty}\right)$ for $a \leq E_{j}(h) \leq b$; and hence only the $B_{0}$ part contributes to the limit.

### 8.4 A QUANTUM ERGODIC THEOREM

Assume now that $A \in \Psi(M)$ has the symbol $\sigma(A)$ with the property that

$$
\begin{equation*}
\alpha:=\int_{\Sigma_{c}} \sigma(A) d \mu \text { is the same for all } c \in[a, b], \tag{8.19}
\end{equation*}
$$

where the slash through the integral denotes the average. In other words, we are requiring that the averages of the symbol of $A$ over each level surface $p^{-1}(c)$ are equal.

THEOREM 8.4 (Quantum ergodicity). Assume the ergodic condition (8.5) and that $A \in \Psi(M)$ satisfies the condition (8.19).
(i) Then

$$
\begin{equation*}
(2 \pi h)^{n} \sum_{a \leq E_{j} \leq b}\left|\left\langle A u_{j}, u_{j}\right\rangle-f_{\{a \leq p \leq b\}} \sigma(A) d x d \xi\right|^{2} \longrightarrow 0 \tag{8.20}
\end{equation*}
$$

(ii) In addition, there exists a family of subsets $\Lambda(h) \subseteq\left\{a \leq E_{j} \leq b\right\}$ such that

$$
\begin{equation*}
\lim _{h \rightarrow 0} \frac{\# \Lambda(h)}{\#\left\{a \leq E_{j} \leq b\right\}}=1 ; \tag{8.21}
\end{equation*}
$$

and for each $A \in \Psi(M)$ satisfying (8.19), we have

$$
\begin{equation*}
\left\langle A u_{j}, u_{j}\right\rangle \rightarrow f_{\{a \leq p \leq b\}} \sigma(A) d x d \xi \quad \text { as } h \rightarrow 0 \tag{8.22}
\end{equation*}
$$

for $E_{j} \in \Lambda(h)$.

Proof. 1. We first show that assertion (i) implies (ii). For this, let

$$
\begin{equation*}
B:=A-\alpha I, \tag{8.23}
\end{equation*}
$$

$\alpha$ defined by (8.19). Then $\int_{\{a \leq p \leq b\}} \sigma(B) d x d \xi=0$. According to (8.20),

$$
(2 \pi h)^{n} \sum_{a \leq E_{j} \leq b}\left\langle B u_{j}, u_{j}\right\rangle^{2}=: \epsilon(h) \rightarrow 0 .
$$

Define

$$
\Gamma(h):=\left\{a \leq E_{j} \leq b \mid\left\langle B u_{j}, u_{j}\right\rangle^{2} \geq \epsilon^{1 / 2}(h)\right\} ;
$$

so that

$$
(2 \pi h)^{n} \# \Gamma(h) \leq \epsilon(h)^{1 / 2}
$$

Next, write

$$
\Lambda(h):=\left\{a \leq E_{j} \leq b\right\}-\Gamma(h) .
$$

Then if $E_{j} \in \Lambda(h)$,

$$
\left|\left\langle B u_{j}, u_{j}\right\rangle\right| \leq \epsilon^{1 / 4}(h)
$$

and so

$$
\left|\left\langle A u_{j}, u_{j}\right\rangle-\alpha\right| \leq \epsilon^{1 / 4}(h) .
$$

Also,

$$
\frac{\# \Lambda(h)}{\#\left\{a \leq E_{j} \leq b\right\}}=1-\frac{\# \Gamma(h)}{\#\left\{a \leq E_{j}<b\right\}}
$$

But according to Weyl's law,

$$
\frac{\# \Gamma(h)}{\#\left\{a \leq E_{j} \leq b\right\}}=\frac{(2 \pi h)^{n} \# \Gamma(h)}{\operatorname{Vol}(\{a \leq p \leq b\})+o(1)} \leq C \epsilon(h)^{1 / 2} \rightarrow 0
$$

This proves (ii).
2. Next we establish assertion (i). Let $B$ be again given by (8.23); so that in view of our hypothesis (8.19)

$$
\begin{equation*}
\int_{\Sigma_{c}} \sigma(B) d \mu=0 \quad \text { for each } c \in[a, b] . \tag{8.24}
\end{equation*}
$$

Define

$$
\epsilon(h):=(2 \pi h)^{n} \sum_{a \leq E_{j} \leq b}\left\langle B u_{j}, u_{j}\right\rangle^{2} ;
$$

we must show $\epsilon(h) \rightarrow 0$.

Now

$$
\left\langle B u_{j}, u_{j}\right\rangle=\left\langle B e^{-\frac{i t E_{j}}{h}} u_{j}, e^{-\frac{i t E_{j}}{h}} u_{j}\right\rangle=\left\langle B e^{-\frac{i t P(h)}{h}} u_{j}, e^{-\frac{i t P(h)}{h}} u_{j}\right\rangle
$$

according to (8.9). Consequently

$$
\begin{equation*}
\left\langle B u_{j}, u_{j}\right\rangle=\left\langle e^{\frac{i t P(h)}{h}} B e^{-\frac{i t P(h)}{h}} u_{j}, u_{j}\right\rangle=\left\langle B_{t} u_{j}, u_{j}\right\rangle \tag{8.25}
\end{equation*}
$$

in the notation of (8.10). This identity is valid for each time $t \in \mathbb{R}$.
We can therefore average:

$$
\begin{equation*}
\left\langle B u_{j}, u_{j}\right\rangle=\left\langle f_{0}^{T} B_{t} d t u_{j}, u_{j}\right\rangle=\left\langle\langle B\rangle_{T} u_{j}, u_{j}\right\rangle \tag{8.26}
\end{equation*}
$$

for

$$
\langle B\rangle_{T}:=\frac{1}{T} \int_{0}^{T} B_{t} d t=f_{0}^{T} B_{t} d t
$$

Now since $\left\|u_{j}\right\|^{2}=1$, (8.26) implies

$$
\left\langle B u_{j}, u_{j}\right\rangle^{2}=\left\langle\langle B\rangle_{T} u_{j}, u_{j}\right\rangle^{2} \leq\left\|\langle B\rangle_{T} u_{j}\right\|^{2}=\left\langle\left\langle B^{*}\right\rangle_{T}\langle B\rangle_{T} u_{j}, u_{j}\right\rangle .
$$

Hence

$$
\begin{equation*}
\epsilon(h) \leq(2 \pi h)^{n} \sum_{a \leq E_{j} \leq b}\left\langle\left\langle B^{*}\right\rangle_{T}\langle B\rangle_{T} u_{j}, u_{j}\right\rangle \tag{8.27}
\end{equation*}
$$

3. Theorem 8.2 tells us that

$$
\langle B\rangle_{T}=\langle\tilde{B}\rangle_{T}+O_{T}(h), \quad\langle\tilde{B}\rangle_{T}:=\int_{0}^{T} \tilde{B}_{t} d t
$$

where $\tilde{B}_{t} \in \Psi(M)$ and $\sigma\left(\tilde{B}_{t}\right)=\varphi_{t}^{*} \sigma(B)$. Hence

$$
\sigma\left(\langle\tilde{B}\rangle_{T}\right)=f_{0}^{T} \sigma(B) \circ \varphi_{t} d t+O_{T}(h)=\langle\sigma(B)\rangle_{T}+O_{T}(h)
$$

as $h \rightarrow 0$.
Since modulo $O(h)$ errors we can replace $e^{i t P(h) / h} B e^{-i t P(h) / h}$ by $\tilde{B}_{t}$, Theorem 8.3 shows that

$$
\begin{align*}
\limsup _{h \rightarrow 0} \epsilon(h) & \leq \iint_{\{a \leq p \leq b\}} \sigma\left(\left\langle\left(\tilde{B}^{*}\right\rangle_{T}\langle\tilde{B}\rangle_{T}\right) d x d \xi\right.  \tag{8.28}\\
& \left.=\iint_{\{a \leq p \leq b\}} \mid \sigma\left(\langle B\rangle_{T}\right)\right)\left.\right|^{2} d x d \xi
\end{align*}
$$

as the symbol map is multiplicative and the symbol of an adjoint is given by the complex conjugate.
4. We can now apply Theorem 8.1 with $f=\sigma(B)$, to conclude that

$$
\int_{p^{-1}[a, b]}\left|\langle\sigma(B)\rangle_{T}\right|^{2} d x d \xi \rightarrow 0
$$

as $T \rightarrow \infty$. Since the left hand side of (8.28) is independent of $T$, this calculation shows that the limit must in fact be zero.

APPLICATION. The simplest and most striking application concerns the complete set of eigenfuctions of the Laplacian on a compact Riemannian manifold:

$$
-\Delta_{g} u_{j}=\lambda_{j} u_{j} \quad(j=1, \ldots)
$$

normalized so that

$$
\left\|u_{j}\right\|_{L^{2}(M)}=1 .
$$

THEOREM 8.5 (Equidistribution of eigenfunctions). There exists a sequence $j_{k} \rightarrow \infty$ of density one,

$$
\lim _{m \rightarrow \infty} \frac{\#\left\{k \mid j_{k} \leq m\right\}}{m}=1
$$

such that for any $f \in C^{\infty}(M)$,

$$
\begin{equation*}
\int_{M}\left|u_{j_{k}}\right|^{2} f d \operatorname{vol}_{g} \rightarrow \int_{M} f d \operatorname{vol}_{g} \tag{8.29}
\end{equation*}
$$

## 9. More on the symbol calculus

9.1 Wavefront sets, essential support
9.2 Application: $L^{\infty}$ bounds
9.3 Beals's Theorem
9.4 Application: exponentiation of operators
9.5 Invariance, half-densities
9.6 Changing variables
9.7 New symbol classes

This chapter collects together various more advanced topics concerning the symbol calculus, discussing in particular a semiclassical version of Beals's characterization of pseudodifferential operators and invariance properties under changes of variable. Chapter 10 will provide further applications.

### 9.1 WAVEFRONT SETS, ESSENTIAL SUPPORT

We devote this section to a few concepts built around the semiclassical wavefront set of a collection of functions bounded in $L^{2}$ :
DEFINITION. Let $u=\{u(h)\}_{0<h \leq h_{0}}$ be a family of functions bounded in $L^{2}\left(\mathbb{R}^{n}\right)$. We define the semiclassical wavefront set

$$
\operatorname{WF}_{h}(u)
$$

to the complement of the set of points $\left(x_{0}, \xi_{0}\right) \in \mathbb{R}^{2 n}$ for which there exists a symbol $a \in S$ such that

$$
\begin{equation*}
a\left(x_{0}, \xi_{0}\right) \neq 0 \tag{9.1}
\end{equation*}
$$

and

$$
\begin{equation*}
\left\|a^{\mathrm{w}}(x, h D) u(h)\right\|_{L^{2}}=O\left(h^{\infty}\right) \tag{9.2}
\end{equation*}
$$

The definition of wavefront sets does not depend on the choice of coordinates. We note that this is a local property of the family $\{u(h)\}_{0<h \leq h_{0}}$ in phase space: see Theorem 9.2 below. The meaning of the wavefront set is elucidated by the following
THEOREM 9.1 (Localization and wavefront sets). Suppose that

$$
\left(x_{0}, \xi_{0}\right) \notin \mathrm{WF}_{h}(u)
$$

Then for any $b \in C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{2 n}\right)$ with support sufficiently close to $\left(x_{0}, \xi_{0}\right)$, we have

$$
b^{\mathrm{w}}(x, h D) u(h)=O_{L^{2}}\left(h^{\infty}\right) .
$$

Proof. 1. Suppose $a \in S, a\left(x_{0}, \xi_{0}\right) \neq 0$. There exists $\chi \in C^{\infty}\left(\mathbb{R}^{2 n}\right)$ supported near $\left(x_{0}, \xi_{0}\right)$ such that

$$
\left|\chi(x, \xi)\left(a(x, \xi)-a\left(x_{0}, \xi_{0}\right)\right)+a\left(x_{0}, \xi_{0}\right)\right| \geq \gamma>0
$$

for $(x, \xi) \in \mathbb{R}^{2 n}$. Then according to Theorem 4.19 there exists $c \in S$ such that

$$
c^{\mathrm{w}}(x, h D)\left(\chi^{\mathrm{w}}(x, h D) a^{\mathrm{w}}(x, h D)+a\left(x_{0}, \xi_{0}\right)\left(I-\chi^{\mathrm{w}}(x, h D)\right)\right)=I .
$$

2. Now consider

$$
\begin{aligned}
b^{\mathrm{w}}(x, h D) u(h)= & b^{\mathrm{w}}(x, h D) c^{\mathrm{w}}(x, h D) \chi^{\mathrm{w}}(x, h D) a^{\mathrm{w}}(x, h D) u(h) \\
& \left.+b^{\mathrm{w}}(x, h D) a\left(x_{0}, \xi_{0}\right)\left(I-\chi^{\mathrm{w}}(x, h D)\right)\right) u(h) .
\end{aligned}
$$

If we choose $a$ to be the symbol appearing in (9.2), then the first term on the right hand side is bounded by $O\left(h^{\infty}\right)$ in $L^{2}$. If the support of $b$ is sufficiently close to $\left(x_{0}, \xi_{0}\right)$ then $\operatorname{spt}(b) \cap \operatorname{spt}(1-\chi)=\emptyset$ and the second term has the same property, according to Theorem ?????

Since compactness of the support is not preserved by changes of variables or by other operations such as composition, we introduce the more flexible notion of the essential support.

DEFINITION. Let $a=\{a(x, \xi, h)\}_{0<h \leq h_{0}}$ be an $h$-dependent family of symbols in $\mathcal{S}$. The essential support of $a$ is the smallest closed set $K \subseteq \mathbb{R}^{n} \times \mathbb{R}^{n}$ such that

$$
\begin{equation*}
\text { spt } \chi \cap K=\emptyset \text { implies } \chi a=O_{S}\left(h^{\infty}\right) \tag{9.3}
\end{equation*}
$$

for each $\chi \in S$. We write

$$
\text { ess-spt }(a)
$$

for the essential support.
Recall that a symbol $b=b(x, \xi, h)$ satisfies $b=O_{S}\left(h^{\infty}\right)$ if

$$
\left|\partial^{\alpha} b\right| \leq C_{\alpha, N} h^{N}
$$

for all multiindices $\alpha$ and nonnegative integers $N$.

REMARK. We will see later, in the proof of Theorem 9.12, that this notion does not depend on the choice of coordinates. In particular, if $\gamma$ is a diffeomorphism and $\operatorname{Op}\left(a_{\gamma}\right)=\operatorname{Op}\left(\left(\gamma^{-1}\right)^{*} \operatorname{Op}(a) \gamma^{*}\right)$, then

$$
\begin{equation*}
\operatorname{ess-spt}\left(a_{\gamma}\right)=\left\{\left(\gamma(x),\left(\partial \gamma(x)^{T}\right)^{-1} \xi\right) \mid(x, \xi) \in \operatorname{ess-spt}(a)\right\} \tag{9.4}
\end{equation*}
$$

THEOREM 9.2 (Wavefront sets and pseudodifferential operators). Suppose that $a=\{a(x, \xi, h)\}_{0<h \leq h_{0}} \subset S(m)$ for some order function $m$, and that $u=\{u(h)\}_{0<h \leq h_{0}}$ is bounded in $L^{2}\left(\mathbb{R}^{n}\right)$.

Then

$$
\begin{equation*}
\mathrm{WF}_{h}\left(a^{\mathrm{w}} u\right) \subset \mathrm{WF}_{h}(u) \cap \operatorname{ess-spt}(a) \tag{9.5}
\end{equation*}
$$

for $a^{\mathrm{w}} u:=\left\{a^{\mathrm{w}}(x, h D, h) u(h)\right\}_{0<h \leq h_{0}}$.

Proof: 1. We need to show that if $\left(x_{0}, \xi_{0}\right) \notin \mathrm{WF}_{h}(u)$ or if $(x, \xi) \notin$ ess-spt $(a)$, then $(x, \xi) \notin \mathrm{WF}_{h}\left(a^{\mathrm{w}} u\right)$.
2. Suppose first that $\left(x_{0}, \xi_{0}\right) \notin \mathrm{WF}_{h}(u)$. Choose $b \in C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{2 n}\right)$, with $b\left(x_{0}, \xi_{0}\right) \neq 0$ and $b^{\mathrm{w}}(x, h D) u(h)=O_{L^{2}}\left(h^{\infty}\right)$. The existence of such $b$ is clear from Theorem 9.1. Now the pseudodifferential calculus shows that

$$
b^{\mathrm{w}}(x, h D) a^{\mathrm{w}}(x, h D, h)=c^{\mathrm{w}}(x, h D, h)+r^{\mathrm{w}}(x, h D, h),
$$

where $\operatorname{spt} c \subset \operatorname{spt} b$ and $r \in S^{-\infty}(1)$. Theorem 9.1 implies that

$$
\left\|b^{\mathrm{w}}(x, h D) a^{\mathrm{w}}(x, h D, h) u(h)\right\|_{L^{2}}=O\left(h^{\infty}\right) .
$$

This shows that $\left(x_{0}, \xi_{0}\right) \notin \mathrm{WF}_{h}\left(a^{\mathrm{w}} u\right)$.
3. Now assume that $\left(x_{0}, \xi_{0}\right) \notin \operatorname{ess-spt}(a)$ and use the same $b$ as above. If the support of $b$ is sufficiently close to $\left(x_{0}, \xi_{0}\right)$, (9.3) implies that $b^{\mathrm{w}}(x, h D) a^{\mathrm{w}}(x, h D, h)=c^{\mathrm{w}}(x, h D, h)$, where $c=O_{L^{2}}\left(h^{\infty}\right)$. Consequently

$$
\left\|b^{\mathrm{w}}(x, h D) a^{\mathrm{w}}(x, h D, h) u(h)\right\|_{L^{2}}=O\left(h^{\infty}\right) .
$$

REMARK. In view of (9.5) we have an alternative definition of ess-spt ( $a$ ), which makes sense on manifolds:

$$
\left\{\begin{array}{c}
(x, \xi) \notin \operatorname{ess-spt}(a) \text { if and only if }(x, \xi) \notin \mathrm{WF}_{h}\left(a^{\mathrm{w}} u\right)  \tag{9.6}\\
\text { for each family } u=\{u(h)\}_{0<h \leq h_{0}} \text { bounded in } L^{2} .
\end{array}\right.
$$

Theorem 9.2 also motivates the following
DEFINITION. Let $A$ be an $h$-dependent family of operators. We define the wavefront set of $A$ to be

$$
\begin{equation*}
\mathrm{WF}_{h}(A):=\bigcup \mathrm{WF}_{h}(u) \cap \mathrm{WF}_{h}(A u), \tag{9.7}
\end{equation*}
$$

the union taken over all families $\{u(h)\}_{0<h \leq h_{0}}$ bounded in $L^{2}$.

In view of (9.6), if $A=a^{\mathrm{w}}$, then

$$
\mathrm{WF}_{h}(A)=\operatorname{ess}-\operatorname{spt}(a)
$$

and hence is a closed set.

### 9.2 APPLICATION: $L^{\infty}$ BOUNDS

Here we will show how a natural frequency localization condition on approximate solutions to pseudodifferential equations implies $h$ dependent $L^{\infty}$ bounds. As an application we will provide bounds on eigenfuction clusters for compact Riemannian manifolds.

We start with the following semiclassical version of the Sobolev inequality (Lemma 3.5):

THEOREM 9.3 (Basic $\boldsymbol{L}^{\boldsymbol{\infty}}$ bounds). Suppose that $\{u(h)\}_{0<h \leq h_{0}}$ is bounded in $L^{2}\left(\mathbb{R}^{n}\right)$ and there exists $\psi \in C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{n}\right)$ such that

$$
\begin{equation*}
\|(1-\psi(h D)) u(h)\|_{L^{2}\left(\mathbb{R}^{n}\right)}=O\left(h^{\infty}\right)\|u(h)\|_{H^{k}\left(\mathbb{R}^{n}\right)}, \text { for all } k . \tag{9.8}
\end{equation*}
$$

Then

$$
\begin{equation*}
\|u(h)\|_{L^{\infty}\left(\mathbb{R}^{n}\right)} \leq C h^{-n / 2}\|u(h)\|_{L^{2}\left(\mathbb{R}^{n}\right)} \tag{9.9}
\end{equation*}
$$

We regard (9.8) as a frequency localization condition.
Proof: 1 . We can assume that $\|u(h)\|_{L^{2}}=1$.
We can also suppose that $u(h)$ is compactly supported. Indeed, if $\varphi \in C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{n}\right)$, then

$$
(1-\psi(h D)) \varphi=\varphi(1-\psi(h D))+r(x, h D)
$$

with $r \in S$ and ess-spt $(r)$ compact. We now choose $\psi_{1} \in C_{\mathrm{c}}^{\infty}$ for which

$$
\left(1-\psi_{1}\right)(1-\psi)=\left(1-\psi_{1}\right),\left.\quad\left(1-\psi_{1}\right)\right|_{\mathrm{ess}-\mathrm{spt}(r)}=0
$$

Then

$$
\begin{aligned}
\left(1-\psi_{1}(h D)\right) \varphi u(h)= & \left(1-\psi_{1}(h D)\right) \psi(1-\psi(h D)) u(h) \\
& +\left(1-\psi_{1}(h D)\right) r(x, h D) u(h) \\
= & O_{L^{2}}\left(h^{\infty}\right)
\end{aligned}
$$

2. The condition (9.8) implies that

$$
\left\|\langle h D\rangle^{k} u(h)\right\| \leq h^{-N_{k}}
$$

for every $k$. Hence

$$
\begin{aligned}
\left\|\langle h D\rangle^{k}(1-\psi(h D)) u(h)\right\|_{L^{2}}^{2} & =\frac{1}{(2 \pi)^{n}}\left\|\langle\xi\rangle^{k}(1-\psi(\xi)) \hat{u}\right\|_{L^{2}}^{2} \\
& \leq \frac{1}{(2 \pi)^{n}}\left\|\langle\xi\rangle^{2 k} \hat{u}\right\|_{L^{2}}\left\|(1-\psi(\xi))^{2} \hat{u}\right\|_{L^{2}} \\
& =\left\|\langle h D\rangle^{2 k} u(h)\right\|_{L^{2}}\left\|(1-\psi(h D))^{2} u(h)\right\|_{L^{2}} \\
& =O\left(h^{\infty}\right),
\end{aligned}
$$

Lemma 3.5 then implies

$$
\|(1-\psi(h D)) u(h)\|_{L^{\infty}}=O\left(h^{\infty}\right) .
$$

3. It remains to estimate $\|\psi(h D) u\|_{L^{\infty}}$. For this we use the semiclassical inverse Fourier transform (3.22):

$$
\|\psi(h D) u\|_{L^{\infty}} \leq \frac{1}{(2 \pi h)^{n}}\|\psi\|_{L^{2}}\left\|\mathcal{F}_{h} u\right\|_{L^{2}}=\frac{1}{(2 \pi h)^{n / 2}}\|\psi\|_{L^{2}}\|u\|_{L^{2}}
$$

We will later need the following
LEMMA 9.4 (A simple $\boldsymbol{L}^{2}$ estimate). Suppose that $a \in S\left(\mathbb{R}^{n+1} \times\right.$ $\left.\mathbb{R}^{n}\right)$ is real valued, and that

$$
\left\{\begin{array}{l}
\left(h D_{t}+a^{\mathrm{w}}\left(x, t, h D_{x}\right)\right) u=f \\
u(\cdot, 0)=u_{0}
\end{array}\right.
$$

Then

$$
\begin{equation*}
\|u(\cdot, t)\|_{L^{2}\left(\mathbb{R}^{n}\right)} \leq \frac{\sqrt{|t|}}{h}\|f\|_{L^{2}\left(\mathbb{R} \times \mathbb{R}^{n}\right)}+\left\|u_{0}\right\|_{L^{2}\left(\mathbb{R}^{n}\right)} . \tag{9.10}
\end{equation*}
$$

Proof: Since $a^{\mathrm{w}}(t, x, h D)$ is family of bounded operators on $L^{2}\left(\mathbb{R}^{n}\right)$, the existence of solutions follows from existence theory for (linear) ordinary differential equations in the variable $t$.

Suppose first that $f \equiv 0$. Then

$$
\begin{aligned}
\frac{1}{2} \frac{d}{d t}\|u(\cdot, t)\|_{L^{2}\left(\mathbb{R}^{n}\right)}^{2} & =\operatorname{Re}\left\langle\partial_{t} u(\cdot, t), u(\cdot, t)\right\rangle_{L^{2}\left(\mathbb{R}^{n}\right)} \\
& =\frac{1}{h} \operatorname{Re}\left\langle i a^{\mathrm{w}}(x, t, h D) u(\cdot, t), u(\cdot, t)\right\rangle=0 .
\end{aligned}
$$

Thus, if we set $E(t) u_{0}:=u(t)$, we have

$$
\left\|E(t) u_{0}\right\|_{L^{2}\left(\mathbb{R}^{n}\right)}=\left\|u_{0}\right\|_{L^{2}\left(\mathbb{R}^{n}\right)} .
$$

If $f \neq 0$, Duhamel's formula gives

$$
u(\cdot, t)=E(t) u_{0}+\frac{i}{h} \int_{0}^{t} E(t-s) f(\cdot, s) d s
$$

Hence

$$
\|u(\cdot, t)\|_{L^{2}\left(\mathbb{R}^{n}\right)} \leq\left\|u_{0}\right\|_{L^{2}\left(\mathbb{R}^{n}\right)}+\frac{1}{h} \int_{0}^{t}\|f(\cdot, s)\|_{L^{2}\left(\mathbb{R}^{n}\right)} d s
$$

The estimate (9.10) is an immediate consequence.

THEOREM 9.5 ( $L^{\infty}$ bounds for approximate solutions). Let $m=m(x, \xi)$ an order function. Suppose that $p \in S(m)$ is real valued, and for some compact set $K \subset \mathbb{R}^{n} \times \mathbb{R}^{n}$,

$$
\begin{equation*}
\partial_{\xi} p(x, \xi) \neq 0 \quad \text { if } \quad(x, \xi) \in K, p(x, \xi)=0 \tag{9.11}
\end{equation*}
$$

If $u=\{u(h)\}_{0<h \leq h_{0}}$ is bounded in $L^{2}\left(\mathbb{R}^{n}\right)$ and satisfies the frequency localization condition (9.8), if

$$
W F_{h}(u) \subset K
$$

and if

$$
\begin{equation*}
\left\|p^{\mathrm{w}}(x, h D) u(h)\right\|_{L^{2}\left(\mathbb{R}^{n}\right)}=O(h)\|u(h)\|_{L^{2}\left(\mathbb{R}^{n}\right)} \tag{9.12}
\end{equation*}
$$

then

$$
\begin{equation*}
\|u(h)\|_{L^{\infty}\left(\mathbb{R}^{n}\right)} \leq C h^{-(n-1) / 2}\|u(h)\|_{L^{2}\left(\mathbb{R}^{n}\right)} . \tag{9.13}
\end{equation*}
$$

The point is that if $u(h)$ is an approximate solution in the sense of satisfying the estimate (9.12), we can improve the earlier $L^{\infty}$ estimate (9.9) by a factor of $h^{\frac{1}{2}}$.

Proof: 1. First we observe that, as in Theorem 9.3, we can assume that the functions $u(h)$ are uniformly compactly supported. We also note that the hypothesis on $u(h)$ is local in phase space: if $\chi \in C_{\mathrm{c}}^{\infty}\left(T^{*} \mathbb{R}^{n}\right)$ then, normalizing to $\|u(h)\|_{L^{2}}=1$, we have

$$
\begin{aligned}
P(h) \chi^{\mathrm{w}}(x, h D) u(h) & =\chi^{\mathrm{w}}(x, h D) P(h) u(h)+\left[P(h), \chi^{\mathrm{w}}(x, h D)\right] u(h) \\
& =O_{L^{2}}(h)
\end{aligned}
$$

According to Theorem 9.2, $\mathrm{WF}_{h}\left(\chi^{\mathrm{w}} u\right) \subset K$.
2. Hence it is enough to prove the theorem for $u(h)$ replaced by $\chi^{\mathrm{w}} u(h)$, where $\chi$ is supported near a given point in $K$ as a partition of unity argument will then give the bound on $u(h)$.

Suppose that $p \neq 0$ on the support of $\chi$. Then we can use Theorem 4.19 as in part 1 of the proof of Lemma 9.1 to see that $P(h) \chi^{\mathrm{w}} u(h)=$ $O_{L^{2}}(h)$ implies that $\chi^{\mathrm{w}} u(h)=O_{L^{2}}(h)$. Theorem 9.3 then shows that

$$
\left\|\chi^{\mathrm{w}} u(h)\right\|_{L^{\infty}} \leq C h h^{-n / 2} \leq C h^{-(n-1) / 2}
$$

3. Now suppose that $p$ vanishes in the support of $\chi$. By applying a linear change of variables we can assume that $p_{\xi_{1}} \neq 0$ there. The Implicit Function Theorem shows that in a neighborhood of spt $\chi$, we have

$$
\begin{equation*}
p(x, \xi)=e(x, \xi)\left(\xi_{1}-a\left(x, \xi^{\prime}\right)\right) \tag{9.14}
\end{equation*}
$$

where $\xi=\left(\xi_{1}, \xi^{\prime}\right)$ and $e(x, \xi)>0$.
We extend $e$ arbitrarily to $e \in S$ so that $e \geq \gamma>0$ and extend $a\left(x, \xi^{\prime}\right)$ to a real valued $a\left(x, \xi^{\prime}\right) \in S$. The pseudodifferential calculus shows that

$$
\begin{aligned}
e^{\mathrm{w}}(x, h D)\left(h D_{x_{1}}-a\left(x, h D_{x^{\prime}}\right)\right)\left(\chi^{\mathrm{w}} u(h)\right) & =P(h)\left(\chi^{\mathrm{w}} u(h)\right)+O_{L^{2}}(h) \\
& =O_{L^{2}}(h) .
\end{aligned}
$$

Since $e^{\mathrm{w}}$ is elliptic,

$$
\begin{equation*}
\left(h D_{x_{1}}-a\left(x, h D_{x^{\prime}}\right)\right)\left(\chi^{\mathrm{w}} u(h)\right)=O_{L^{2}}(h) . \tag{9.15}
\end{equation*}
$$

4. The proof will be completed once we show

$$
\begin{equation*}
\left\|\left(\chi^{\mathrm{w}} u\right)\left(x_{1}, \cdot\right)\right\|_{L^{2}\left(\mathbb{R}^{n-1}\right)}=O(1) \tag{9.16}
\end{equation*}
$$

and for that we use (9.15) and Lemma 9.4. We now apply Theorem 9.3 in $x^{\prime}$ variables only, that is with $n-1$ replacing $n$ and $t$ replacing $x_{1}$. That is allowed since we clearly have

$$
\left\|\left(1-\psi\left(h D^{\prime}\right)\right) \chi^{\mathrm{w}} u(h)\left(x_{1}, \cdot\right)\right\|_{L^{2}\left(\mathbb{R}^{n-1}\right)}=O\left(h^{\infty}\right),
$$

uniformly in $x_{1}$.
REMARK. The bound given in Theorem 9.5 is already optimal in the simplest case in which the assumptions are satisfied: $p(x, \xi)=\xi_{1}$. Indeed, write $x=\left(x_{1}, x^{\prime}\right)$ and let $\phi \in C_{\mathrm{c}}^{\infty}(\mathbb{R})$, and $\chi \in C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{n-1}\right)$. Then

$$
u(h):=h^{-(n-1) / 2} \phi\left(x_{1}\right) \chi\left(x^{\prime} / h\right)
$$

satisfies $\|u(h)\|_{L^{2}}=O(1)$,

$$
P(h) u(h)=h D_{x_{1}} u(h)=O_{L^{2}}(h) ;
$$

and for any non-trivial choices of $\phi$ and $\chi,\|u(h)\|_{L^{\infty}} \simeq h^{-(n-1) / 2}$.
The condition (9.11) is in general necessary as shown by another simple example. Let $p(x, \xi)=x_{1}$, and

$$
u(h)=h^{-n / 2} \phi\left(x_{1} / h\right) \chi\left(x^{\prime} / h\right) .
$$

Then $\|u(h)\|_{L^{2}}=O(1)$,

$$
P(h) u(h)=\left.h h^{-n / 2}(t \phi(t))\right|_{t=x_{1} / h} \chi\left(x^{\prime} / h\right)=O_{L^{2}}(h),
$$

and $\|u(h)\|_{L^{\infty}} \simeq h^{-n / 2}$. This is the general bound of Lemma 9.3.

As an application we present an $L^{\infty}$ bound on "spectral clusters", that is, linear combinations of eigenfunctions for the Laplacian on a compact manifold. The statement requires the material presented in Appendix E.3.

THEOREM 9.6 (Bounds on eigenfuction clusters). Suppose that $M$ is an n-dimensional compact Riemannian manifold and let $\Delta_{g}$ be its Laplace-Beltrami operator. Assume that

$$
0=\lambda_{0}<\lambda_{1} \leq \cdots \lambda_{j} \rightarrow \infty
$$

are the eigenvalues of $-\Delta_{g}$ and that

$$
-\Delta_{g} \varphi_{j}=\lambda_{j} \varphi_{j} \quad(j=1, \ldots)
$$

are a corresponding orthonormal basis of eigenfunctions.
There exists a constant $C$ such that for any choices of constants $c_{j} \in \mathbb{C}$, we have the inequality

$$
\begin{equation*}
\left\|\sum_{\mu \leq \sqrt{\lambda_{j} \leq \mu+1}} c_{j} \varphi_{j}\right\|_{L^{\infty}} \leq C \mu^{(n-1) / 2}\left\|\sum_{\mu \leq \sqrt{\lambda_{j}} \leq \mu+1} c_{j} \varphi_{j}\right\|_{L^{2}} . \tag{9.17}
\end{equation*}
$$

In particular,

$$
\begin{equation*}
\left\|\varphi_{j}\right\|_{L^{\infty}} \leq C \lambda_{j}^{(n-1) / 4}\left\|\varphi_{j}\right\|_{L^{2}} \tag{9.18}
\end{equation*}
$$

Proof: Put $h=1 / \mu, P(h):=-h^{2} \Delta_{g}-1$, and

$$
u(h):=\sum_{\mu \leq \sqrt{\lambda_{j} \leq \mu+1}} c_{j} \varphi_{j} .
$$

Then the assumption (9.11) holds everywhere. Also

$$
\begin{aligned}
\|P(h) u(h)\|_{L^{2}} & =\left\|\sum_{\mu \leq \sqrt{\lambda_{j}} \leq \mu+1} c_{j}\left(h^{2} \lambda_{j}-1\right) \varphi_{j}\right\|_{L^{2}} \\
& =\left(\sum_{\mu \leq \sqrt{\lambda_{j}} \leq \mu+1}\left|c_{j}\right|^{2}\left(h^{2} \lambda_{j}-1\right)^{2}\left\|\varphi_{j}\right\|_{L^{2}}^{2}\right)^{\frac{1}{2}} \\
& \leq 2 h\|u(h)\|_{L^{2}} .
\end{aligned}
$$

Thus the assumption (9.12) holds. On a compact manifold the frequency localization condition (9.8) follows from

$$
\left\|\left(1-\varphi\left(-h^{2} \Delta_{g}\right)\right) u(h)\right\|=O\left(h^{\infty}\right)\|u(h)\|_{L}^{2}
$$

for $\varphi \in C_{\mathrm{c}}^{\infty}(\mathbb{R})$ satisfying $\varphi(t) \equiv 1$ for $|t| \leq 2$. But this is a consequence of the Spectral Theorem.

The estimate (9.17) is essentially optimal. On the other hand the optimality of (9.18) is very rare: see [S-Z] for a recent discussion.

### 9.3 BEALS'S THEOREM

We next present a semiclassical version of Beals's Theorem, a characterization of pseudodifferential operators in terms of $h$-dependent bounds on commutators. Beals's Theorem answers a fundamental question: when can a given linear operator be represented using our symbol calculus?

We start with

THEOREM 9.7 (Estimating a symbol by operator norms).
Take $h=1$. There exist constants $C, M>0$ such that

$$
\begin{equation*}
\|b\|_{L^{\infty}} \leq C \sum_{|\alpha| \leq M}\left\|\left(\partial^{\alpha} b\right)^{\mathrm{w}}(x, D)\right\|_{L^{2} \rightarrow L^{2}} \tag{9.19}
\end{equation*}
$$

for all $b \in \mathcal{S}^{\prime}\left(\mathbb{R}^{2 n}\right)$.

Proof. 1. We will first consider the classical quantization

$$
b(x, D) u(x)=\frac{1}{(2 \pi)^{n}} \int_{\mathbb{R}^{n}} b(x, \xi) e^{i\langle x, \xi\rangle} \hat{u}(\xi) d \xi
$$

where by the integration we mean the Fourier transform in $\mathcal{S}^{\prime}$.
Then if $\varphi=\varphi(x), \psi=\psi(\xi)$ are functions in the Schwartz space $\mathcal{S}$, we can regard $\mathcal{F}\left(b \bar{\varphi} \hat{\psi} e^{i\langle x, \xi\rangle}\right)$ as a function of the dual variables $\left(x^{*}, \xi^{*}\right) \in$ $\mathbb{R}^{2 n}$. We have

$$
\begin{aligned}
\left.\frac{1}{(2 \pi)^{n}} \right\rvert\, & \mathcal{F} \\
& \left(b \bar{\varphi} \hat{\psi} e^{i\langle x, \xi\rangle}\right)(0,0) \mid \\
& =\frac{1}{(2 \pi)^{n}}\left|\int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} b(x, \xi) \bar{\varphi}(x) \hat{\psi}(\xi) e^{i\langle x, \xi\rangle} d x d \xi\right| \\
& =|\langle b(x, D) \psi, \varphi\rangle| \leq\|b\|_{L^{2} \rightarrow L^{2}}\|\varphi\|_{L^{2}}\|\psi\|_{L^{2}} .
\end{aligned}
$$

Fix $\left(x^{*}, \xi^{*}\right) \in \mathbb{R}^{2 n}$ and rewrite this inequality with $\varphi(x) e^{i\left\langle x^{*}, x\right\rangle}$ replacing $\varphi(x)$ and $\psi(\xi) e^{-i\left\langle\xi^{*}, \xi\right\rangle}$ replacing $\psi(\xi)$, a procedure which does not change the $L^{2}$ norms. It follows that

$$
\begin{equation*}
\frac{1}{(2 \pi)^{n}}\left|\mathcal{F}\left(b \bar{\varphi} \hat{\psi} e^{i\langle x, \xi\rangle}\right)\left(x^{*}, \xi^{*}\right)\right| \leq\|b\|_{L^{2} \rightarrow L^{2}}\|\varphi\|_{L^{2}}\|\psi\|_{L^{2}} \tag{9.20}
\end{equation*}
$$

2. Now take $\chi \in C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{2 n}\right)$. Select $\varphi, \psi \in \mathcal{S}$ so that

$$
\begin{cases}\varphi(x)=1 & \text { if }(x, \xi) \in \operatorname{spt} \chi \\ \hat{\psi}(\xi)=1 & \text { if }(x, \xi) \in \operatorname{spt} \chi\end{cases}
$$

Write

$$
\begin{equation*}
\varphi=\chi e^{-i\langle x, \xi\rangle} \tag{9.21}
\end{equation*}
$$

Then

$$
\chi(x, \xi)=\varphi(x, \xi) \varphi(x) \hat{\psi}(\xi) e^{i\langle x, \xi\rangle}
$$

According to (3.18),

$$
\|\mathcal{F} \chi\|_{L^{1}} \leq C \sum_{|\alpha| \leq 2 n+1}\left\|\partial^{\alpha} \chi\right\|_{L^{1}}
$$

and so (9.21) implies

$$
\begin{equation*}
\|\mathcal{F} \varphi\|_{L^{1}} \leq C \sum_{|\alpha| \leq 2 n+1}\left\|\partial^{\alpha} \chi\right\|_{L^{1}} \tag{9.22}
\end{equation*}
$$

Thus (9.20) shows that for any $\left(x^{*}, \xi^{*}\right) \in \mathbb{R}^{2 n}$ we have

$$
\begin{aligned}
\left|\mathcal{F}(\chi b)\left(x^{*}, \xi^{*}\right)\right| & \leq\left\|\mathcal{F}\left(\varphi b \bar{\varphi} \hat{\psi} e^{i\langle x, \xi\rangle}\right)\right\|_{L^{\infty}} \\
& =\frac{1}{(2 \pi)^{n}}\left\|\mathcal{F}(\varphi) * \mathcal{F}\left(b \bar{\varphi} \hat{\psi} e^{i\langle x, \xi\rangle}\right)\right\|_{L^{\infty}} \\
& \leq \frac{1}{(2 \pi)^{n}}\left\|\mathcal{F}\left(b \bar{\varphi} \hat{\psi} e^{i\langle x, \xi\rangle}\right)\right\|_{L^{\infty}}\|\mathcal{F}(\varphi)\|_{L^{1}} \\
& \leq C\|b\|_{L^{2} \rightarrow L^{2}},
\end{aligned}
$$

the constant $C$ depending on $\varphi, \psi$ and $\chi$, but not $\left(x^{*}, \xi^{*}\right)$. Hence

$$
\begin{equation*}
\|\mathcal{F}(\chi b)\|_{L^{\infty}} \leq C\|b\|_{L^{2} \rightarrow L^{2}} \tag{9.23}
\end{equation*}
$$

with the same constant for any translate of $\chi$.
3. Next, we assert that

$$
\begin{equation*}
\left|\mathcal{F}(\chi b)\left(x^{*}, \xi^{*}\right)\right| \leq C\left\langle\left(x^{*}, \xi^{*}\right)\right\rangle^{-2 n-1} \sum_{|\alpha| \leq 2 n+1}\left\|\left(\partial^{\alpha} b\right)(x, h D)\right\|_{L^{2} \rightarrow L^{2}} \tag{9.24}
\end{equation*}
$$

To see this, compute

$$
\begin{aligned}
\left(x^{*}\right)^{\alpha}\left(\xi^{*}\right)^{\beta} \mathcal{F}(\chi b)\left(x^{*}, \xi^{*}\right) & =\int_{\mathbb{R}^{2 n}}\left(x^{*}\right)^{\alpha}\left(\xi^{*}\right)^{\beta} e^{-i\left(\left\langle x^{*}, x\right\rangle+\left\langle\xi^{*}, \xi\right\rangle\right)} \chi b(x, \xi) d x d \xi \\
& =\int_{\mathbb{R}^{2 n}} e^{-i\left(\left\langle x^{*}, x\right\rangle+\left\langle\xi^{*}, \xi\right\rangle\right)} D_{x}^{\alpha} D_{\xi}^{\beta}(\chi b) d x d \xi
\end{aligned}
$$

Summing absolute values of the left hand side over all $(\alpha, \beta)$ with $|\alpha|+$ $|\beta| \leq 2 n+1$ and using the estimate (9.23), we obtain the bound

$$
\begin{aligned}
\left\|\left\langle\left(x^{*}, \xi^{*}\right)\right\rangle^{2 n+1} \mathcal{F}(\chi b)\right\|_{L^{\infty}} & \leq C_{1} \sum_{|\alpha|+|\beta| \leq 2 n+1}\left\|\mathcal{F}\left(D_{x}^{\alpha} D_{\xi}^{\beta}(\chi b)\right)\right\|_{L^{\infty}} \\
& \leq C_{2} \sum_{|\alpha| \leq 2 n+1}\left\|\left(\partial^{\alpha} b\right)(x, h D)\right\|_{L^{2} \rightarrow L^{2}}
\end{aligned}
$$

This gives (9.24).

Consequently,

$$
\|\chi b\|_{L^{\infty}} \leq C\|\mathcal{F}(\chi b)\|_{L^{1}} \leq C \sum_{|\alpha| \leq 2 n+1}\left\|\left(\partial^{\alpha} b\right)(x, h D)\right\|_{L^{2} \rightarrow L^{2}} .
$$

4. This implies the desired inequality (9.19), except that we used the classical $(t=1)$, and not the Weyl $(t=1 / 2)$ quantization. To remedy this, recall from Theorem ?? that if

$$
b=e^{\frac{i}{2}\left\langle D_{x}, D_{\xi}\right\rangle} \tilde{b},
$$

then

$$
\left\{\begin{array}{l}
b^{\mathrm{w}}(x, D)=\tilde{b}(x, D) \\
\left(\partial^{\alpha} b\right)^{\mathrm{w}}(x, D)=\left(\partial^{\alpha} \tilde{b}\right)(x, D)
\end{array}\right.
$$

The continuity statement in Theorem ?? shows that

$$
\|b\|_{L^{\infty}} \leq C \sum_{|\alpha| \leq K}\left\|\partial^{\alpha} \tilde{b}\right\|_{L^{\infty}}
$$

and reduces the argument to the classical quantization.

NOTATION. We henceforth write

$$
\operatorname{ad}_{B} A:=[B, A] ;
$$

"ad" is called the adjoint action.
Recall also that we identify a pair $\left(x^{*}, \xi^{*}\right) \in \mathbb{R}^{2 n}$ with the linear operator $l(x, \xi)=\left\langle x^{*}, x\right\rangle+\left\langle\xi^{*}, \xi\right\rangle$.

THEOREM 9.8 (Semiclassical form of Beals's Theorem). Let $A: \mathcal{S} \rightarrow \mathcal{S}^{\prime}$ be a continuous linear operator. Then
(i) $A=a^{\mathrm{w}}$ for a symbol $a \in S$
if and only if
(ii) for all $N=0,1,2, \ldots$ and all linear functions $l_{1}, \ldots, l_{N}$, we have

$$
\begin{equation*}
\left\|\operatorname{ad}_{l_{1}(x, h D)} \circ \cdots \circ \operatorname{ad}_{l_{N}(x, h D)} A\right\|_{L^{2} \rightarrow L^{2}}=O\left(h^{N}\right) \tag{9.25}
\end{equation*}
$$

Proof. 1. That (i) implies (ii) follows from the symbol calculus developed in Chapter 4. Indeed, $\|A\|_{L^{2} \rightarrow L^{2}}=O(1)$ and each commutator with an operator $l_{j}(x, h D)$ yields a term of order $h$.
2. That (ii) implies (i) is harder. First of all, the Schwartz Kernel Theorem (Theorem C.1) asserts that we can write

$$
\begin{equation*}
A u(x)=\int_{\mathbb{R}^{n}} K_{A}(x, y) u(y) d y \tag{9.26}
\end{equation*}
$$

for $K_{A} \in \mathcal{S}^{\prime}\left(\mathbb{R}^{n} \times \mathbb{R}^{n}\right)$. We call $K_{A}$ the kernel of $A$.
3. We now claim that if we define $a \in \mathcal{S}^{\prime}\left(\mathbb{R}^{2 n}\right)$ by

$$
\begin{equation*}
a(x, \xi):=\int_{\mathbb{R}^{n}} e^{-\frac{i}{h}\langle w, \xi\rangle} K_{A}\left(x+\frac{w}{2}, x-\frac{w}{2}\right) d w, \tag{9.27}
\end{equation*}
$$

then

$$
\begin{equation*}
K_{A}(x, y)=\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} a\left(\frac{x+y}{2}, \xi\right) e^{\frac{i}{h}\langle x-y, \xi\rangle} d \xi \tag{9.28}
\end{equation*}
$$

where the integrals are a shorthand for the Fourier transforms defined on $\mathcal{S}^{\prime}$.

To confirm this, we calculate that

$$
\begin{aligned}
& \frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} a\left(\frac{x+y}{2}, \xi\right) e^{\frac{i}{h}\langle x-y, \xi\rangle} d \xi \\
& \quad=\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}\langle x-y-w, \xi\rangle} K_{A}\left(\frac{x+y}{2}+\frac{w}{2}, \frac{x+y}{2}-\frac{w}{2}\right) d w d \xi \\
& \quad=K_{A}(x, y)
\end{aligned}
$$

since

$$
\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}\langle x-y-w, \xi\rangle} d \xi=\delta_{x-y}(w) \quad \text { in } \mathcal{S}^{\prime}
$$

In view of (9.26) and (9.28), we see that $A=a^{\mathrm{w}}$, for $a$ defined by (9.27).
4. Now we must show that $a$ belongs to the symbol class $S$; that is,

$$
\begin{equation*}
\sup _{\mathbb{R}^{2 n}}\left|\partial^{\alpha} a\right| \leq C_{\alpha} \tag{9.29}
\end{equation*}
$$

for each multiindex $\alpha$.

To do so we will make use of our hypothesis (9.25) with $l=x_{j}, \xi_{j}$, that is, with $l(x, h D)=x_{j}, h D_{j}$. We compute

$$
\begin{aligned}
\operatorname{Op}\left(h D_{\xi_{j}} a\right) & u(x) \\
= & \frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} h D_{\xi_{j}}\left(a\left(\frac{x+y}{2}, \xi\right)\right) e^{i \frac{\langle x-y, \xi\rangle}{h}} u(y) d \xi d y \\
= & -\frac{1}{(2 \pi h)} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} a\left(\frac{x+y}{2}, \xi\right) h D_{\xi_{j}}\left(e^{\frac{i\langle x-y, \xi\rangle}{h}}\right) u(y) d \xi d y \\
= & -\frac{1}{(2 \pi h)} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} a\left(\frac{x+y}{2}, \xi\right) e^{\frac{i\langle x-y, \xi\rangle}{h}}\left(x_{j}-y_{j}\right) u(y) d \xi d y \\
= & -\left[x_{j}, A\right] u=-\operatorname{ad}_{x_{j}} A u(x) .
\end{aligned}
$$

Likewise,

$$
\begin{aligned}
& \operatorname{Op}\left(h D_{x_{j}} a\right) u(x) \\
&=\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} a_{x_{j}}\left(\frac{x+y}{2}, \xi\right) e^{\frac{i\langle x-y, \xi\rangle}{h}} u(y) d \xi d y \\
&= \frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} h\left(D_{x_{j}}+D_{y_{j}}\right)\left(a\left(\frac{x+y}{2}, \xi\right)\right) \\
& \quad=\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} h D_{x_{j}}\left(a\left(\frac{x+y}{2}, \xi\right)\right) e^{\frac{i\langle x-y, \xi\rangle}{h}} u(y) d \xi d y \\
&+\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} a\left(\frac{x+y}{h}, \xi\right) e^{\frac{i\langle x-y, \xi\rangle}{h}} u(y) d \xi d y \\
&=h D_{x_{j}}(A u)-A\left(h D_{x_{j}} u\right) \\
&\left.=\left[h D_{x_{j}}, A\right] u=D_{y_{j}} u(y)\right) d \xi d y \\
& \operatorname{ad}_{x_{x_{j}}} A u(x) .
\end{aligned}
$$

In summary, for $j=1, \ldots, n$,

$$
\left\{\begin{array}{l}
\operatorname{ad}_{x_{j}} A=-\operatorname{Op}\left(h D_{\xi_{j}} a\right)  \tag{9.30}\\
\operatorname{ad}_{h D_{x_{j}}} A=\operatorname{Op}\left(h D_{x_{j}} a\right) .
\end{array}\right.
$$

5. Next we convert to the case $h=1$ by rescaling. For this, define

$$
U_{h} u(x):=h^{n / 4} u\left(h^{1 / 2} x\right)
$$

and check that $U_{h}: L^{2} \rightarrow L^{2}$ is unitary. Then

$$
U_{h} a^{\mathrm{w}}(x, h D) U_{h}^{-1}=a^{\mathrm{w}}\left(h^{1 / 2} x, h^{1 / 2} D\right)=\mathrm{Op}\left(a_{h}\right)
$$

for

$$
\begin{equation*}
a_{h}(x, \xi):=a\left(h^{1 / 2} x, h^{1 / 2} \xi\right) . \tag{9.31}
\end{equation*}
$$

Our hypothesis (9.25) is invariant under conjugation by $U_{h}$, and is consequently equivalent to

$$
\begin{equation*}
\operatorname{ad}_{l_{1}\left(h^{1 / 2} x, h^{1 / 2} D\right)} \circ \cdots \circ a_{l_{N}\left(h^{1 / 2} x, h^{1 / 2} D\right)} \operatorname{Op}\left(a_{h}\right)=O\left(h^{N}\right) . \tag{9.32}
\end{equation*}
$$

But since $l_{j}$ is linear, $l_{j}\left(h^{1 / 2} x, h^{1 / 2} D\right)=h^{1 / 2} l(x, D)$. Thus (9.32) is equivalent to

$$
\begin{equation*}
\operatorname{ad}_{l_{1}(x, D)} \circ \cdots \circ a_{l_{N}(x, D)} \circ \operatorname{Op}\left(a_{h}\right)=O\left(h^{N / 2}\right) . \tag{9.33}
\end{equation*}
$$

Taking $l_{k}(x, \xi)=x_{j}$ or $\xi_{j}$, it follows from (9.33) that

$$
\begin{equation*}
\left\|\operatorname{Op}\left(\partial^{\beta} a_{h}\right)\right\| \leq C h^{\frac{|\beta|}{2}} \tag{9.34}
\end{equation*}
$$

for all multiindices $\beta$.
6. Finally, we claim that

$$
\begin{equation*}
\left|\partial^{\alpha} a_{h}\right| \leq C_{\alpha} h^{|\alpha| / 2} \text { for each multiindex } \alpha . \tag{9.35}
\end{equation*}
$$

But this follows from Theorem 9.7, owing to estimate (9.34):

$$
\left\|a_{h}^{\alpha}\right\|_{L^{\infty}} \leq C \sum_{|\beta| \leq n+1}\left\|\operatorname{Op}\left(\partial^{\alpha+\beta} a_{h}\right)\right\|_{L^{2} \rightarrow L^{2}} \leq C_{\alpha} h^{|\alpha|} .
$$

Recalling (9.31), we rescale to derive the desired inequality (9.29).
EXAMPLE: resolvents. Suppose $a \in S$ is real-valued, so that $A=$ $a^{\mathrm{w}}(x, h D)$ is a self-adjoint operator on $L^{2}\left(\mathbb{R}^{n}\right)$. The resolvent $(A+i)^{-1}$ is then a bounded operator on $L^{2}\left(\mathbb{R}^{n}\right)$. Can we represent $(A+i)^{-1}$ as a pseudodifferential operator?

Since

$$
\operatorname{ad}_{l} B=-B\left(\operatorname{ad}_{l} A\right) B
$$

we see that the assumptions of Beals's Theorem are satisfied, and consequently $(A+i)^{-1}=b^{\mathrm{w}}(x, h D)$ for some symbol $b \in S$.

### 9.4 APPLICATION: EXPONENTIATION OF OPERATORS

In this section we will consider one parameter families of operators which give exponentials of self-adjoint pseudodifferential operators. As we have seen in Theorem 4.4, quantization of exponentiation commutes with quantization for linear symbols. This is of course not true for nonlinear symbols: see Section 10.2 for the subtleties involved in exponention of skew-adjoint pseudodifferential operators.

THEOREM 9.9 (Exponentials and order functions). Let $m=$ $m(x, \xi)$ be an order function and suppose that $g=g(x, \xi, h)$ satisfies

$$
\begin{equation*}
g(x, \xi)-\log m(x, \xi)=O(1) \tag{9.36}
\end{equation*}
$$

and

$$
\begin{equation*}
\partial^{\alpha} g \in S_{\delta} \tag{9.37}
\end{equation*}
$$

for $|\alpha|=1$ and $0 \leq \delta \leq \frac{1}{2}$.
Then the equation

$$
\left\{\begin{array}{l}
\frac{d}{d t} B(t)=g^{\mathrm{w}}(x, h D) B(t),  \tag{9.38}\\
B(0)=I,
\end{array}\right.
$$

has a unique solution $B(\cdot)$. Furthermore, for each $t \geq 0$, we have

$$
\begin{equation*}
B(t)=b_{t}^{\mathrm{w}}(x, h D) \tag{9.39}
\end{equation*}
$$

for a symbol

$$
b_{t} \in S\left(m^{t}\right) .
$$

Here $m^{t}$ means $m$ raised to the power $t$.
Theorem 9.9 identifies $\exp \left(\operatorname{tg}^{\mathrm{w}}(x, h D)\right)$ as a quantization of an element of $S_{\delta}\left(m^{t}\right)$. Thus we are asserting that on the level of order functions exponentiation commutes with quantization. Here is an application. Given an operator $P$, it is often very useful to consider its conjugations of the form

$$
P(t):=e^{-t g^{\mathrm{w}}(x, h D)} P e^{\operatorname{tg}^{\mathrm{w}}(x, h D)} .
$$

As an application of Theorem 9.9 we observe that if $g$ is as above and $P$ is bounded on $L^{2}$, then $P(t)$ is bounded on $L^{2}$. To see this, apply Theorems 4.13 and 4.16.

To prove Theorem 9.9 we start with
LEMMA 9.10 (Inverting exponentials). Consider

$$
U(t):=(\exp t g)^{\mathrm{w}}(x, D)
$$

as a mapping from $\mathcal{S}\left(\mathbb{R}^{n}\right)$ to itself. There exists $\epsilon_{0}=\epsilon_{0}(g)>0$ such that the operator $U(t)$ is invertible for $|t|<\epsilon_{0}$ and

$$
U(t)^{-1}=b_{t}^{\mathrm{w}}(x, D)
$$

for a symbol

$$
b_{t} \in S\left(m^{-t}\right)
$$

Proof: 1. We apply the composition formula given in Theorem 4.13 to obtain

$$
U(-t) U(t)=I+e_{t}^{\mathrm{w}}(x, D)
$$

where $e_{t} \in S$.
More explicitly, we write

$$
\begin{aligned}
e_{t}\left(x_{1}, \xi\right) & =\left.\int_{0}^{s} e^{s A(D)} A(D)\left(e^{-t g\left(x_{1}, \xi_{1}\right)+\operatorname{tg}\left(x_{2}, \xi_{2}\right)}\right)\right|_{x_{2}=x_{1}=x, \xi_{2}=\xi_{1}=\xi} d s \\
& =\left.\int_{0}^{s} i t e^{s A(D)} F e^{-\operatorname{tg}\left(x_{1}, \xi_{1}\right)+\operatorname{tg}\left(x_{2}, \xi_{2}\right)}\right|_{x_{2}=x_{1}=x, \xi_{2}=\xi_{1}=\xi} d s / 2,
\end{aligned}
$$

where

$$
\begin{gathered}
A(D)=i \sigma\left(D_{x_{1}}, D_{\xi_{1}} ; D_{x_{2}}, D_{\xi_{2}}\right) / 2, \\
F=\partial_{x_{1}} g\left(x_{1}, \xi_{1}\right) \cdot \partial_{\xi_{2}} g\left(x_{2}, \xi_{2}\right)-\partial_{\xi_{1}} g\left(x_{1}, \xi_{1}\right) \cdot \partial_{x_{2}} g\left(x_{2}, \xi_{2}\right) .
\end{gathered}
$$

2. It follows that $e_{t}=t \widetilde{e}_{t}$, for $\widetilde{e}_{t} \in S$. Therefore

$$
e_{t}^{\mathrm{w}}(x, D)=O_{L^{2} \rightarrow L^{2}}(t) ;
$$

this shows that $I+e_{t}^{\mathrm{w}}(x, D)$ is invertible for $|t|$ small enough. Then Theorem 9.8 implies

$$
\left(I+e_{t}^{\mathrm{w}}(x, D)\right)^{-1}=c_{t}^{\mathrm{w}}(x, D)
$$

for a symbol $c_{t} \in S$. Hence $b_{t}=c_{t} \# \exp (-t g(x, \xi)) \in S\left(m^{-t}\right)$.
Proof of Therem 9.9: 1. We first note that we only need to prove the result in the case $h=1$ by using the rescaling given in (4.29).
2. The hypotheses on $g$ in (9.36) are equivalent to the statement that $\exp (t g) \in S\left(m^{t}\right)$ for all $t \in \mathbb{R}$. We now observe that

$$
\begin{equation*}
\frac{d}{d t}\left(U(-t) \exp \left(t g^{\mathrm{w}}(x, D)\right)\right)=V(t) \exp \left(t g^{\mathrm{w}}(x, D)\right) \tag{9.40}
\end{equation*}
$$

where $V(t)=a_{t}^{\mathrm{w}}(x, D)$ and $a_{t} \in S\left(m^{-t}\right)$. In fact, we see that

$$
\begin{equation*}
\frac{d}{d t} U(-t)=-(g \exp (-t g))^{\mathrm{w}}(x, D) \tag{9.41}
\end{equation*}
$$

and

$$
\begin{equation*}
U(-t) g^{\mathrm{w}}(x, D)=(\exp (t g) \# g)^{\mathrm{w}}(x, D) \tag{9.42}
\end{equation*}
$$

As before, the composition formula (??) gives

$$
\begin{gathered}
\exp (-t g) \# g-g \exp (-t g)= \\
\int_{0}^{1} \exp (s A(D)) A(D) \exp \left(-\left.t g\left(x^{1}, \xi^{1}\right) g\left(x^{2}, \xi^{2}\right)\right|_{x^{1}=x^{2}=x, \xi^{1}=\xi^{2}=\xi},\right. \\
A(D)=i \sigma\left(D_{x^{1}}, D_{\xi^{1}} ; D_{x^{2}}, D_{\xi^{2}}\right) / 2
\end{gathered}
$$

From the hypothesis on $g$ we see that $A(D) \exp \left(t g\left(x^{1}, \xi^{1}\right)\right) g\left(x^{2}, \xi^{2}\right)$ is a sum of terms of the form $a\left(x^{1}, \xi^{1}\right) b\left(x^{2}, \xi^{2}\right)$ where $a \in S\left(m^{-t}\right)$ and $b \in S(1)$. The continuity of $\exp (A(D))$ on the spaces of symbols in Theorem ?? gives (9.40).
3. If we put

$$
C(t):=-V(t) U(-t)^{-1}
$$

then by Lemma 9.10, $C(t)=c_{t}^{\mathrm{w}}$ where $c_{t} \in S(1)$. Symbolic calculus shows that $c_{t}$ depends smoothly on $t$ and

$$
\left(\partial_{t}+C(t)\right)\left(U(-t) \exp \left(t g^{\mathrm{w}}(x, D)\right)\right)=0 .
$$

4. The proof of Theorem 9.9 is now reduced to showing

## LEMMA 9.11 (Solving an operator equation). Suppose that

$$
C(t)=c_{t}^{\mathrm{w}}(x, D),
$$

where the symbols $c_{t} \in S$ depends continuously on $t \in\left(-\epsilon_{0}, \epsilon_{0}\right)$.
Assume $q \in S$. Then the solution of

$$
\left\{\begin{align*}
\left(\partial_{t}+C(t)\right) Q(t) & =0,  \tag{9.43}\\
Q(0) & =q^{\mathrm{w}}(x, D)
\end{align*}\right.
$$

$i s$

$$
Q(t)=q_{t}^{\mathrm{w}}(x, D)
$$

where $q_{t} \in S$ depend continuously on $t \in\left(-\epsilon_{0}, \epsilon_{0}\right)$.
Proof: 1. The Picard existence theorem for ODE shows that $Q(t)$ is bounded on $L^{2}$. Assume now that $l_{j}(x, \xi)$ are linear functions on $\mathbb{R}^{n} \times \mathbb{R}^{n}$. Then

$$
\begin{aligned}
& \frac{d}{d t} \operatorname{ad}_{l_{1}(x, D)} \circ \cdots \circ \operatorname{ad}_{l_{N}(x, D)} Q(t)+ \\
& \quad \operatorname{ad}_{l_{1}(x, D)} \circ \cdots \circ \operatorname{ad}_{l_{N}(x, D)}(C(t) Q(t))=0, \\
& \operatorname{ad}_{l_{1}(x, D)} \circ \cdots \circ \operatorname{ad}_{l_{N}(x, D)} Q(0): L^{2}\left(\mathbb{R}^{n}\right) \rightarrow L^{2}\left(\mathbb{R}^{n}\right) .
\end{aligned}
$$

If we show that for any choice of $l_{j}^{\prime} s$ and any $N$

$$
\begin{equation*}
\operatorname{ad}_{l_{1}(x, D)} \circ \cdots \circ \operatorname{ad}_{l_{N}(x, D)} Q(t): L^{2}\left(\mathbb{R}^{n}\right) \rightarrow L^{2}\left(\mathbb{R}^{n}\right) \tag{9.44}
\end{equation*}
$$

then Beals's Theorem concludes the proof.
2 . We proceed by induction on $N$ :

$$
\begin{aligned}
& \operatorname{ad}_{l_{1}(x, D)} \circ \cdots \circ \operatorname{ad}_{l_{N}(x, D)}(C(t) Q(t))= \\
& \quad C(t) \operatorname{ad}_{l_{1}(x, D)} \circ \cdots \circ \operatorname{ad}_{l_{N}(x, D)} Q(t)+R(t),
\end{aligned}
$$

where $R(t)$ is the sum of terms of the form

$$
A_{k}(t) \operatorname{ad}_{l_{1}(x, D)} \circ \operatorname{ad}_{l_{k}(x, D)} Q(t)
$$

for $k<N, A_{k}(t)=a_{k}(t)^{\mathrm{w}}$, and $a_{k}(t) \in S$ depend continuously on $t$. This also follows by an inductive based on the derivation property of $a d_{l}$ :

$$
\operatorname{ad}_{l}(C D)=\left(\operatorname{ad}_{l} C\right) D+C\left(\operatorname{ad}_{l} D\right)
$$

Hence by the induction hypothesis $R(t)$ is bounded on $L^{2}$, and depends continuously on $t$. Thus

$$
\left(\partial_{t}+C(t)\right) \operatorname{ad}_{l_{1}(x, D)} \circ \cdots \circ \operatorname{ad}_{l_{N}(x, D)} Q(t)=R(t)
$$

is bounded on $L^{2}$. Since (9.44) is valid at $t=0$, we obtain it for all $t \in\left(-\epsilon_{0}, \epsilon_{0}\right)$.

### 9.5 INVARIANCE, HALF-DENSITIES

Invariance. We begin with a general discussion concerning the invariance of various quantities under the change of variables

$$
\begin{equation*}
\tilde{x}=\gamma(x), \tag{9.45}
\end{equation*}
$$

where $\gamma: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ is a diffeomorphism.
Functions. We note first that functions transform under (9.45) by pull-back. This means that we transform $u$ into a function of the new variables $\tilde{x}$ by the rule

$$
\begin{equation*}
\tilde{u}(\tilde{x})=\tilde{u}(\gamma(x)):=u(x) \tag{9.46}
\end{equation*}
$$

for $x \in \mathbb{R}^{n}$. Observe however that in general the integral of $u$ over a Borel set $E$ is not then invariant:

$$
\int_{\gamma(E)} \tilde{u}(\tilde{x}) d \tilde{x} \neq \int_{E} u(x) d x
$$

Densities. One way to repair this defect is to change our definition (9.46) to include the Jacobian of the transformation $\gamma$. We elegantly accomplish this by turning our attention to densities, which we denote symbolically as

$$
u(x)|d x| .
$$

We therefore modify our earlier definition (9.46), now to read

$$
\begin{equation*}
\tilde{u}(\tilde{x})=\tilde{u}(\gamma(x)):=u(x)|\operatorname{det}(\partial \gamma(x))|^{-1} . \tag{9.47}
\end{equation*}
$$

Then we have the invariance assertion

$$
" \tilde{u}(\tilde{x})|d \tilde{x}|=u(x)|d x| ",
$$

meaning that

$$
\int_{\gamma(E)} \tilde{u}(\tilde{x}) d \tilde{x}=\int_{E} u(x) d x
$$

for all Borel sets $E \subseteq \mathbb{R}^{n}$.

Half-densities. Next recall our general motivation coming from quantum mechanics. The eigenfunctions $u$ we study are then interpreted as wave functions and the squares of their moduli are the probability densities in the position representation: the probability of "finding our state in the set $E "$ is given by

$$
\int_{E}|u(x)|^{2} d x .
$$

This probability density should be invariantly defined, and so should not depend on the choice of coordinates $x$. As above, this means that it is not the function $u(x)$ which should be defined invariantly but rather the density $|u(x)|^{2} d x$, or up to the phase information, the half-density

$$
u(x)|d x|^{\frac{1}{2}} .
$$

For half-densities we therefore demand that

$$
" \tilde{u}(\tilde{x})|d \tilde{x}|^{\frac{1}{2}}=u(x)|d x|^{\frac{1}{2}} ",
$$

which means that integrals of the squares should be invariantly defined. To accomplish this, we once again modify our original definition (9.46), this time to become

$$
\begin{equation*}
\tilde{u}(\tilde{x})=\tilde{u}(\gamma(x)):=u(x)|\operatorname{det}(\partial \gamma(x))|^{-\frac{1}{2}} . \tag{9.48}
\end{equation*}
$$

Then

$$
\int_{\gamma(E)}|\tilde{u}(\tilde{x})|^{2} d \tilde{x}=\int_{E}|u(x)|^{2} d x .
$$

for all Borel subsets $E \subseteq \mathbb{R}^{n}$.
DISCUSSION. The foregoing formalism is at first rather unintuitive, but turns out later to play a crucial role in the rigorous semiclassical calculus, in particular in the theory of Fourier integral operators, which we will touch upon later. Section 9.2 will demonstrate how the halfdensity viewpoint fits naturally within the Weyl calculus, and Section 10.2 will explain how half-densities simplify some related calculations for a propagator.

Our Appendix E provides a more careful foundation of these concepts in terms of the s-density line bundles over $\mathbb{R}^{n}$, denoted $\Omega^{s}\left(\mathbb{R}^{n}\right)$. In this notation, a density is a smooth section of $\Omega^{1}\left(\mathbb{R}^{n}\right)$ and a half-density is a smooth section of $\Omega^{\frac{1}{2}}\left(\mathbb{R}^{n}\right)$.

We therefore write

$$
u|d x| \in C^{\infty}\left(\mathbb{R}^{n}, \Omega^{1}\left(\mathbb{R}^{n}\right)\right)
$$

for densities, and

$$
u|d x|^{\frac{1}{2}} \in C^{\infty}\left(\mathbb{R}^{n}, \Omega^{\frac{1}{2}}\left(\mathbb{R}^{n}\right)\right)
$$

for half-densities.
REMARK: Half density operator kernels. Half-densities elegantly appear when we use introduce operator kernels. Suppose that

$$
K \in C^{\infty}\left(\mathbb{R}^{n} \times \mathbb{R}^{n} ; \Omega^{\frac{1}{2}}\left(\mathbb{R}^{n} \times \mathbb{R}^{n}\right)\right)
$$

Then $K$, acting as an integral kernel, defines a map

$$
K: C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{n}, \Omega^{\frac{1}{2}}\left(\mathbb{R}^{n}\right)\right) \rightarrow C^{\infty}\left(\mathbb{R}^{n}, \Omega^{\frac{1}{2}}\left(\mathbb{R}^{n}\right)\right)
$$

in an invariant way, independently of the choice of densities:

$$
\begin{align*}
K u(x)|d x|^{\frac{1}{2}}= & \int_{\mathbb{R}^{n}} K(x, y)|d x|^{\frac{1}{2}}|d y|^{\frac{1}{2}} u(y)|d y|^{\frac{1}{2}} \\
& :=\left(\int_{\mathbb{R}^{n}} K(x, y) u(y) d y\right)|d x|^{\frac{1}{2}} \tag{9.49}
\end{align*}
$$

### 9.6 CHANGING VARIABLES

In this section we illustrate the usefulness of half-densities in charterizing invariance properties of quantization under changes of variables.

When we fix the symplectic form $\sigma=d \xi \wedge d x$ on $\mathbb{R}^{2 n}=\mathbb{R}^{n} \times \mathbb{R}^{n}$, the half-density sections over $\mathbb{R}^{n 2}$ are identified with functions using the canononical density

$$
d \xi_{1} \wedge \cdots \wedge d \xi_{n} \wedge d x_{1} \wedge \cdots \wedge d x_{n}=\frac{1}{n!} \sigma^{n}
$$

In other words, half-densities transform as functions under symplectic changes of variables, and in particular for symplectic transformations arising as in Example 1 of Section 2.3:

$$
\begin{equation*}
(x, \xi) \mapsto\left(\gamma(x),\left(\partial \gamma(x)^{T}\right)^{-1} \xi\right) . \tag{9.50}
\end{equation*}
$$

We will consider the Weyl quantization of a symbol $a \in \mathcal{S}\left(\mathbb{R}^{2 n}\right)$ as an operator acting on half-densities. That is done as in (9.49) by defining

$$
\begin{align*}
& K_{a}(x, y)|d x|^{\frac{1}{2}}|d y|^{\frac{1}{2}} \\
& \quad:=\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} a\left(\frac{x+y}{2}, \xi\right) e^{i \frac{i x-y, \xi\rangle}{h}} d \xi|d x|^{\frac{1}{2}}|d y|^{\frac{1}{2}} \tag{9.51}
\end{align*}
$$

and

$$
\begin{equation*}
\operatorname{Op}(a)\left(u|d y|^{\frac{1}{2}}\right)|d x|^{\frac{1}{2}}:=\int_{\mathbb{R}^{n}} K_{a}(x, y) u(y) d y|d x|^{\frac{1}{2}} \tag{9.52}
\end{equation*}
$$

The arguments of Chapter 4 show that for $a \in \mathcal{S}$ we obtain a bounded operator that quantizes $a$ :

$$
\operatorname{Op}(a): L^{2}\left(\mathbb{R}^{n}, \Omega^{\frac{1}{2}}\left(\mathbb{R}^{n}\right)\right) \rightarrow L^{2}\left(\mathbb{R}^{n}, \Omega^{\frac{1}{2}}\left(\mathbb{R}^{n}\right)\right)
$$

Next, let $\gamma: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ be a smooth diffeomorphism which for simplicity we assume to be the identity outside a compact set. Take $a \in \mathcal{S}$. We write $A=\operatorname{Op}(a)$ for the operator acting on half-densities. As above, if we write $\tilde{x}=\gamma(x)$, we define $\tilde{u}$ by

$$
\begin{equation*}
\tilde{u}(\tilde{x})|d \tilde{x}|^{\frac{1}{2}}=u(x)|d x|^{\frac{1}{2}} . \tag{9.53}
\end{equation*}
$$

Then $\tilde{A}=\left(\gamma^{-1}\right)^{*} A \gamma^{*}$, acting on half-densities, is given by the rule

$$
\begin{equation*}
\tilde{A} \tilde{u}(\tilde{x})=A u(x), \tag{9.54}
\end{equation*}
$$

when acting on functions.

THEOREM 9.12 (Operators and half-densities). Let $a \in \mathcal{S}\left(\mathbb{R}^{2 n}\right)$ and let $A$ be its quantization acting on half-densities.
(i) Then

$$
\begin{equation*}
\left(\gamma^{-1}\right)^{*} A \gamma^{*}=\operatorname{Op}(\tilde{a}) \tag{9.55}
\end{equation*}
$$

for the symbol

$$
\begin{equation*}
\tilde{a}(x, \xi):=a\left(\gamma^{-1}(x), \partial \gamma(x)^{T} \xi\right)+O_{\mathcal{S}}\left(h^{2}\right) . \tag{9.56}
\end{equation*}
$$

That is,

$$
\begin{equation*}
a(x, \xi)=\tilde{a}\left(\gamma(x),\left(\partial \gamma(x)^{T}\right)^{-1} \xi\right)+O_{\mathcal{S}}\left(h^{2}\right) \tag{9.57}
\end{equation*}
$$

(ii) Consider $A$ acting on functions and define

$$
A_{1}=\left(\gamma^{-1}\right)^{*} A \gamma^{*},
$$

then

$$
A_{1}=\mathrm{Op}\left(a_{1}\right)
$$

for the symbol

$$
\begin{equation*}
a_{1}(x, \xi):=a\left(\gamma^{-1}(x), \partial \gamma(x)^{T} \xi\right)+O_{\mathcal{S}}(h) . \tag{9.58}
\end{equation*}
$$

That is,

$$
\begin{equation*}
a(x, \xi)=a_{1}\left(\gamma(x),\left(\partial \gamma(x)^{T}\right)^{-1} \xi\right)+O_{\mathcal{S}}(h) \tag{9.59}
\end{equation*}
$$

INTERPRETATION. A further motivation for half-densities is that assertion (i) for half-densities (with error term of order $O_{\mathcal{S}}\left(h^{2}\right)$ ) is more precise than the assertion (ii) for functions (with error term $O_{\mathcal{S}}(h)$ ). The notation $b=O_{\mathcal{S}}(h)$ means that

$$
\left|x^{\alpha} \xi^{\beta} \partial^{\gamma} b\right| \leq C_{\alpha \beta, \gamma} h
$$

for all multiindices $\alpha, \beta, \gamma$ : see $\S$ A. 5 .
Proof. 1. Since $a \in \mathcal{S}$, we have $K_{a} \in \mathcal{S}$. Take $\tilde{a} \in \mathcal{S}\left(\mathbb{R}^{2 n}\right)$ for which $\left(\kappa^{-1}\right)^{*} A \kappa^{*}=\operatorname{Op}(\tilde{a})$.
2. Remember that

$$
A u(x)|d x|^{\frac{1}{2}}=\int_{\mathbb{R}^{n}} K_{a}(x, y)|d x|^{\frac{1}{2}}|d y|^{\frac{1}{2}} u(y)|d y|^{\frac{1}{2}}
$$

for

$$
K_{a}(x, y):=\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} a\left(\frac{x+y}{2}, \xi\right) e^{\frac{i}{h}\langle x-y, \xi\rangle} d \xi
$$

Likewise

$$
\tilde{A} \tilde{u}(\tilde{x})=\int_{\mathbb{R}^{n}} K_{\tilde{a}}(\tilde{x}, \tilde{y})|d \tilde{x}|^{\frac{1}{2}}|d \tilde{y}|^{\frac{1}{2}} \tilde{u}(\tilde{y})|d \tilde{y}|^{\frac{1}{2}}
$$

for

$$
K_{\tilde{a}}(\tilde{x}, \tilde{y}):=\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \tilde{a}\left(\frac{\tilde{x}+\tilde{y}}{2}, \tilde{\xi}\right) e^{\frac{i}{h}\langle\tilde{x}-\tilde{y}, \tilde{\xi}\rangle} d \tilde{\xi}
$$

Since $\tilde{u}(\tilde{y})|d \tilde{y}|^{\frac{1}{2}}=u(y)|d y|^{\frac{1}{2}}$ and $d \tilde{y}=|\operatorname{det} \partial \gamma(y)| d y$, it follows that

$$
\tilde{A} \tilde{u}(x)=\int_{\mathbb{R}^{n}} K_{\tilde{a}}(\tilde{x}, \tilde{y})|\operatorname{det} \partial \gamma(y)|^{\frac{1}{2}}|\operatorname{det} \partial \gamma(x)|^{\frac{1}{2}} u(y) d y .
$$

Hence

$$
\begin{equation*}
K_{a}(x, y)=K_{\tilde{a}}(\tilde{x}, \tilde{y})|\operatorname{det} \partial \gamma(y)|^{\frac{1}{2}}|\operatorname{det} \partial \gamma(x)|^{\frac{1}{2}} . \tag{9.60}
\end{equation*}
$$

3. Now

$$
\begin{aligned}
K_{\tilde{a}}(\tilde{x}, \tilde{y}) & =\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \tilde{a}\left(\frac{\tilde{x}+\tilde{y}}{2}, \tilde{\xi}\right) e^{\frac{i}{h}\langle\tilde{x}-\tilde{y}, \tilde{\xi}\rangle} d \tilde{\xi} \\
& =\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \tilde{a}\left(\frac{\gamma(x)+\gamma(y)}{2} \tilde{\xi}\right) e^{\frac{i}{h}\langle\gamma(x)-\gamma(y), \tilde{\xi}\rangle} d \tilde{\xi}
\end{aligned}
$$

We have

$$
\begin{equation*}
\gamma(x)-\gamma(y)=\langle g(x, y), x-y\rangle \tag{9.61}
\end{equation*}
$$

where

$$
\begin{equation*}
g(x, y)=\partial \gamma\left(\frac{x+y}{2}\right)+O\left(|x-y|^{2}\right) . \tag{9.62}
\end{equation*}
$$

Also

$$
\begin{equation*}
\gamma(x)+\gamma(y)=2 \gamma\left(\frac{x+y}{2}\right)+O\left(|x-y|^{2}\right) \tag{9.63}
\end{equation*}
$$

Let us also write

$$
\begin{equation*}
\tilde{\xi}=\left(g(x, y)^{T}\right)^{-1} \xi \tag{9.64}
\end{equation*}
$$

Substituting above, we deduce that

$$
\begin{aligned}
& K_{\tilde{a}}(\tilde{x}, \tilde{y}) \\
& =\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}}\left[\tilde{a}\left(\gamma\left(\frac{x+y}{2}\right),\left(g(x, y)^{T}\right)^{-1} \xi\right)+O\left(|x-y|^{2}\right)\right] \\
& \quad e^{\frac{i}{h}\left\langle\kappa(x)-\kappa(y),\left(g(x, y)^{T}\right)^{-1} \xi\right\rangle} d \tilde{\xi} .
\end{aligned}
$$

We now use the so-called "Kuranishi trick" to rewrite this expression as a pseudodifferential operator. First,

$$
\left\langle\gamma(x)-\gamma(y),\left(g(x, y)^{T}\right)^{-1} \xi\right\rangle=\left\langle g(x, y)^{-1}(\gamma(x)-\gamma(y)), \xi\right\rangle=\langle x-y, \xi\rangle
$$

according to (9.61). Remembering also (9.62), we compute

$$
\begin{aligned}
& K_{\tilde{a}}(\tilde{x}, \tilde{y})= \\
& \frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}}\left[\tilde{a}\left(\gamma\left(\frac{x+y}{2}\right),\left(\partial \gamma\left(\frac{x+y}{2}\right)^{T}\right)^{-1} \xi\right)+O\left(|x-y|^{2}\right)\right] e^{\frac{i}{h}\langle x-y, \xi\rangle} d \tilde{\xi} \\
& \quad=\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}}\left[a\left(\frac{x+y}{2}, \xi\right)+O\left(|x-y|^{2}\right)\right] e^{\frac{i}{h}\langle x-y, \xi\rangle} d \tilde{\xi}
\end{aligned}
$$

Furthermore $d \tilde{\xi}=|\operatorname{det} g(x, y)|^{-1} d \xi$ and

$$
\operatorname{det} g(x, y)=\operatorname{det} \partial \gamma\left(\frac{x+y}{2}\right)+O\left(|x-y|^{2}\right)
$$

Also, we claim that

$$
\left|\operatorname{det} \partial \gamma\left(\frac{x+y}{2}\right)\right|^{2}=|\operatorname{det} \partial \gamma(x)||\operatorname{det} \partial \gamma(y)|+O\left(|x-y|^{2}\right) .
$$

This identity is clear if we add a term $\left\langle A\left(\frac{x+y}{2}\right), x-y\right\rangle$ on the right hand side. But the symmetry under switching $x$ and $y$ shows that $A \equiv 0$.
4. Finally we observe that

$$
\begin{equation*}
(x-y)^{\alpha} e^{\frac{i}{h}\langle x-y, \xi\rangle}=\left(h D_{\xi}\right)^{\alpha} e^{\frac{i}{h}\langle x-y, \xi\rangle} . \tag{9.65}
\end{equation*}
$$

Hence integrating by parts in the terms with $O\left(|x-y|^{2}\right)$ gives us terms of order $O\left(h^{2}\right)$. So

$$
\begin{aligned}
& K_{\tilde{a}}(\tilde{x}, \tilde{y})= \\
& \frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}}\left(a\left(\frac{x+y}{2}, \xi\right)+O\left(h^{2}\right)\right) e^{\frac{i}{h}\langle x-y, \xi\rangle} d \xi \\
& \quad|\operatorname{det} \partial \gamma(x)|^{-1 / 2}|\operatorname{det} \partial \gamma(y)|^{-1 / 2} .
\end{aligned}
$$

This proves (9.60) with $\tilde{a}$ satisfying (9.56).
5. When $A$ acts on functions, then $K_{a}$ has to transform as a density. In other words, we need to show

$$
\begin{equation*}
K_{a}(x, y)=K_{a_{1}}(\tilde{x}, \tilde{y})|\operatorname{det} \partial \gamma(y)|+O(h) \tag{9.66}
\end{equation*}
$$

instead of (9.60). Since
$|\operatorname{det} \partial \gamma(y)|=|\operatorname{det} \partial \gamma(y)|^{1 / 2}|\operatorname{det} \partial \gamma(x)|^{1 / 2}+O(|x-y|)$,
we see from (9.65) that (9.66) follows from (9.60) with $a_{1}=\tilde{a}+O(h)$.

### 9.7 NEW SYMBOL CLASSES

Suppose that

$$
\gamma: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}
$$

is a diffeomorphism equal to the identity outside of a compact set. In $\S 9.6$ we discussed the operator $\left(\gamma^{-1}\right)^{*} \operatorname{Op}(a) \gamma^{*}$ for $a \in \mathcal{S}\left(\mathbb{R}^{n}\right)$. We want now to introduce more general classes of symbols $a$ than those provided by Chapter 4 , and to discuss their invariance under the mapping $\kappa$.
DISCUSSION. To motivate the need for these new classes, let $m$ be an order function and recall the class $S(m)$ introduced in Chapter 4:

$$
S(m):=\left\{a \in C^{\infty}\left(\mathbb{R}^{2 n}\right)| | \partial^{\alpha} a(x, \xi) \mid \leq C_{\alpha} m(x, \xi)\right\} .
$$

In view of Theorem 9.12, for a symbol $a$ in $S(m)$ to be invariant under $\gamma$, we would need

$$
\begin{equation*}
\left|\partial^{\alpha}\left(a\left(\gamma^{-1}(x), \partial \gamma(x)^{T} \xi\right)\right)\right| \leq C_{\alpha} m(x, \xi) . \tag{9.67}
\end{equation*}
$$

Since differentiation in $x$ falling on the second set of variables produces factors of the form

$$
\partial_{x_{j}}\left(\partial \gamma(x)^{T} \xi\right),
$$

the bound (9.67) is in general false unless

$$
|m(x, \xi)| \leq C_{N}\langle\xi\rangle^{-N}
$$

for all $N$. This requirement is clearly too restrictive, as it would exclude all differential operators.

An estimate of the type (9.67) would however hold if differentiation in $\xi$ improves the decay in $\xi$. This observation leads us to the following definition, in which we restrict to the simplest order functions $\langle\xi\rangle^{m}$ :

DEFINITIONS. (i) The classical symbols are

$$
\begin{equation*}
S^{m, k}=\left\{a \in C^{\infty}\left(\mathbb{R}^{2 n}\right)| | \partial_{x}^{\alpha} \partial_{\xi}^{\beta} a \mid \leq C_{\alpha \beta} h^{-k}\langle\xi\rangle^{m-|\beta|}\right\} . \tag{9.68}
\end{equation*}
$$

(ii) We also write

$$
\Psi^{m, k}=\left\{a^{\mathrm{w}}(x, h D) \mid a \in S^{m, k}\right\}
$$

to denote the corresponding class of operators obtained by the quantization procedure described in Theorem 4.10.

Since $S^{m, k} \subset S^{k}\left(\langle\xi\rangle^{m}\right)$ all the results of Chapter 4 are applicable but due to the improvement under differentiation in $\xi$ there are many new features, important in the study of partial differential equations. We will only present the improved composition formula and the change of variables formula.

THEOREM 9.13 (Composition for $\Psi^{m, k}$ ). Suppose that $a \in S^{m_{1}, k_{1}}$ and $b \in S^{m_{2}, k_{2}}$. Then

$$
a^{\mathrm{w}}(x, h D) \circ b^{\mathrm{w}}(x, h D)=c^{\mathrm{w}}(x, h D)
$$

where $c \in S^{m_{1}+m_{2}, k_{1}+k_{2}}$ is given by (??).
Moreover,

$$
\begin{align*}
& c(x, \xi)= \\
& \qquad \begin{array}{l}
\left.\sum_{k=0}^{N} \frac{1}{k!}\left(\frac{i h}{2} \sigma\left(D_{x}, D_{\xi}, D_{y}, D_{\eta}\right)\right)^{k} a(x, \xi) b(y, \eta)\right|_{x=\xi, y=\eta} \\
\\
+O_{S^{m}+m_{2}-N-1, k_{1}+k_{2}}\left(h^{N+1}\right)
\end{array} \tag{9.69}
\end{align*}
$$

where

$$
\sigma\left(D_{x}, D_{\xi}, D_{y}, D_{\eta}\right):=\left\langle D_{\xi}, D_{y}\right\rangle-\left\langle D_{x}, D_{\eta}\right\rangle .
$$

REMARK. Similar statements hold for the usual quantization:

$$
a(x, h D) \circ b(x, h D)=c_{1}(x, h D)=\left.e^{i\left\langle\left\langle D_{\xi}, D_{y}\right\rangle\right.} a(x, \xi) b(y, \eta)\right|_{y=x, \eta=\xi}
$$

and

$$
c_{1}(x, \xi)=\sum_{|\alpha| \leq N} \frac{1}{\alpha!} \partial_{\xi}^{\alpha} a(x, \xi)\left(h D_{x}\right)^{\alpha} b(x, \xi)+O_{S^{m_{1}+m_{2}-N-1, k_{1}+k_{2}}}\left(h^{N+1}\right)
$$

Proof: 1. To simplify the notation let us make the harmless assumption that $k_{1}=k_{2}=0$. Since $S^{m_{j}, 0} \subset S\left(\langle\xi\rangle^{m_{j}}\right)$ the validity of (??) follows from Theorem 4.6. Similarly, (9.69) is valid but with an error $O_{S\left(\langle\xi\rangle^{m_{1}+m_{2}}\right)}\left(h^{N+1}\right)$.
2. Now we observe that

$$
\left.\left(\frac{i h}{2} \sigma\left(D_{x}, D_{\xi}, D_{y}, D_{\eta}\right)\right)^{k} a(x, \xi) b(y, \eta)\right|_{x=\xi, y=\eta} \in S^{m_{1}+m_{2}-k,-k}
$$

and we will show that the remainder satisfies

$$
\begin{gather*}
c(x, \xi)-\left.\sum_{k=0}^{N} \frac{1}{k!}\left(\frac{i h}{2} \sigma\left(D_{x}, D_{\xi}, D_{y}, D_{\eta}\right)\right)^{k} a(x, \xi) b(y, \eta)\right|_{x=\xi, y=\eta}  \tag{9.70}\\
\in S^{N+1}\left(\langle\xi\rangle^{m_{1}+m_{2}-N-1}\right)
\end{gather*}
$$

Since $N$ is arbitrary it follows that

$$
\left|\partial_{x}^{\alpha} \partial_{\xi}^{\beta} c(x, \xi)\right| \leq C_{\alpha, \beta}\langle\xi\rangle^{m_{1}+m_{2}-|\beta|}
$$

uniformly for $0<h \leq 1$.
3. To check (9.70) we recall that (??) (and hence (9.69)) was proved by consider the action of $\exp A(D)$, where $A$ is a quadratic form, on symbol classes - see Theorem ??. Using Taylor's formula we write $\exp (i h A(D))=$

$$
\sum_{k=0}^{N} \frac{(i h A(D))^{k}}{k!}+\frac{1}{N!} \int_{0}^{1}(1-t)^{N} \exp (i t h A(D))(i h A(D))^{N+1} d t
$$

In our case we have $A(D)=h \sigma\left(D_{x}, D_{\xi} ; D_{y}, D_{\eta}\right) / 2$, so that

$$
c(x, \xi)=\left.\exp (i h A(D)) a(x, \xi) b(y, \eta)\right|_{x=y, \eta=\xi}
$$

and the remainder in (9.70) is given by

$$
\left.\frac{1}{N!} \int_{0}^{1}(1-t)^{N} \exp (i t h A(D))(i h A(D))^{N+1}(a(x, \xi) b(y, \eta))\right|_{x=y, \xi=\eta} d t
$$

Since $a \in S^{m_{1}, 0}$ and $b \in S^{m_{2}, 0}$, we see that

$$
(h A(D))^{N+1}(a(x, \xi) b(y, \eta)) \in \sum_{k=0}^{N+1} S^{N+1}\left(\langle\xi\rangle^{m_{1}-k}\langle\eta\rangle^{m_{2}-N-1+k}\right) .
$$

Theorem ?? shows that $\exp (i t h A(D))$ :

$$
S^{N+1}\left(\langle\xi\rangle^{m_{1}-k}\langle\eta\rangle^{m_{2}-N-1+k}\right) \longrightarrow S^{N+1}\left(\langle\xi\rangle^{m_{1}-k}\langle\eta\rangle^{m_{2}-N-1+k}\right),
$$

with bounds uniform for $0 \leq t \leq 1,0<h \leq 1$. Also,

$$
\begin{gathered}
\left.S^{N+1}\left(\mathbb{R}_{(x, \xi)}^{2 n} \times \mathbb{R}_{(y, \eta)}^{2 n},\langle\xi\rangle^{m_{1}-k}\langle\eta\rangle^{m_{2}-N-1+k}\right)\right|_{x=y, \xi=\eta}= \\
S^{N+1}\left(\mathbb{R}_{(x, \xi)}^{2 n},\langle\xi\rangle^{m_{1}+m_{2}-N-1}\right),
\end{gathered}
$$

from which (9.70) follows.
We also record the following useful lemma:

## LEMMA 9.14 (Schwartz kernels of operators in $\Psi^{m, k}$ ).

(i) Suppose that $a \in S^{m, k}$ and that $K_{a} \in \mathcal{S}^{\prime}\left(\mathbb{R}^{n} \times \mathbb{R}^{n}\right)$ is the Schwartz kernel of $\mathrm{Op}(a)$. Then

$$
\begin{equation*}
K_{a}(x, y) \in C^{\infty}\left(\mathbb{R}^{n} \times \mathbb{R}^{n} \backslash \Delta\right) \tag{9.71}
\end{equation*}
$$

for the diagonal $\Delta:=\left\{(x, x): x \in \mathbb{R}^{n}\right\}$. Furthermore, we have the estimates

$$
\begin{equation*}
\left|\left(h D_{x}\right)^{\alpha}\left(h D_{y}\right)^{\beta} K_{a}(x, y)\right| \leq C_{N} h^{-k}\left(\frac{h}{|x-y|}\right)^{N} \tag{9.72}
\end{equation*}
$$

for $(x, y) \in \mathbb{R}^{n} \times \mathbb{R}^{n} \backslash \Delta$ and $N>|\alpha|+|\beta|+m+n$.
(ii) If $K_{a}$ satisfies

$$
\begin{equation*}
\left|\partial_{x}^{\alpha} \partial_{y}^{\beta} K_{a}(x, y)\right| \leq C_{N}\left(\frac{h}{\langle x-y\rangle}\right)^{N} \tag{9.73}
\end{equation*}
$$

for all $(x, y) \in \mathbb{R}^{n} \times \mathbb{R}^{n}$, then $K_{a}$ is the Schwartz kernel of $\operatorname{Op}(a)$ for $a \in S^{-\infty,-\infty}$.

The lemma shows one of the many advantages of symbol classes $S^{m, k}$ since we now have smoothness and rapid decay away from the diagonal.
Proof. 1. We can consider either the Weyl quantization, Op, or the standard quantization, $\mathrm{Op}_{1}$. That follows from Theorem ?? since as in Part 3 of the proof of Theorem 9.13 we see that

$$
\begin{equation*}
\exp \left(i(t-s) h\left\langle D_{x}, D_{\xi}\right\rangle\right): S^{m, k} \longrightarrow S^{m, k} \tag{9.74}
\end{equation*}
$$

So for simplicity of notation we opt for $\mathrm{Op}_{1}(a)$.
2. Suppose first that $a \in \mathcal{S}$ so that

$$
K_{a}(x, y)=\frac{1}{(2 \pi h)^{n}} \int a(x, \xi) e^{i\langle x-y, \xi\rangle / h} d \xi
$$

and the integral is taken in the usual sense. We note that

$$
(x-y)^{\gamma} K_{a}(x, y)=\frac{1}{(2 \pi h)^{n}} \int\left(-h D_{\xi}\right)^{\gamma} a(x, \xi) e^{i\langle x-y, \xi\rangle / h} d \xi
$$

and hence

$$
\left|(x-y)^{\gamma} K_{a}(x, y)\right| \leq C_{\gamma} h^{|\gamma|}\left\|\langle\xi\rangle^{n+1} \partial_{\xi}^{\gamma} a\right\|_{\infty} \int\langle\xi\rangle^{-n-1} d \xi
$$

Observe next that

$$
\sup _{|\gamma|=N}\left|(x-y)^{\gamma}\right| \geq n^{-\frac{N}{2}}|x-y|^{N}
$$

(It is enough to prove this inequality for $y=0,|x|=1$. It then says that we can choose $\gamma, \gamma_{1}+\cdots \gamma_{n}=N$, so that $\left|x_{1}\right|^{\gamma_{1}} \cdots\left|x_{n}\right|^{\gamma_{n}} \geq n^{-N / 2}$
for any $x_{1}^{2}+\cdots x_{n}^{2}=1$. The last condition implies that there exists $m$, $1 \leq m \leq n$ such that $\left|x_{m}\right| \geq 1 / \sqrt{n}$. We then take $\gamma_{j}=\delta_{j m} N$.)

Therefore we obtain

$$
\begin{equation*}
\left|K_{a}(x, y)\right| \leq C_{N}\left(\frac{h}{|x-y|}\right)^{N} \sup _{|\gamma|=N}\left\|\langle\xi\rangle^{n+1} \partial_{\xi}^{\gamma} a\right\|_{\infty} \tag{9.75}
\end{equation*}
$$

Similarly,

$$
\begin{gathered}
\left|\left(h D_{x}\right)^{\alpha}\left(h D_{y}\right)^{\beta}\left((x-y)^{\gamma} K_{a}(x, y)\right)\right| \leq \\
C_{\gamma, \alpha, \beta} \sup _{|\rho| \leq|\alpha|} h^{|\gamma|}\left\|\langle\xi\rangle^{n+1+|\alpha|+|\beta|} \partial_{\xi}^{\gamma} \partial_{x}^{\rho} a\right\|_{\infty}
\end{gathered}
$$

and consequently,

$$
\begin{gather*}
\left|\left(h D_{x}\right)^{\alpha}\left(h D_{y}\right)^{\beta} K_{a}(x, y)\right| \leq \\
C_{N, \alpha, \beta}\left(\frac{h}{|x-y|}\right)^{N} \sup _{|\rho| \leq|\alpha|,|\gamma|=N}\left\|\langle\xi\rangle^{n+1+|\alpha|+|\beta|} \partial_{x}^{\rho} \partial_{\xi}^{\gamma} a\right\|_{\infty} . \tag{9.76}
\end{gather*}
$$

3. If $a \in S^{m, k}$ we observe that seminorms appearing on the right hand side of (9.76) are finite and bounded by $h^{-k}$ if $N>|\alpha|+|\beta|+m+n$. Approximation by symbols in $\mathcal{S}$ concludes the proof of (??).
4. The second conclusion follows from applying the inverse Fourier transform:

$$
a(x, \xi)=\int K_{a}(x, x-z) e^{i\langle z, \xi\rangle / h} d z
$$

Since

$$
\left|\partial_{x}^{\alpha} \partial_{z}^{\beta} K_{a}(x, x-z)\right| \leq C_{N} \frac{h^{N}}{\langle z\rangle^{N}}
$$

we see that $a \in S^{-\infty,-\infty}$.

Before restating Theorem 9.12 in this more general setting we will discuss usual quantization acting on functions. Suppose first that $a \in$ $\mathcal{S}$. Then

$$
\begin{equation*}
\left(\gamma^{-1}\right)^{*} \mathrm{Op}_{1}(a) \gamma^{*}=\mathrm{Op}_{1}\left(a_{\gamma}\right), \tag{9.77}
\end{equation*}
$$

where

$$
\begin{equation*}
a_{\gamma}(\gamma(x), \eta)=e^{-\frac{i}{h}\langle\gamma(x), \eta\rangle} a(x, h D) e^{\frac{i}{h}\langle\gamma(\cdot), \eta\rangle} \tag{9.78}
\end{equation*}
$$

In fact, Theorem ?? shows that

$$
\begin{aligned}
a_{\gamma}(y, \eta) & =e^{-\frac{i}{h}\langle y, \eta\rangle} \mathrm{Op}_{1}\left(a_{\gamma}\right) e^{\frac{i}{h}\langle y, \eta\rangle} \\
& =e^{-\frac{i}{h}\langle y, \eta\rangle}\left(\gamma^{-1}\right)^{*} a(x, h D) \gamma^{*}\left(e^{-\frac{i}{h}\langle\cdot, \eta\rangle}\right)
\end{aligned}
$$

which is the same as (9.78).

THEOREM 9.15 (Changing variables I). Suppose that $a \in S^{m, k}$ where $S^{m, k}$ is defined by (9.68). Then (9.77) defines $a_{\gamma} \in S^{m, k}$ for which (9.78) holds. Moreover,

$$
\begin{align*}
a_{\gamma}(\gamma(x), \eta)=\sum_{|\alpha| \leq N} & \left.\frac{1}{\alpha!} \partial_{\xi}^{\alpha} a\left(x, \partial \gamma(x)^{T} \eta\right)\left(h D_{y}\right)^{\alpha} e^{\frac{i}{h}\left\langle\rho_{x}(y), \eta\right\rangle}\right|_{y=x}  \tag{9.79}\\
& +O_{S^{m-N-1, k}}\left(h^{N+1}\right)
\end{align*}
$$

where

$$
\rho_{x}(y)=\gamma(y)-\gamma(x)-\partial \gamma(x)(y-x) .
$$

In particular,

$$
\begin{equation*}
a_{\gamma}(\gamma(x), \eta)=a\left(x, \partial \gamma(x)^{T} \eta\right)+O_{S^{m-1, k}}(h) . \tag{9.80}
\end{equation*}
$$

Proof. 1. We need to show is that $a_{\gamma}$ defined by (9.78) is in $S^{m, k}$. Since that will imply

$$
\left(\gamma^{-1}\right)^{*} \mathrm{Op}_{1}(a) \gamma^{*} u=\mathrm{Op}_{1}\left(a_{\gamma}\right) u, \quad u(x)=\exp (i\langle x, \eta\rangle / h),
$$

the operator identity (9.77) will follow since $\exp (i\langle x, \eta\rangle / h), \eta \in \mathbb{R}^{n}$, are dense in $\mathcal{S}^{\prime}\left(\mathbb{R}_{x}^{n}\right)$.
2. We now claim that $a_{\gamma}(\gamma(x), \eta)=$

$$
\begin{align*}
& \frac{1}{(2 \pi h)^{n}} \iint a(x, \xi) \chi_{x}(y) \chi_{\eta}(\xi) e^{i(\langle x-y, \xi\rangle+\langle\gamma(y)-\gamma(x), \eta\rangle) / h} d y d \xi  \tag{9.81}\\
& \quad+O_{S}\left(\langle\eta\rangle^{-\infty} h^{\infty}\right)
\end{align*}
$$

where

$$
\begin{gathered}
\chi_{x}(y):=\chi(x-y), \quad \chi_{\eta}(\xi):=\chi\left(\left(\xi-\partial \gamma(x)^{T} \eta\right) /\langle\eta\rangle\right), \\
\chi \in C_{\mathrm{c}}^{\infty}(\mathbb{R},[0,1]),\left.\quad \chi\right|_{[-1,1]}=1,\left.\quad \chi\right|_{\mathbf{c}[-2,2]}=0 .
\end{gathered}
$$

In fact, on the support of $1-\chi_{x}(y) \chi_{\eta}(\xi)$ the phase is not stationary:

$$
d_{y, \xi}(\langle x-y, \xi\rangle+\langle\kappa(y)-\kappa(x), \eta\rangle)=0 \Longleftrightarrow\left\{\begin{array}{l}
x=y \\
\text { and } \\
\xi=\partial \kappa(x)^{T} \eta
\end{array}\right.
$$

Consequently the now standard integration-by-parts argument, which we leave to the reader, gives (9.81).
3. We rewrite the main part in (9.81) as follows:

$$
\begin{gathered}
\frac{1}{(2 \pi \tilde{h})^{n}} \iint a_{\eta}(x, \tilde{\xi}) \chi_{x}(y) \tilde{\chi}_{\eta}(\tilde{\xi}) e^{i(\langle x-y, \tilde{\xi})+\langle\gamma(y)-\gamma(x), \eta /\langle\eta\rangle\rangle) / \bar{h}} d y d \tilde{\xi} \\
a_{\eta}(x, \tilde{\xi}):=a(x,\langle\eta\rangle \tilde{\xi}), \quad \tilde{\chi}_{\eta}(\tilde{\xi}):=\chi_{\eta}(\tilde{\xi}\langle\eta\rangle), \quad \tilde{h}:=h /\langle\eta\rangle .
\end{gathered}
$$

The support of the integrand is contained in a fixed compact set, and

$$
h^{k}\langle\eta\rangle^{-m} a_{\eta} \chi_{\eta} \in S(1)
$$

uniformly in $h$ and $\eta$. Hence we can apply the method of stationary phase (Theorem 3.15) to obtain an expansion in powers of $\tilde{h}=h /\langle\eta\rangle$. By computing the leading term we easily check that (9.80) holds.
4. To obtain a formula for terms in the full expansion (9.79) we use the method which already appeared in the second proof of Theorem 3.10 .

Since we will not use (9.79) the argument is only sketched with full details given in [H2, Theorem 18.1.17]: the integrand in (9.81) can be rewritten as follows

$$
\frac{1}{(2 \pi h)^{n}}\left[a(x, \xi) e^{i\left\langle\rho_{x}(y), \eta\right\rangle / h} \chi_{x}(y) \chi_{\eta}(\xi)\right] e^{i\left\langle x-y, \xi-\partial \gamma(x)^{T} \eta\right\rangle / h},
$$

where $\rho_{x}(y)$ is given by (9.79). If we consider the term in square brackets as the amplitude, and change variables to

$$
z:=x-y, \quad w:=\xi-\partial \gamma(x)^{T} \eta,
$$

this becomes the integral in the statement of Theorem ?? (with $z$ and $w$ playing the roles of $x$ and $y$ there). Since $\rho_{x}(y)$ vanishes to second order,

$$
\rho_{x}(x)=0, \quad d_{y} \rho_{x}(x)=0,
$$

at the critical point $x=y$, the differentiation of the oscillatory term in the amplitude combined with the decay of $\partial_{\xi}^{\alpha} a$ shows that the terms in (9.79) are in $S^{m-|\alpha| / 2, k-|\alpha| / 2}$. Hence the formal expansion makes sense but to control the error terms we need arguments similar to those in second proof of Theorem 3.10.

We can now give the generalization of Theorem 9.12. The proof given in $\S 9.6$ can be adapted to the present setting using the integration by parts arguments from the proof of Lemmas 9.14.

THEOREM 9.16 (Changing variables II). Let $a \in S^{m, k}\left(\mathbb{R}^{2 n}\right)$ and let $A$ be its quantization acting on half-densities.
(i) Consider $A$ acting on half-densities. Then

$$
\begin{equation*}
\left(\gamma^{-1}\right)^{*} A \gamma^{*}=\operatorname{Op}(\tilde{a}) \tag{9.82}
\end{equation*}
$$

for

$$
\begin{equation*}
\tilde{a}(x, \xi):=a\left(\gamma^{-1}(x), \partial \gamma(x)^{T} \xi\right)+O_{S^{m-2, k}}\left(h^{2}\right) \tag{9.83}
\end{equation*}
$$

That is,

$$
\begin{equation*}
a(x, \xi)=\tilde{a}\left(\gamma(x),\left(\partial \gamma(x)^{T}\right)^{-1} \xi\right)+O_{S^{m-2, k}}\left(h^{2}\right) \tag{9.84}
\end{equation*}
$$

(ii) When we consider $A$ acting on functions and define

$$
A_{1}=\left(\gamma^{-1}\right)^{*} A \gamma^{*}
$$

then

$$
A_{1}=\mathrm{Op}\left(a_{1}\right)
$$

for

$$
\begin{equation*}
a_{1}(x, \xi):=a\left(\gamma^{-1}(x), \partial \gamma(x)^{T} \xi\right)+O_{S^{m-1, k}}(h) \tag{9.85}
\end{equation*}
$$

That is,
(9.86) $\quad a(x, \xi)=a_{1}\left(\gamma(x),\left(\partial \gamma(x)^{T}\right)^{-1} \xi\right)+O_{S^{m-1, k}}(h)$.

## 10. Quantizing symplectic transformations

10.1 Deformation and quantization
10.2 Semiclassical analysis of propagators
10.3 Semiclassical Strichartz estimates, $L^{p}$ bounds
10.4 More symplectic geometry
10.5 Normal forms for operators with real symbols
10.6 Normal forms for operators with complex symbols
10.7 Application: semiclassical pseudospectra

This final chapter presents some more advanced topics, mostly concerning how (and why) to quantize symplectic transformations.

### 10.1 DEFORMATION AND QUANTIZATION

Throughout this chapter, we identify $\mathbb{R}^{2 n}=\mathbb{R}^{n} \times \mathbb{R}^{n}$. In this section $\gamma: \mathbb{R}^{2 n} \rightarrow \mathbb{R}^{2 n}$ denotes a symplectomorphism:

$$
\gamma^{*} \sigma=\sigma \quad \text { for } \quad \sigma=\sum_{j=1}^{n} d \xi_{j} \wedge d x_{j}
$$

normalized so that $\gamma(0,0)=(0,0)$. Our goal is to quantize $\gamma$ locally, meaning to find a unitary operator $F: L^{2}\left(\mathbb{R}^{n}\right) \rightarrow L^{2}\left(\mathbb{R}^{n}\right)$ such that

$$
F^{-1} A F=B \quad \text { near }(0,0)
$$

for $A=\operatorname{Op}(a)$, where $a \in S$ and $B=\operatorname{Op}(b)$ for

$$
b=\gamma^{*} a+O(h)
$$

This can be useful in practice, since sometimes we can design $\kappa$ so that $\kappa^{*} a$ is more tractable than $a$.

The basic strategy will be (i) finding a family $\left\{\kappa_{t}\right\}_{0 \leq t \leq 1}$ of symplectomorphisms so that $\gamma_{0}=I$ and $\gamma_{1}=\gamma$; (ii) quantizing the functions $q_{t}$ generating this flow of mappings; and then (iii) solving an associated operator ODE (10.7).
10.1.1 Deformations. We begin by deforming $\gamma$ to the identity mapping. So assume $U_{0}$ and $U_{1}$ are simply connected neighborhoods of $(0,0)$ and $\gamma: U_{0} \rightarrow U_{1}$ is a symplectomorphism such that $\kappa(0,0)=$ $(0,0)$.

THEOREM 10.1 (Deforming symplectomorphisms). There exists a continuous, piecewise smooth family

$$
\left\{\gamma_{t}\right\}_{0 \leq t \leq 1}
$$

of local symplectomorphisms $\gamma_{t}: U_{0} \rightarrow U_{t}=: \gamma_{t}\left(U_{0}\right)$ such that
(i) $\gamma_{t}(0,0)=0 \quad(0 \leq t \leq 1)$
(ii) $\gamma_{1}=\gamma, \gamma_{0}=I$.
(iii) Also,

$$
\begin{equation*}
\frac{d}{d t} \gamma_{t}=\left(\gamma_{t}\right)_{*} H_{q_{t}} \quad(0 \leq t \leq 1) \tag{10.1}
\end{equation*}
$$

for a smooth family of functions $\left\{q_{t}\right\}_{0 \leq t \leq 1}$.

REMARK. The statement (10.1) means that for each function $a \in$ $C^{\infty}\left(U_{1}\right)$, we have

$$
\begin{equation*}
\frac{d}{d t} \gamma_{t}^{*} a=H_{q_{t}} \gamma_{t}^{*} a . \tag{10.2}
\end{equation*}
$$

In fact,

$$
\frac{d}{d t} \gamma_{t}^{*} a=\left\langle d a, d \gamma_{t} / d t\right\rangle=\left\langle d a,\left(\gamma_{t}\right)_{*} H_{q_{t}}\right\rangle=H_{q_{t}} \gamma_{t}^{*} a
$$

where $\langle\cdot, \cdot\rangle$ is the pairing of differential 1-forms and vectorfields on $U_{t}$.

Proof. 1. We first consider the case that $\gamma$ is given by a linear symplectomorphism $K: \mathbb{R}^{2 n} \rightarrow \mathbb{R}^{2 n}$ :

$$
\begin{equation*}
K^{*} J K=J \tag{10.3}
\end{equation*}
$$

for

$$
J:=\left(\begin{array}{cc}
0 & I \\
-I & 0
\end{array}\right) .
$$

Since $K$ is an invertible matrix, we have the unique polar decomposition

$$
K=Q P,
$$

where $Q$ is orthogonal and $P$ is positive definite. From (10.3) we deduce that

$$
Q^{*-1} P^{*-1}=K^{*-1}=J Q J^{-1} J P J^{-1}
$$

whence the uniqueness of $Q$ and $P$ implies

$$
Q^{*-1}=J Q J^{-1}, \quad P^{*-1}=J P J^{-1}
$$

That is, both $Q$ and $P$ are symplectic. Furthermore, we can write

$$
P=\exp A
$$

where $A=A^{*}$ and $J A+A J=0$.
2. We identify $\mathbb{R}^{2 n}=\mathbb{R}^{n} \times \mathbb{R}^{n}$ with $\mathbb{C}^{n}$, under the relation $(x, y) \leftrightarrow$ $x+i y$. Since

$$
\left\langle x+i y, x^{\prime}+i y^{\prime}\right\rangle_{\mathbb{C}^{n}}=\left\langle(x, y),\left(x^{\prime}, y^{\prime}\right)\right\rangle_{\mathbb{R}^{n}}+i \sigma\left((x, y),\left(x^{\prime}, y^{\prime}\right)\right),
$$

the fact that $Q$ is orthogonal and symplectic implies it is unitary:

$$
Q=Q^{*-1}=-J Q J
$$

(Similarly, any unitary transformation on $\mathbb{C}^{n}$ gives an orthogonal symplectic transformation in $\mathbb{R}^{n} \times \mathbb{R}^{n}$.)

We can now write

$$
Q=\exp i B
$$

where $B^{*}=B$ is Hermitian on $\mathbb{C}^{n}$. A smooth deformation to the identity is now clear:

$$
K_{t}:=\exp (i t B) \exp (t A) \quad(0 \leq t \leq 1)
$$

3. For the general case that $\gamma$ is nonlinear, set $K:=\partial \gamma(0,0)$. Then for $1 / 2 \leq t \leq 1$,

$$
\gamma_{t}:=K_{2-2 t}^{-1} \circ \gamma
$$

is a piecewise smooth family of symplectomorphisms satisfying

$$
\gamma_{1}=\gamma, \quad \partial \gamma_{1 / 2}(0,0)=I
$$

For $0 \leq t \leq 1 / 2$, we set

$$
\gamma_{t}(m):=\frac{1}{2 t} \gamma_{1 / 2+t}(2 t m)
$$

4. Define $V_{t}:=\frac{d}{d t} \gamma_{t}$; we must show

$$
V_{t}=\left(\kappa_{t}\right)_{*} H_{q_{t}}
$$

for some function $q_{t}$. According to Cartan's formula (Theorem B.3):

$$
\left.\left.\mathcal{L}_{V_{t}} \sigma=d \sigma\right\lrcorner V_{t}+d(\sigma\lrcorner V_{t}\right)
$$

But $\mathcal{L}_{V_{t}} \sigma=\frac{d}{d t} \gamma_{t}^{*} \sigma=\frac{d}{d t} \sigma=0$, since $\gamma_{t}^{*} \sigma=\sigma$. Furthermore, $d \sigma=0$, and consequently $\left.d(\sigma\lrcorner V_{t}\right)=0$. Owing to Poincaré's Lemma (Theorem B.4), we have

$$
\left.\gamma_{t}^{*}(\sigma\lrcorner V_{t}\right)=d q_{t}
$$

for a function $q_{t}$; and this means that $V_{t}=\left(\kappa_{t}\right)_{*} H_{q_{t}}$.
To define our symbol classes, we hereafter consider the order function

$$
m:=\left(1+|x|^{2}+|\xi|^{2}\right)^{\frac{k}{2}}
$$

for some positive integer $k$.

THEOREM 10.2 (Quantizing one parameter families of symplectomorphisms). Let $\left\{\gamma_{t}\right\}_{0 \leq t \leq 1}$ be a smooth family of symplectomorphisms of $\mathbb{R}^{2 n}$, such that

$$
\gamma_{0}=I, \quad \frac{d}{d t} \gamma_{t}=\left(\gamma_{t}\right)_{*} H_{q_{t}}
$$

where $q_{t} \in S(m)$ is a smooth family of real valued symbols.
Then there exists a family of unitary operators

$$
F(t): L^{2}\left(\mathbb{R}^{n}\right) \rightarrow L^{2}\left(\mathbb{R}^{n}\right)
$$

such that

$$
F(0)=I,
$$

and for all $A=\operatorname{Op}(a)$ with $a \in \mathcal{S}$, we have

$$
\begin{equation*}
F(t)^{-1} \circ A \circ F(t)=B(t) \quad(0 \leq t \leq 1) \tag{10.4}
\end{equation*}
$$

for

$$
\begin{equation*}
B(t)=\mathrm{Op}\left(b_{t}\right), \tag{10.5}
\end{equation*}
$$

where

$$
\begin{equation*}
b_{t}=\gamma_{t}^{*} a+h c_{t} \tag{10.6}
\end{equation*}
$$

for $c_{t} \in \mathcal{S} \cap S$.

Proof. 1. We define

$$
Q(t):=\operatorname{Op}\left(q_{t}\right): \mathcal{S} \rightarrow \mathcal{S} \quad(0 \leq t \leq 1)
$$

and recall that

$$
Q(t)^{*}=Q(t)
$$

Since $Q(t)$ depends smoothly on $t$ as an operator on $\mathcal{S}$, we can solve the operator ODE

$$
\left\{\begin{align*}
h D_{t} F(t)+F(t) Q(t) & =0 \quad(0 \leq t \leq 1)  \tag{10.7}\\
F(0) & =I
\end{align*}\right.
$$

for $F(t): \mathcal{S} \rightarrow \mathcal{S}$. Then

$$
\left\{\begin{align*}
h D_{t} F(t)^{*}-Q(t) F(t)^{*} & =0 \quad(0 \leq t \leq 1)  \tag{10.8}\\
F(0)^{*} & =I
\end{align*}\right.
$$

2. We claim that

$$
F(t) \text { is unitary on } L^{2}\left(\mathbb{R}^{n}\right) .
$$

To confirm this, let us calculate using (10.7) and (10.8):

$$
\begin{aligned}
h D_{t}\left(F(t) F(t)^{*}\right) & =h D_{t} F(t) F(t)^{*}+F(t) h D_{t} F(t)^{*} \\
& =-F(t) Q(t) F(t)^{*}+F(t) Q(t) F(t)^{*}=0 .
\end{aligned}
$$

Hence $F(t) F(t)^{*} \equiv I$. On the other hand,

$$
\begin{aligned}
h D_{t}\left(F(t)^{*} F(t)-I\right) & =Q(t) F(t)^{*} F(t)-F(t)^{*} F(t) Q(t) \\
& =\left[Q(t), F(t)^{*} F(t)-I\right] .
\end{aligned}
$$

with $F(0)^{*} F(0)-I=0$. Since this equation for $F(t)^{*} F(t)-I$ is homogeneous, it follows that $F(t)^{*} F(t) \equiv I$.
3. Now define

$$
\begin{equation*}
B(t):=F(t)^{-1} A F(t), \tag{10.9}
\end{equation*}
$$

and first show that

$$
\begin{equation*}
B(t): \mathcal{S}^{\prime} \longrightarrow \mathcal{S} \tag{10.10}
\end{equation*}
$$

This from a stronger statement showing that for any $N$,

$$
\begin{equation*}
B_{N}(t):=\langle x\rangle^{N}\left\langle h D_{x}\right\rangle^{N} B(t)\langle x\rangle^{N}\left\langle h D_{x}\right\rangle^{N}=O(1): L^{2} \longrightarrow L^{2}, \tag{10.11}
\end{equation*}
$$

that is the bound does not depend on $h$. To see this we note that

$$
B_{N}(t)=F_{N}(t)^{-1} B_{N}(0) F_{N}(t),
$$

where $F_{N}(t)$ is defined in the same way as $F(t)$ but with $Q(t)$ replaced by

$$
Q_{N}(t):=\left\langle h D_{x}\right\rangle^{-N}\langle x\rangle^{-N} Q(t)\langle x\rangle^{N}\left\langle h D_{x}\right\rangle .
$$

Theorem 4.13 shows that $Q_{N}(t)=\operatorname{Op}\left(q_{N}(t)\right), q_{N}(t) \in S(1)$, and hence it is bounded on $L^{2}$. The inverse, $F_{N}(t)^{-1}=G_{N}(t): \mathcal{S} \rightarrow \mathcal{S}$, is obtained by solving

$$
h D_{t} G_{N}(t)-Q_{N}(t) G_{N}(t)=0 \quad(0 \leq t \leq 1), \quad G_{N}(0)=I
$$

Since $Q_{N}(t)=Q_{N}(t)^{*}+O_{L^{2} \rightarrow L^{2}}(h)$, we see that

$$
\frac{d}{d t}\left\|G_{N}(t) u\right\|^{2}=\frac{2}{h} \operatorname{Im}\left\langle Q_{N}(t) G_{N}(t) u, G_{N}(t)\right\rangle \leq C\left\|G_{N}(t) u\right\|^{2}
$$

and hence, by Gronwall's inequality, $G_{N}(t)$, and hence $F_{N}(t)$, are bounded on $L^{2}$, uniformly with respect to $h$.

This concludes the proof of (10.11) since

$$
B_{N}(0)=\langle x\rangle^{N}\left\langle h D_{x}\right\rangle^{N} A\langle x\rangle^{N}\left\langle h D_{x}\right\rangle^{N}=O(1): L^{2} \longrightarrow L^{2}
$$

by the assumption that $A=\operatorname{Op}(a), a \in \mathcal{S}$.
4. We assert that

$$
\begin{equation*}
B(t)=\mathrm{Op}\left(b_{t}\right) \tag{10.12}
\end{equation*}
$$

for

$$
\begin{equation*}
b_{t}=\gamma_{t}^{*} a+O(h), \quad b_{t} \in \mathcal{S} \cap S(1) . \tag{10.13}
\end{equation*}
$$

To prove this, define the family of pseudodifferential operators

$$
\tilde{B}(t):=\mathrm{Op}\left(\gamma_{t}^{*} a\right) .
$$

We calculate

$$
\begin{aligned}
h D_{t} \tilde{B}(t) & =\frac{h}{i} \mathrm{Op}\left(\frac{d}{d t} \gamma_{t}^{*} a\right)=\frac{h}{i} \mathrm{Op}\left(H_{q_{t}} \gamma_{t}^{*} a\right) \\
& =\frac{h}{i} \mathrm{Op}\left(\left\{q_{t}, \gamma_{t}^{*} a\right\}\right)=[Q(t), \tilde{B}(t)]+E(t),
\end{aligned}
$$

and the pseudodifferential calculus implies that

$$
\|E(t)\|_{L^{2} \rightarrow L^{2}}=O\left(h^{2}\right)
$$

where $E(t)=\operatorname{Op}(e(t))$ for a symbol $e(t) \in S^{-2}$.
Therefore

$$
\begin{aligned}
h D_{t}\left(F(t) \tilde{B}(t) F(t)^{-1}\right)= & \left(h D_{t} F(t)\right) \tilde{B}(t) F(t)^{-1}+F(t)\left(h D_{t} \tilde{B}(t)\right) F(t)^{-1} \\
& +F(t) \tilde{B}(t) h D_{t}\left(F(t)^{-1}\right) \\
= & -F(t) Q(t) \tilde{B}(t) F(t)^{-1}+F(t)([Q(t), \tilde{B}(t)] \\
& +E(t)) F(t)^{-1}+F(t) \tilde{B}(t) Q(t) F(t)^{-1} \\
= & F(t) E(t) F(t)^{-1}=O\left(h^{2}\right) .
\end{aligned}
$$

Integrating and dividing by $h$ gives

$$
\begin{equation*}
F(t) \tilde{B}(t) F(t)^{-1}=A+\frac{i}{h} \int_{0}^{t} F(s) E(s) F(s)^{-1} d s=A+O(h) \tag{10.14}
\end{equation*}
$$

so that $\tilde{B}(t)-B(t)=O(h)$.
5. We will now construct families of pseudodifferential operators $B_{k}(t)$ so that for each $m$

$$
\begin{equation*}
B(t)=\tilde{B}(t)+B_{1}(t)+\cdots+B_{m}(t)+O_{L^{2} \rightarrow L^{2}}\left(h^{m+1}\right), \quad B_{j} \in \Psi^{-j} \tag{10.15}
\end{equation*}
$$

For that let

$$
\tilde{e}(t)=\left(\kappa_{t}\right)^{*} \int_{0}^{t}\left(\kappa_{s}^{-1}\right)^{*} e(s) d s
$$

and set $\tilde{E}(t)=\operatorname{Op}(\tilde{e}(t))$. We observe that

$$
h D_{t} \tilde{E}(t)=[Q(t), \tilde{E}]+\frac{h}{i}\left(E(t)+E_{1}(t)\right),
$$

where $E_{1}(t)=\mathrm{Op}\left(e_{1}(t)\right), e_{1}(t) \in S^{-3}$ by the pseudodifferential calculus. Then, as in Step 4 above,

$$
\begin{aligned}
h D_{t}\left(F(t) \tilde{E}(t) F(t)^{-1}\right)= & -F(t)[Q(t), \tilde{E}(t)] F(t)^{-1} \\
& \quad+F(t) h D_{t}(\tilde{E}(t)) F(t)^{-1} \\
& =\frac{h}{i}\left(F(t) E(t) F(t)^{-1}+F(t) E_{1}(t) F(t)^{-1}\right) .
\end{aligned}
$$

Integrating in $t$ gives

$$
F(t) \tilde{E}(t) F(t)^{-1}=\int_{0}^{t} F(s) E(s) F(s)^{-1} d s+\int_{0}^{t} F(s) E_{1}(s) F(s)^{-1} d s
$$

This we now substitute in (10.14) obtaining

$$
\begin{aligned}
\tilde{B}(t)-B(t) & =\frac{i}{h} \tilde{E}(t)-F(t)^{-1}\left(\frac{i}{h} \int_{0}^{t} F(s) E_{1}(s) F(s)^{-1} d s\right) F(t) \\
& =\frac{i}{h} \tilde{E}(t)+O_{L^{2} \rightarrow L^{2}}\left(h^{2}\right)
\end{aligned}
$$

Setting $B_{1}(t)=i \tilde{E}(t) / h \in \Psi^{-1}$, and continuing inductively gives $B_{k}(t)$ satisfying (10.15).
6. It remains to show that $B(t)$ is a pseudodifferential operator. To do so, we invoke Beals's Theorem 9.8 by showing that for any linear $l_{1}, \cdots, l_{M}$, we have the estimate

$$
\begin{equation*}
\operatorname{ad}_{l_{M}} \cdots \operatorname{ad}_{l_{1}} B(t)=O_{L^{2} \rightarrow L^{2}}\left(h^{M}\right) . \tag{10.16}
\end{equation*}
$$

But this statement is clear from Steps 3 and 5 : for any $P$ we can find a pseudodifferential operator $\operatorname{Op}\left(b_{t}^{P}\right)$, with $b_{t}^{N} \in S^{-1}$, such that

$$
\begin{gathered}
B(t)=\mathrm{Op}\left(b_{t}^{P}\right)+R_{P}(t) \\
\langle x\rangle^{N}\left\langle h D_{x}\right\rangle^{N} R_{P}(t)\langle x\rangle^{N}\left\langle h D_{x}\right\rangle^{N}=O\left(h^{P}\right): L^{2} \rightarrow L^{2} .
\end{gathered}
$$

Since

$$
\operatorname{ad}_{l_{M}} \cdots \operatorname{ad}_{l_{1}} \operatorname{Op}\left(b_{t}^{N}\right)=O\left(h^{M}\right)
$$

and, by a trivial estimate,

$$
\operatorname{ad}_{l_{M}} \cdots \operatorname{ad}_{l_{1}} R_{P}(t)=O\left(h^{P}\right)
$$

(10.16) follows by choosing $M \geq P$.

REMARK. The argument used in Step 2 of the proof shows that if in Theorem 10.2 we have

$$
a(x, \xi ; h) \sim a_{0}(x, \xi)+h a_{1}(x, \xi)+\cdots+h^{N} a_{N}(x, \xi)+\cdots,
$$

for $a_{j} \in S$, then

$$
b_{t}(x, \xi ; h) \sim \gamma_{t}^{*} a_{0}(x, \xi)+h b_{t}^{1}(x, \xi)+\cdots h^{N} b_{t}^{N}(x, \xi)+\cdots
$$

However, the higher order terms are difficult to compute.
10.1.2 Locally defined symplectomorphisms. The requirement that the family of symplectomorphism be global on $\mathbb{R}^{2 n}$ is very strong and often invalid in interesting situations. So we now discuss quantization of locally defined symplectomorphisms, for which the quantization formula (10.4) holds only locally.

THEOREM 10.3 (Local quantization). Let $\gamma: U_{0} \rightarrow U_{1}$ be a symplectomorphism fixing $(0,0)$ and defined in a neighbourhood of $U_{0}$.

Then there exists a unitary operator

$$
F: L^{2}\left(\mathbb{R}^{n}\right) \rightarrow L^{2}\left(\mathbb{R}^{n}\right)
$$

such that for all $A=\operatorname{Op}(a)$ with $a \in S$, we have

$$
\begin{equation*}
F^{-1} A F=B \tag{10.17}
\end{equation*}
$$

where $B=\operatorname{Op}(b)$ for a symbol $b \in S$ satisfying

$$
\begin{equation*}
\left.b\right|_{U_{0}}:=\left.\gamma^{*}\left(\left.a\right|_{U_{1}}\right)\right|_{U_{0}}+O(h) . \tag{10.18}
\end{equation*}
$$

Proof. 1. According to Theorem 10.1, there exists a piecewise smooth family of symplectomorphisms $\gamma_{t}: U_{0} \rightarrow U_{t},(0 \leq t \leq 1)$ such that $\gamma=\gamma_{1}, \gamma_{0}=I$, and

$$
\frac{d}{d t} \gamma_{t}=\left(\gamma_{t}\right)_{*} H_{q_{t}} \quad(0 \leq t \leq 1)
$$

within $U$, for a smooth family $\left\{q_{t}\right\}_{0 \leq t \leq 1}$.
We extend $q_{t}$ smoothly to be equal to 0 in $\mathbb{R}^{2 n}-U_{0}$ and then define a family of global symplectomorphisms $\tilde{\gamma}_{t}$ using the now globally defined functions $q_{t}$. Observe that

$$
\left.\tilde{\gamma}_{t}\right|_{U_{0}}=\gamma_{t}: U_{0} \rightarrow U_{t}
$$

and hence

$$
\begin{equation*}
\left.\tilde{\gamma}_{t}^{*}(a)\right|_{U_{0}}=\left.\gamma_{t}^{*}\left(\left.a\right|_{U_{t}}\right)\right|_{U_{0}} . \tag{10.19}
\end{equation*}
$$

2. We now apply Theorem 10.2 , to obtain the family of operators $\{F(t)\}_{0 \leq t \leq 1}$. We observe that since the supports of the functions $q_{t}$ lie in a fixed compact set, the proof of Theorem 10.2 shows that (10.4) holds for $a \in S$. That is,

$$
F(t)^{-1} A F(t)=\operatorname{Op}(b(t))=B(t)
$$

for

$$
b(t)=\tilde{\gamma}_{t}^{*} a+O(h)
$$

We now put

$$
F:=F(1), \quad B:=B(1) .
$$

Then (10.19) shows that formula (10.17) is valid.
10.1.3 Microlocality. It will prove useful to formulate the theorems above without reference to the global properties of the operator $F$.

DEFINITIONS. (i) Let $U, V$ be open, bounded subsets of $\mathbb{R}^{2 n}$, and assume

$$
T: \mathcal{S}\left(\mathbb{R}^{n}\right) \rightarrow \mathcal{S}\left(\mathbb{R}^{n}\right)
$$

is linear.
We say that $T$ is tempered if for each seminorm $\|\cdot\|_{1}$ on $\mathcal{S}\left(\mathbb{R}^{n}\right)$, there exists another seminorm $\|\cdot\|_{2}$ and a constant $N \in \mathbb{R}$ such that

$$
\begin{equation*}
\|T u\|_{1}=O\left(h^{-N}\right)\|u\|_{2} \tag{10.20}
\end{equation*}
$$

for all $u \in \mathcal{S}$.
(ii) Given two tempered operators $T$ and $S$, we say that

$$
\begin{equation*}
T \equiv S \quad \text { microlocally on } \quad U \times V \tag{10.21}
\end{equation*}
$$

if there exist open sets $\tilde{U} \supseteq U$ and $\tilde{V} \supseteq V$ such that

$$
A(T-S) B=O\left(h^{\infty}\right)
$$

as a mapping $\mathcal{S} \rightarrow \mathcal{S}$, for all $A, B$ such that

$$
\mathrm{WF}_{h}(A) \subset \tilde{V}, \mathrm{WF}_{h}(B) \subset \tilde{U}
$$

(iii) In particular, we say

$$
T \equiv I \text { microlocally near } U \times U
$$

if there exists an open set $\tilde{U} \supseteq U$ such that

$$
A-T A=A-A T=O\left(h^{\infty}\right)
$$

as mappings $\mathcal{S} \rightarrow \mathcal{S}$, for all $A$ with $\mathrm{WF}_{h}(A) \subset \tilde{U}$.
(iv) We will say that $T$ is microlocally invertible near $U \times U$ if there exists an operator $S$ such that $T S \equiv I$ and $S T \equiv I$ microlocally near $U \times U$.

When no confusion is likely, we write

$$
S=T^{-1}
$$

and call $S$ a microlocal inverse of $T$.

LEMMA 10.4 (Wavefront sets and composition). If

$$
\mathrm{WF}_{h}(A) \cap U=\emptyset
$$

and $B=\mathrm{Op}(b)$ for $b \in S$, then

$$
\begin{equation*}
\mathrm{WF}_{h}(B A) \cap U=\emptyset \tag{10.22}
\end{equation*}
$$

Proof. The symbol of $B A$ is $b \# a=O\left(h^{\infty}\right)$ in $U$.
LEMMA 10.5 (Tempered unitary transformations). The unitary transformations $F(t)$ given by Theorem 10.2 are tempered.

Proof. Up to powers of $h$, each seminorm on $\mathcal{S}$ is bounded from above and below by these specific seminorms:

$$
u \mapsto\left\|A_{N} u\right\| \quad \text { for } \quad A_{N}:=\left(1+|x|^{2}+|h D|^{2}\right)^{N}
$$

We observe that the operators $A_{N}$ are invertible and selfadjoint and that, in the notation of the proof of Theorem 10.2

$$
A_{N} Q(t) A_{N}^{-1}=Q_{N}(t)=\operatorname{Op}\left(q_{t}^{N}\right)
$$

for $q_{t}^{N} \in S(m)$ such that $q_{t}^{N}-q_{t} \in S^{-1}(m)$.
We then have

$$
h D_{t} A_{N} F(t) A_{N}^{-1}=A_{N} F(t) A_{N}^{-1} Q_{N}(t)
$$

and hence the same arguments as in Step 3 of the proof of Theorem 10.4,

$$
\left\|A_{N} F(t) u\right\|^{2} \leq C\left\|A_{N} u\right\|^{2}
$$

Consequently for any seminorm $\|\cdot\|_{1}$ on $\mathcal{S}$, there exists a seminorm $\|\cdot\|_{2}$ and $N$ such that

$$
\|F(t) u\|_{1} \leq O\left(h^{-N}\right)\|u\|_{2} .
$$

The previous two lemmas and Theorem 10.3 give
THEOREM 10.6 (More on local quantization). Let $\gamma: U_{0} \rightarrow U_{1}$ be a symplectomorphism fixing $(0,0)$ and defined in a neighbourhood of $U_{0}$. Suppose $U$ is open, $\bar{U} \subset \subset U_{0} \cap U_{1}$.

Then there exists a tempered operator

$$
F: L^{2}\left(\mathbb{R}^{n}\right) \rightarrow L^{2}\left(\mathbb{R}^{n}\right)
$$

such that $F$ is microlocally invertible near $U \times U$ and for all $A=\operatorname{Op}(a)$, with $a \in S$,

$$
\begin{equation*}
F^{-1} A F=B \quad \text { microlocally near } U \times U \tag{10.23}
\end{equation*}
$$

where $B=\mathrm{Op}(b)$ for a symbol $b \in S$ satisfying

$$
\begin{equation*}
b:=\gamma^{*} a+O(h) . \tag{10.24}
\end{equation*}
$$

In (10.24) we do not specify the neighbourhoods, as we did in (10.19), since the statement needs to make sense only locally near $U \times U$.

The last theorem has the following converse which we include for completeness:

THEOREM 10.7 (Converse). Suppose that $F: L^{2}\left(\mathbb{R}^{n}\right) \rightarrow L^{2}\left(\mathbb{R}^{n}\right)$ is a tempered operator such that for every $A=\operatorname{Op}(a)$ with $a \in S$, we have

$$
A F \equiv F B
$$

microlocally near ( 0,0 ), for

$$
B=\mathrm{Op}(b), \quad b=\gamma^{*} a+O(h),
$$

where $\kappa: \mathbb{R}^{2 n} \rightarrow \mathbb{R}^{2 n}$ is a symplectomorphism, defined locally near $U$, with $\kappa(0,0)=(0,0)$.

Then there exists a pseudodifferential operator $F_{0}$, elliptic near $U$, and a family of self-adjoint pseudodifferential operators $Q(t)$, such that

$$
F=F(1) \quad \text { microlocally near } U \times U,
$$

where

$$
\left\{\begin{aligned}
h D_{t} F(t)+F(t) Q(t) & =0 \quad(0 \leq t \leq 1) \\
F(0) & =F_{0} .
\end{aligned}\right.
$$

Proof. 1. From Theorem 10.1 we know that there exists a family of local symplectomorphisms, $\gamma_{t}$, satisfying $\gamma_{t}(0,0)=(0,0), \gamma_{1}=\gamma$ and $\gamma_{0}=I$. Since we are working locally, there exists a function $q_{t}$ so that $\gamma_{t}$ is generated by its Hamiltonian vectorfield $H_{q_{t}}$.

As in the proof of Theorem 10.3 we extend this function to be zero outside a compact set. Let us now consider the dynamics

$$
\left\{\begin{aligned}
h D_{t} F(t) & =Q(t) F(t) \quad(0 \leq t \leq 1) \\
F(1) & =C F C
\end{aligned}\right.
$$

where $C$ is a pseudodifferential operator with $\mathrm{WF}_{h}(I-C) \cap U=\emptyset$.
2. We claim that $F(0)$ satisfies

$$
\begin{equation*}
\operatorname{Op}(a) F(0)=F(0) \operatorname{Op}(a+h \tilde{a}) \tag{10.25}
\end{equation*}
$$

for $a, \tilde{a} \in \mathcal{S} \cap S^{0}(1)$. To establish this, let us introduce $V(t)$ satisfying

$$
\left\{\begin{aligned}
h D_{t} V(t)+V(t) Q(t) & =0 \quad(0 \leq t \leq 1) \\
V(0) & =I .
\end{aligned}\right.
$$

Then using Theorem 10.3 and the assumption that $b=\gamma^{*} a+0(h)$, we deduce that

$$
\begin{aligned}
\mathrm{Op}(a) F(t) V(t) & =F(t) \operatorname{Op}(b) V(t) \\
& =F(t) V(t)\left(V(t)^{-1} \mathrm{Op}(b) V(t)\right) \\
& =F(t) V(t) \operatorname{Op}(a+h \tilde{a})
\end{aligned}
$$

Putting $t=0$ gives (10.25).
3. We now use Beals's Theorem to conclude that $F(0) \in \Psi^{0}$. We verify the hypothesis by induction: suppose we know that

$$
\operatorname{ad}_{\mathrm{Op}\left(b_{1}\right)} \cdots a d_{\mathrm{Op}\left(b_{N}\right)} F(0)=O\left(h^{N}\right)
$$

for any $b_{j} \in S^{0}(1)$. Then by (10.25)

$$
\operatorname{Op}\left(b_{N+1}\right) F(0)-F(0) \operatorname{Op}\left(b_{N+1}\right)=h \operatorname{Op}\left(\tilde{b}_{N+1}\right) F(0) ;
$$

and hence

$$
\begin{aligned}
& \left\|\operatorname{ad}_{\mathrm{Op}\left(b_{1}\right)} \cdots a d_{\mathrm{Op}\left(b_{N}\right)} \operatorname{ad}_{\mathrm{Op}\left(b_{N+1}\right)} F(0)\right\|_{L^{2} \rightarrow L^{2}}= \\
& \quad\left\|h \operatorname{ad}_{\mathrm{Op}\left(b_{1}\right)} \cdots \operatorname{ad}_{\mathrm{Op}\left(b_{N}\right)}\left(\operatorname{Op}\left(\tilde{b}_{N+1}\right) F(0)\right)\right\|_{L^{2} \rightarrow L^{2}}=O\left(h^{N+1}\right),
\end{aligned}
$$

according to the induction hypothesis and the derivation property

$$
\operatorname{ad}_{A}(B C)=B\left(\operatorname{ad}_{A} C\right)+\left(\operatorname{ad}_{A} B\right) C .
$$

Hence Beals's Theorem applies and shows that $F(0)$ is a pseudodifferential operator. By construction, $F(1)=C F C \equiv F$ near $(0,0)$.

### 10.1.4. Quantization of linear symplectic maps.

CHECK DEFINITION OF J Consider first the simple linear symplectic transformation $\gamma=J$; that is,

$$
\begin{equation*}
\gamma(x, \xi)=(-\xi, x) \tag{10.26}
\end{equation*}
$$

on $\mathbb{R}^{2 n}=\mathbb{R}^{n} \times \mathbb{R}^{n}$.
Then we can take for $0 \leq t \leq 1$,

$$
\gamma_{t}(x, \xi)=\left(\cos \left(\frac{t \pi}{2}\right) x-\sin \left(\frac{t \pi}{2}\right) \xi, \sin \left(\frac{t \pi}{2}\right) x+\cos \left(\frac{t \pi}{2}\right) \xi\right)
$$

so that

$$
\frac{d \gamma_{t}}{d t}=\left(\gamma_{t}\right)_{*} H_{q}
$$

for

$$
q:=\frac{\pi}{4}\left(|x|^{2}+|\xi|^{2}\right)
$$

THEOREM 10.8 (J quantized). The operator $F$ associated with the transformation (10.26) as in Theorem 10.3 is

$$
\begin{equation*}
F u(x):=\frac{e^{-\frac{\pi}{4} i}}{(2 \pi h)^{\frac{n}{2}}} \int_{\mathbb{R}^{n}} e^{-\frac{i\langle x, y\rangle}{h}} u(y) d y=\frac{e^{-\frac{\pi}{4} i}}{(2 \pi h)^{\frac{n}{2}}} \mathcal{F}_{h} u . \tag{10.27}
\end{equation*}
$$

Proof. 1. To verify this, we first show that for $a \in \mathcal{S}^{\prime}$ we have

$$
\begin{equation*}
a^{\mathrm{w}}(x, h D) \circ F=F \circ a^{\mathrm{w}}(-h D, x) ; \tag{10.28}
\end{equation*}
$$

that is, the conclusion of Theorem 10.2 holds without any error terms. As in the proof of that theorem, we see that

$$
h D_{t} A_{t}=\frac{\pi}{4}\left[-h^{2} \Delta+|x|^{2}, A_{t}\right]
$$

for

$$
A_{t}:=F(t)^{-1} a^{\mathrm{w}}(x, h D) F(t) .
$$

Let $l(x, \xi)$ be a linear function on $\mathbb{R}^{2 n}$ and consider the exponential symbol

$$
a_{t}(x, \xi):=\exp \left(\gamma_{t}^{*} l(x, \xi) / h\right)
$$

and its Weyl quantization

$$
a_{t}^{\mathrm{w}}(x, h D)=\exp \left(\gamma_{t}^{*} l(x, h D) / h\right) .
$$

An explicit computation reveals that

$$
h D_{t} a_{t}(x, h D)=\frac{\pi}{4}\left[-h^{2} \Delta+|x|^{2}, a_{t}(x, h D)\right] .
$$

Since any Weyl operator is a superposition of exponentials of $l$ 's (recall (??)), assertion (10.28) follows.
2. Suppose now that $\widetilde{F}$ is another unitary operator for which (10.28) holds. Then $\tilde{F}=c F$ for $c \in \mathbb{C},|c|=1$, as follows from applying Lemma 3.3 to $L=F^{*} \widetilde{F}$. Since the Fourier transform satisfies (10.28) and $(2 \pi h)^{-n / 2} \mathcal{F}_{h}$ is unitary, we deduce that

$$
F=\frac{c}{(2 \pi h)^{\frac{n}{2}}} \mathcal{F}_{h} .
$$

3. Thus it remains to compute the constant $c$. For this, let us put $u_{0}=\exp \left(-|x|^{2} / 2\right)$ and consider the ODE

$$
\left\{\begin{aligned}
h D_{t} u(t) & =\frac{\pi}{4}\left(-h^{2} \Delta+|x|^{2}\right) u(t) \\
u(0) & =u_{0} .
\end{aligned}\right.
$$

Recalling (10.7), we see that $u(t)=F(t)^{*} u_{0}$. Since $u_{0}$ is the ground state of the harmonic oscillator with eigenvalue $h$, we learn that $u(t)=$ $a(t) u_{0}$, where $a(t)$ solves the ODE

$$
\left\{\begin{aligned}
\frac{d}{d t} a(t) & =\frac{\pi i}{4} a(t) \\
a(0) & =1 ;
\end{aligned}\right.
$$

that is, $a(t)=\exp (\pi i t / 4)$. Finally, we note that

$$
e^{\pi i / 4} u_{0}=F(1)^{*} u_{0}=\bar{c}(2 \pi h)^{-n / 2} \mathcal{F}_{h} u_{0}=\bar{c} u_{0}
$$

whence

$$
c=\exp (-\pi i / 4)
$$

REMARK. The family of canonical transformations $\gamma_{t}(0 \leq t \leq 1)$, used here can be extended to a periodic family of canonical transformations: $\gamma_{t+4}=\gamma_{t}(t \in \mathbb{R})$. Extending $F(t)$ using (10.7), we see that the argument above gives

$$
F(4 k)=(-1)^{k} I, \quad \gamma_{4 k}=I .
$$

Consequently on the quantum level the deformation produces an additional shift in the phase. This shift has an important geometric and physical interpretation and is related to the Maslov index. For a brief discussion and references see [?, Sect.7].

REMARK: Quantizing linear symplectic mappings. Using Step 1 in the proof of Theorem 10.1, we can in fact quantize any linear symplectic transformation. So given

$$
K: \mathbb{R}^{2 n} \rightarrow \mathbb{R}^{2 n}, \quad K=\left(\begin{array}{cc}
A & B \\
C & D
\end{array}\right)
$$

where

$$
C^{*} A=A^{*} C, \quad D^{*} B=B^{*} D, \quad D^{*} A-B^{*} C=I,
$$

we can construct $F_{K}: L^{2}\left(\mathbb{R}^{n}\right) \rightarrow L^{2}\left(\mathbb{R}^{n}\right)$ satisfying

$$
F_{K}^{*} F_{K}=F_{K} F_{K}^{*}=I, \quad a^{\mathrm{w}}(x, h D) \circ F_{K}=F_{K} \circ\left(K^{*} a\right)^{\mathrm{w}}(x, h D) .
$$

The operator $F_{K}$ is unique up to a multiplicative factor; and hence

$$
F_{K_{1}} \circ F_{K_{2}}=c F_{K_{1} \circ K_{2}}, \quad|c|=1 .
$$

The association $K \mapsto F_{K}$ can in fact be chosen so that $c= \pm 1$; therefore it is almost a representation of the group of symplectic transformations. To make it a representation, one has to move to the double cover of the symplectic group, the so-called the metaplectic group. Unitary operators quantizing linear symplectic transformations are consequently
called metaplectic operators: see Dimassi-Sjöstrand [D-S, Appendix to Chapter 7] for a self-contained presentation in the semiclassical spirit, and Folland [?, Chapter 4] for more and for references.

EXAMPLE: $A$ invertible. For reasons already apparent in the discussion of the Fourier transform, there cannot be a general formula for the kernel $F_{K}$ in terms of the entries $A, B, C, D$ of $K$.

But if $\operatorname{det} A \neq 0$, we have for $u \in \mathcal{S}$ the formula

$$
\begin{equation*}
F_{K} u(x)=\frac{(\operatorname{det} A)^{-\frac{1}{2}}}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}(\varphi(x, \eta)-\langle y, \eta\rangle)} u(y) d y d \xi, \tag{10.29}
\end{equation*}
$$

where

$$
\begin{equation*}
\varphi(x, \eta):=-\frac{1}{2}\left\langle C A^{-1} x, x\right\rangle+\left\langle A^{-1} x, \eta\right\rangle+\frac{1}{2}\left\langle A^{-1} B \eta, \eta\right\rangle . \tag{10.30}
\end{equation*}
$$

We will refer to this formula in our next example.

### 10.2 SEMICLASSICAL ANALYSIS OF PROPAGATORS

In this section we consider the flow of symplectic transformations

$$
\begin{equation*}
\gamma_{t}=\exp \left(t H_{p}\right) \tag{10.31}
\end{equation*}
$$

generated by the real-valued symbol $p \in S(m)$.
Let $P=\operatorname{Op}(p)$. Then in the notation of Theorem 10.2, $F(t)=$ $e^{-i t P / h}$ solves

$$
\left\{\begin{aligned}
\left(h D_{t}+P\right) F(t) u & =0 \\
F(0) u & =u
\end{aligned}\right.
$$

for $u \in \mathcal{S}$. In this case, Theorem 10.2 reproduces Egorov's Theorem 8.2: if $a \in \mathcal{S}$, then

$$
e^{i t P / h} \mathrm{Op}(a) e^{-i t P / h}=\mathrm{Op}\left(b_{t}\right),
$$

for

$$
b_{t}=\left(\exp t H_{p}\right)^{*} a+O(h)
$$

A Fourier integral representation formula. Our goal now is to find for small times $t_{0}>0$ a microlocal representation of $F(t)$ as an oscillatory integral. In other words, we would like to find an operator $U(t)$ so that for each $h$ dependent family, $u \in \mathcal{S}$ with $\mathrm{WF}_{h}(u) \subset \subset \mathbb{R}^{2 n}$, we have

$$
\left\{\begin{align*}
h D_{t} U(t) u+P U(t) u & =O\left(h^{\infty}\right) \quad\left(-t_{0} \leq t \leq t_{0}\right)  \tag{10.32}\\
U(0) u & =u .
\end{align*}\right.
$$

Using Duhamel's formula, we can then deduce that

$$
F(t)-U(t)=O\left(h^{\infty}\right)
$$

THEOREM 10.9 (Oscillatory integral representation). We have the representation

$$
\begin{equation*}
U(t) u(x)=\frac{1}{(2 \pi h)^{n}} \int_{\mathbb{R}^{n}} \int_{\mathbb{R}^{n}} e^{\frac{i}{h}(\varphi(t, x, \eta)-\langle y, \eta\rangle)} b(t, x, \eta ; h) u(y) d y d \eta, \tag{10.33}
\end{equation*}
$$

for the phase $\varphi$ and amplitude $b$ as defined below.

The proof will appear after the following constructions of the phase and amplitude.

Construction of the phase function. We start by finding the phase function $\varphi$ as a local generating function associated with the symplectomorphisms (10.31). (Recall the discussion in $\S 2.3$ of generating functions.)

Let $U$ denote a bounded open set containing $(0,0)$.
LEMMA 10.10 (Hamilton-Jacobi equation). If $t_{0}>0$ is small enough, there exists a smooth function

$$
\varphi=\varphi(t, x, \eta)
$$

defined in $\left(-t_{0}, t_{0}\right) \times U \times U$, such that

$$
\gamma_{t}(y, \eta)=(x, \xi)
$$

locally if and only if

$$
\begin{equation*}
\xi=\partial_{x} \varphi(t, x, \eta), \quad y=\partial_{\eta} \varphi(t, x, \eta) . \tag{10.34}
\end{equation*}
$$

Furthermore, $\varphi$ solves the Hamilton-Jacobi equation

$$
\left\{\begin{array}{l}
\partial_{t} \varphi(t, x, \eta)+p\left(x, \partial_{x} \varphi(t, x, \eta)\right)=0  \tag{10.35}\\
\varphi(0, x, \eta)=\langle x, \eta\rangle
\end{array}\right.
$$

Proof. 1. We know that for points $(y, \eta)$ lying in a compact subset of $\mathbb{R}^{2 n}$, the flow

$$
\begin{equation*}
(y, \eta) \mapsto \gamma_{t}(y, \eta) \tag{10.36}
\end{equation*}
$$

is surjective near $(0,0)$ for times $0 \leq t \leq t_{0}$, provided $t_{0}$ is small enough. This is so since $\gamma_{0}(y, \eta)=(y, \eta)$.
2. To show the existence of $\varphi$, consider

$$
\Lambda:=\left\{\left(t, p(y, \eta) ; \gamma_{t}(y, \eta) ; y, \eta\right): t \in \mathbb{R},(y, \eta) \in \mathbb{R}^{2 n}\right\}
$$

This is a surface in $\mathbb{R}^{2} \times \mathbb{R}^{2 n} \times \mathbb{R}^{2 n}$, a typical point of which we will write as $(t, \tau, x, \xi, y, \eta)$. Introduce the one-form

$$
V:=-\tau d t+\sum_{j=1}^{n} \xi_{j} d x_{j}+\sum_{j=1}^{n} y_{j} d \eta_{j}
$$

That $\gamma_{t}$ is a symplectic implies $\left.d V\right|_{\Lambda}=0$. By Poincaré's Lemma (Theorem B.4), there exists a smooth function $\varphi$ such that

$$
d \varphi=V
$$

In view of (10.36) we can use $(t, x, \eta)$ as coordinates on $\Lambda \cap\left(\left(-t_{0}, t_{0}\right) \times\right.$ $U \times U)$; and hence

$$
-\tau d t+\sum_{j=1}^{n} \xi_{j} d x_{j}+\sum_{j=1}^{n} y_{j} d \eta_{j}=\partial_{t} \varphi d t+\sum_{j=1}^{n} \partial_{x_{j}} \varphi d x_{j}+\sum_{j=1}^{n} \partial_{\eta_{j}} \varphi d \eta_{j} .
$$

Comparing the terms on the two sides gives (10.34) and (10.35).
Construction of the amplitude. The amplitude $b$ in (10.33) must satisfy

$$
\left(h D_{t}+p^{\mathrm{w}}(x, h D)\right)\left(e^{i \varphi(t, x, \eta) / h} b(t, x, \eta ; h)\right)=O\left(h^{\infty}\right)
$$

and so

$$
\begin{equation*}
\left(\partial_{t} \varphi+h D_{t}+e^{-i \varphi / h} p^{\mathrm{w}}(x, h D) e^{i \varphi / h}\right) b(t, x, \eta ; h)=O\left(h^{\infty}\right), \tag{10.37}
\end{equation*}
$$

for $(x, \eta)$ in a neighbourhood of $U \times U, 0 \leq t \leq t_{0}$.
We will build $b$ as an expansion in powers of $h$ :

$$
\begin{equation*}
b(t, x, \eta ; h) \sim b_{0}(t, x, \eta)+h b_{1}(t, x, \eta)+h^{2} b_{2}(t, x, \eta)+\cdots \tag{10.38}
\end{equation*}
$$

Once all the terms $b_{j}$ are computed, Borel's Theorem 4.11 produces the amplitude $b$.

LEMMA 10.11 (Calculation of $b_{0}$ ). We have

$$
\begin{equation*}
b_{0}(t, x, \eta)=\left(\operatorname{det} \partial_{\eta x}^{2} \varphi(t, x, \eta)\right)^{\frac{1}{2}} \tag{10.39}
\end{equation*}
$$

Note that $\operatorname{det} \partial_{\eta x}^{2} \varphi>0$ for $0 \leq t \leq t_{0}$, if $t_{0}$ is sufficiently small.
Proof. 1. We first observe that

$$
e^{-i \varphi / h} p^{\mathrm{w}}(x, h D) e^{i \varphi / h}=q_{t}(x, h D ; h),
$$

where

$$
\begin{equation*}
q_{t}(x, \xi ; h)=p\left(x, \partial_{x} \varphi+\xi\right)+O\left(h^{2}\right) \tag{10.40}
\end{equation*}
$$

In fact, writing $\varphi(x)-\varphi(y)=F(x, y)(x-y)$, we easily check that

$$
\begin{gathered}
e^{-i \varphi / h} p^{\mathrm{w}}(x, h D) e^{i \varphi / h} u= \\
\iint a\left(\frac{x+y}{2}, \xi+F(x, y)\right) e^{i\langle x-y, \xi\rangle / h} u(y) d y d \xi
\end{gathered}
$$

where

$$
F(x, y)=\partial_{x} \varphi\left(\frac{x+y}{2}\right)+O\left((x-y)^{2}\right) .
$$

Hence,

$$
\begin{gathered}
e^{-i \varphi / h} p^{\mathrm{w}}(x, h D) e^{i \varphi / h} u=\iint\left(a\left((x+y) / 2, \xi+\partial_{x} \varphi((x+y) / 2)\right)\right. \\
+\langle e(x, y, \xi)(x-y),(x-y)\rangle) e^{i\langle x-y, \xi\rangle / h} u(y) d y d \xi
\end{gathered}
$$

where the entries of the matrix valued function $e$ are in $S$. Integration by parts based on (9.65) gives (10.40).
2. Recalling from Lemma 10.10 that $\partial_{t} \varphi=-p\left(x, \partial_{x} \varphi\right)$, we then deduce from (10.37) that

$$
\begin{equation*}
\left(h D_{t}+f_{t}^{\mathrm{w}}(x, h D, \eta)\right) b(t, x, \eta)=O\left(h^{2}\right) \tag{10.41}
\end{equation*}
$$

where

$$
f_{t}(x, \xi):=p\left(x, \partial_{x} \varphi(t, x, \eta)+\xi\right)-p\left(x, \partial_{x} \varphi(t, x, \eta)\right)
$$

and where $\eta$ considered as a parameter. So

$$
f_{t}(x, \xi, \eta)=\sum_{j=1}^{n} \xi_{j} \partial_{\xi_{j}} p\left(x, \partial_{x} \varphi(t, x, \eta)\right)+O\left(|\xi|^{2}\right)
$$

Hence for $g=g(t, x, \eta) \in S$,

$$
f_{t}^{\mathrm{w}}(x, h D, \eta) g=\frac{1}{2} \sum_{j=1}^{n}\left(\left(\partial_{\xi_{j}} p\right) h D_{x_{j}} g+h D_{x_{j}}\left(\partial_{\xi_{j}} p g\right)\right)+O\left(h^{2}\right),
$$

in which expression the derivatives of $p$ are evaluated at $\left(x, \partial_{x} \varphi(t, x, \eta)\right)$.
Consequently $b_{0}$ satisfies:

$$
h D_{t} b_{0}+\frac{1}{2} \sum_{j=1}^{n}\left(\partial_{\xi_{j}} p\right) h D_{x_{j}} b_{0}+h D_{x_{j}}\left(\partial_{\xi_{j}} p b_{0}\right)=0
$$

This we rewrite as

$$
\begin{equation*}
\left(\partial_{t}+V_{t}+\frac{1}{2} \operatorname{div} V_{t}\right) b_{0}=0 \tag{10.42}
\end{equation*}
$$

with

$$
V_{t}:=\sum\left(\partial_{\xi_{j}} p\right) \partial_{x_{j}}
$$

3. To understand this equation geometrically, we consider $b_{0}(t, \cdot, \eta)$ as a function on

$$
\Lambda_{t, \eta}:=\left\{\left(x, \partial_{x} \varphi(t, x, \eta)\right)\right\}
$$

Then

$$
\begin{gathered}
\gamma_{s, t}: \Lambda_{t-s, \eta} \rightarrow \Lambda_{t, \eta}, \\
\left.\frac{d}{d s} \gamma_{s, t}^{*} u\right|_{s=0}=\left.H_{p}\right|_{\Lambda_{t, \eta}} u=V_{t} u,
\end{gathered}
$$

for $u \in C^{\infty}$. But equation (10.42) can be further rewritten as

$$
\begin{equation*}
\frac{d}{d t} \gamma_{t}^{*} b_{0}(t, \cdot, \eta)=-\frac{1}{2} \gamma_{t}^{*}\left(\operatorname{div} V_{t} b_{0}(t, \cdot, \eta)\right) \tag{10.43}
\end{equation*}
$$

We claim next that

$$
\begin{equation*}
\gamma_{t}^{*} b_{0}(t, x, \eta)=\left|\partial \gamma_{t}\right|^{-\frac{1}{2}} \tag{10.44}
\end{equation*}
$$

is the solution of (10.43) satisfying $b_{0}(0, x, \eta)=1$. Here $\gamma_{t}$ is considered as a function $\Lambda_{0, \eta} \rightarrow \Lambda_{t, \eta}$. In fact,

$$
\begin{aligned}
\frac{d}{d t}\left|\partial \gamma_{t}\right|^{-\frac{1}{2}} & =\frac{d}{d s}\left|\partial \gamma_{t} \circ \kappa_{s, t}\right|_{s=0}^{-\frac{1}{2}} \\
& =\frac{d}{d s}\left|\partial \gamma_{t}\right|^{-\frac{1}{2}} \gamma_{t}^{*}\left|\partial \gamma_{s, t}\right|_{s=0}^{-\frac{1}{2}} \\
& =-\frac{1}{2} \gamma_{t}^{*} \operatorname{div} V_{t}\left|\partial \gamma_{t}\right|^{-\frac{1}{2}}
\end{aligned}
$$

4. To obtain an explicit formula for $b_{0}$, we recall that

$$
\gamma_{t}^{-1}:\left(x, \partial_{x} \varphi(t, x, \eta)\right) \rightarrow\left(\partial_{\eta} \varphi(t, x, \eta), \eta\right) .
$$

Hence

$$
\partial\left(\left.\gamma_{t}^{-1}\right|_{\Lambda_{t, \eta}}\right)=\partial_{\eta x}^{2} \varphi(t, x, \eta)
$$

and consequently, from (10.44), we see that (10.39) holds.
Proof of Theorem 10.9 Using the same argument for the higher order terms in $b$, we can find its full expansion with all the equations valid in $\left(-t_{0}, t_{0}\right) \times U$. That shows that $U(t)$ given by (10.33) satisfies (10.32), and thereby completes the proof of Theorem 10.9.

EXAMPLE. Revisiting example (10.29), we see that for the phase (10.30) the corresponding amplitude is

$$
b_{0}=\left(\operatorname{det} \partial_{x \eta}^{2} \varphi(x, \eta)\right)^{1 / 2}=(\operatorname{det} A)^{-1 / 2}
$$

REMARK: Amplitudes as half-densities. The somewhat cumbersome derivation of the formula for $b_{0}$, the leading term of the amplitude
$b$ in (10.33), becomes much more natural when we use half-densities, introduced earlier in Section 9.1.

We first make a general observation. If $a:=u|d x|^{\frac{1}{2}}$ is a half-density, and $\gamma_{t}$ is a family of diffeomorphisms generated by a family of vectorfields:

$$
\frac{d}{d t} \gamma_{t}=\left(\gamma_{t}\right)_{*} V_{t}
$$

then

$$
\begin{equation*}
\mathcal{L}_{V_{t}} a:=\frac{d}{d t} \gamma_{t}^{*} a=\left(V_{t} u+\left(\operatorname{div} V_{t} / 2\right) u\right)|d x|^{\frac{1}{2}} \tag{10.45}
\end{equation*}
$$

Indeed,

$$
\gamma_{t}^{*} a=\gamma_{t}^{*} u\left|\partial \gamma_{t}\right|^{\frac{1}{2}}|d x|^{\frac{1}{2}} ;
$$

and if we define

$$
\gamma_{s, t}(x):=\gamma_{t+s}\left(\gamma_{t}^{-1}(x)\right),\left.\quad \frac{d}{d s} \gamma_{s, t}(x)\right|_{s=0}=V_{t}(x)
$$

then

$$
\frac{d}{d t}\left|\partial \gamma_{t}\right|^{\frac{1}{2}}=\frac{d}{d s}\left|\partial \gamma_{t} \circ \kappa_{s, t}\right|^{\frac{1}{2}}=\frac{1}{2}\left|\partial \gamma_{t}\right|^{\frac{1}{2}} \kappa_{t}^{*} \operatorname{div} V_{t}
$$

This means that if we consider $b_{0}(t, x, \eta)|d x|^{\frac{1}{2}}$ as a half-density on $\Lambda_{t, \eta}$ then (10.42) becomes

$$
(d / d t) \gamma_{t}^{*}\left(b_{0}|d x|^{\frac{1}{2}}\right)=\left(\partial_{t}+\mathcal{L}_{V_{t}}\right)\left(b_{0}(t, x, \eta)|d x|^{\frac{1}{2}}\right)=0 .
$$

This is the same as

$$
\gamma_{t}^{*}\left(\left.b_{0}(t, x, \eta)|d x|^{\frac{1}{2}}\right|_{\Lambda_{\varphi_{t, \eta}}}\right)=\left.|d x|^{\frac{1}{2}}\right|_{\Lambda_{\varphi_{0, \eta}}} .
$$

It follows that $\gamma_{t}^{*} b_{0}=\left|\partial \gamma_{t}\right|^{-1 / 2}$, the same conclusion as before.
It is appealing that the amplitude, interpreted as a half-density, is invariant under the flow. When coordinates change, and in particular when we move to larger times at which (10.34) and (10.35) are no longer valid, the statement about the amplitude as a half-density remains simple.

## REMARK: A more general version of oscillatory integral representation.

If we examine the proof of Theorem 10.9 we notice that we did not use the fact that $P=p^{\mathrm{w}}(x, h D)$ is $t$ independent. That means that we can consider the solution of a more general problem,

$$
\left\{\begin{array}{c}
\left(h D_{t}+P(t)\right) F(t) u=0  \tag{10.46}\\
F(0) u=u
\end{array}\right.
$$

where

$$
P(t)=p^{\mathrm{w}}(t, x, h D), \quad p(t, x, \xi) \in C^{\infty}\left(\mathbb{R}_{t}, S\left(\mathbb{R}_{x, \xi}^{2 n}, m\right)\right)
$$

For the approximate solution of this problem we still have the same oscillatory integral representation as the one give in Theorem 10.9. In particular that means that we have an oscillatory integral representation of the family of operators defined in Theorem 10.3 for small values of $t$ there.

For the yet more general problem of $p$ depending on $h$ we refer to [?, Section 7] and references given there. Here we note that the proof works for $P(t)=p^{\mathrm{w}}(t, x, h D)+h^{2} p_{2}^{\mathrm{w}}(t, x, h D)$ and that form of operators acting on half-densities is invariant (see Theorem 9.12).

### 10.3 SEMICLASSICAL STRICHARTZ ESTIMATES, $L^{p}$ BOUNDS

In this section we will use Theorem 10.9 to obtain $L^{p}$ bounds on approximate solutions to

Let $a=a(t, x, \xi) \in C^{\infty}\left(\mathbb{R}, S\left(T^{*} \mathbb{R}^{k}, m\right)\right)$. We introduce the following nondegeneracy condition at $(t, x, \xi)$ :

$$
\begin{equation*}
\partial_{\xi}^{2} a(t, x, \xi) \text { is non-degenerate . } \tag{10.47}
\end{equation*}
$$

REMARK. The Hessian, $\partial_{\xi}^{2} f\left(\xi_{0}\right)$, of a smooth function $f(\xi)$ is not invariantly defined unless $\partial_{\xi} f\left(\xi_{0}\right)=0$. However the statement (10.47) is invariant if only linear transformations in $\xi$ are allowed. That is the case for symbol transformation induced by changes of variables in $x$, see Theorem 9.12.

We consider the problem which essentially the same as (10.46):

$$
\left\{\begin{array}{c}
\left(h D_{t}+A(t)\right) F(t, r) u=0  \tag{10.48}\\
F(r, r) u=u
\end{array}\right.
$$

where $r \in \mathbb{R}$. As discussed in the remark at the end of Section 10.2, Theorem 10.9 gives a description of $F(t, r)$ for small values of $t$.

THEOREM 10.12 (Semiclassical Strichartz estimates). Suppose that $a(t) \in C^{\infty}\left(\mathbb{R}_{t}, S\left(T^{*} \mathbb{R}^{k}, m\right)\right.$ ), is real valued, $\chi \in C_{\mathrm{c}}^{\infty}\left(T^{*} \mathbb{R}^{k}\right)$, and that (10.47) holds in $\operatorname{spt}(\chi), t \in \mathbb{R}$. With $A(t):=a^{\mathrm{w}}(t, x, h D)$, let $F(t, r)$ be the solution of (10.48). Then for $\psi \in C_{\mathrm{c}}^{\infty}(\mathbb{R})$ with support sufficiently close to 0 , any $I \subset \subset \mathbb{R}$, and

$$
U(t, r):=\psi(t) F(t, r) \chi^{\mathrm{w}}(x, h D) \text { or } U(t, r):=\psi(t) \chi^{\mathrm{w}}(x, h D) F(t, r)
$$

we have

$$
\begin{gather*}
\sup _{r \in I}\left(\int_{\mathbb{R}}\|U(t, r) f\|_{L^{q}\left(\mathbb{R}^{k}\right)}^{p} d t\right)^{\frac{1}{p}} \leq B h^{-\frac{1}{p}}\|f\|_{L^{2}\left(\mathbb{R}^{k}\right)},  \tag{10.49}\\
\frac{2}{p}+\frac{k}{q}=\frac{k}{2}, \quad 2 \leq p \leq \infty, \quad 1 \leq q \leq \infty, \quad(p, q) \neq(2, \infty) .
\end{gather*}
$$

Proof: 1. In view of Theorem C. 10 we need to show that

$$
\begin{equation*}
\left\|U(t, r) U(s, r)^{*} f\right\|_{L^{\infty}(X, \mu)} \leq A h^{-k / 2}|t-s|^{-k / 2}, \quad t, s \in \mathbb{R} \tag{10.50}
\end{equation*}
$$ with constants independent of $r \in I$. We can put $r=0$ in the argument and drop the dependence on $r$ in $U$ and $F$.

2. We use Theorem 10.9. The construction there and the assumption that $\chi \in C_{\mathrm{c}}^{\infty}$ show that

$$
U(t)=\widetilde{U}(t)+E(t),
$$

where

$$
E(t)=O\left(h^{\infty}\right): \mathcal{S}^{\prime} \rightarrow \mathcal{S},
$$

and the Schwartz kernel of $\widetilde{U}(t)$ is

$$
\begin{gather*}
\widetilde{U}(t, x, y)=\frac{1}{(2 \pi h)^{k}} \int_{\mathbb{R}^{k}} e^{\frac{i}{h}(\varphi(t, x, \eta)-\langle y, \eta\rangle)} \tilde{b}(t, y, x, \eta ; h) d \eta, \\
\tilde{b} \in S \cap C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{1+3 k}\right), \quad \varphi(0, x, \eta)=\langle x, \eta\rangle,  \tag{10.51}\\
\partial_{t} \varphi(t, x, \eta)+a\left(t, x, \partial_{x} \varphi(t, x, \eta)\right)=0 .
\end{gather*}
$$

3. Hence we only need to prove (10.50) with $U$ replaced by $\widetilde{U}$ and that means that we need an $L^{\infty}$ bound on the Schwartz kernel of $W(t, s):=\widetilde{U}(t) \widetilde{U}(s)^{*}$ :

$$
W(t, s, x, y)=\frac{1}{(2 \pi h)^{2 k}} \int_{\mathbb{R}^{3 k}} e^{\frac{i}{h}(\varphi(t, x, \eta)-\varphi(s, y, \zeta)-\langle z, \eta-\zeta\rangle)} B d z d \zeta d \eta,
$$

where

$$
B=B(t, s, x, y, z, \eta, \zeta ; h) \in S \cap C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{2+6 k}\right) .
$$

4. The phase is nondegenerate in $(z, \zeta)$ variables and stationary for $\zeta=\eta, z=\partial_{\zeta} \varphi(s, y, \zeta)$. Hence we can apply Theorem 3.15 to obtain

$$
W(t, s, x, y)=\frac{1}{(2 \pi h)^{k}} \int_{\mathbb{R}^{k}} e^{\frac{i}{h}(\varphi(t, x, \eta)-\varphi(s, y, \eta))} B_{1}(t, s, x, y, \eta ; h) d \eta,
$$

where $B_{1} \in S \cap C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{2+3 k}\right)$. We now rewrite the phase as follows:

$$
\begin{gathered}
\widetilde{\varphi}:=\varphi(t, x, \eta)-\varphi(s, y, \eta)=(t-s)(a(0, x, \eta)+O(|t|+|s|)) \\
+\langle x-y, \eta+s F(s, x, y, \eta)\rangle, \quad F \in C^{\infty}\left(\mathbb{R}^{1+3 k}\right),
\end{gathered}
$$

where using (10.51) we wrote

$$
\varphi(s, x, \eta)-\varphi(s, y, \eta)=\langle x-y, \eta\rangle+\langle x-y, s F(s, x, y, \eta)\rangle .
$$

5. The phase is stationary when

$$
\partial_{\eta} \widetilde{\varphi}=\left(I+s \partial_{\eta} F\right)(x-y)+(t-s)\left(\partial_{\eta} a+O(|t|+|s|)\right)=0,
$$

and in particular, for $s$ small, having a stationary point implies

$$
x-y=O(t-s),
$$

as then $\left(I+s \partial_{\eta} F\right)$ is invertible. The Hessian is given by

$$
\begin{aligned}
\partial_{\eta}^{2} \widetilde{\varphi} & =s\left(\partial_{\eta}^{2} F\right)(x-y)+(t-s)\left(\partial_{\eta}^{2} a+O(|t|+|s|)\right) \\
& =(t-s)\left(\partial_{\eta}^{2} a+O(|t|+|s|)\right),
\end{aligned}
$$

where $\partial_{\eta}^{2} a=\partial_{\eta}^{2} a(0, x, \eta)$.
6. Hence, for $t$ and $s$ sufficiently small, that is for a suitable choice of the support of $\psi$ in the definition of $U(\cdot)$, the nondegeneracy assumption (10.47) implies that at the critical point

$$
\partial_{\eta}^{2} \widetilde{\varphi}=(t-s) \psi(x, y)
$$

If $|t-s|>M h$ where $M$ is a large constant we can use the stationary phase estimate in Theorem 3.15 to see that

$$
|W(t, s, x, y)| \leq C h^{-k / 2}|t-s|^{-k / 2}
$$

When $|t-s|<M h$ we see that the crude upper bound on the integral gives

$$
\begin{aligned}
|W(t, s, x, y)| & \leq \frac{1}{(2 \pi h)^{2 k}} \int_{\mathbb{R}^{3 k}}|B(t, s, x, y, z, \eta, \zeta ; h)| d z d \zeta d \eta \\
& \leq C h^{-k} \leq C^{\prime} h^{-k / 2}|t-s|^{-k / 2},
\end{aligned}
$$

which is what we need to apply Theorem C.10.
We formulate the following nondegeneracy assumptions at $\left(x_{0}, \xi_{0}\right) \in$ $T^{*} \mathbb{R}^{n}$ :

$$
\begin{equation*}
p\left(x_{0}, \xi_{0}\right)=0 \Longrightarrow \partial_{\xi} p\left(x_{0}, \xi_{0}\right) \neq 0 \tag{10.52}
\end{equation*}
$$

Consequently, the set

$$
\left\{\xi: p\left(x_{0}, \xi\right)=0\right\}
$$

is a smooth hypersurface in $\mathbb{R}^{n}$ near $\xi_{0}$. We then assume that the second fundamental form of this hypersuface is nondegenerate at $\xi_{0}$.

We can reformulate this as follows. By a linear change of variables assume that $\partial_{\xi} p\left(x_{0}, \xi_{0}\right)=(\rho, 0, \cdots, 0), \rho \neq 0$. Then near $\left(x_{0}, \xi_{0}\right)$,

$$
p(x, \xi)=e(x, \xi)\left(\xi_{1}-a\left(x, \xi^{\prime}\right)\right)
$$

and our assumption on the curvature becomes

$$
\begin{equation*}
\partial_{\xi^{\prime}}^{2},\left(x_{0}, \xi_{0}^{\prime}\right) \text { is nondegenerate. } \tag{10.53}
\end{equation*}
$$

THEOREM 10.13 ( $L^{p}$ bounds on approximate solutions). Suppose that $u(h),\|u(h)\|_{L^{2}}=1$, satisfies the frequency localization condition (9.8). Suppose also that (10.53) is satisfied in $W F_{h}(u)$, and that

$$
\begin{equation*}
p^{\mathrm{w}}(x, h D) u(h)=O_{L^{2}}(h) . \tag{10.54}
\end{equation*}
$$

Then for $p=2(n+1) /(n-1)$, and any $K \subset \subset \mathbb{R}^{n}$,

$$
\begin{equation*}
\|u(h)\|_{L^{p}(K)}=O\left(h^{-1 / p}\right) . \tag{10.55}
\end{equation*}
$$

REMARK. The first example in the remark after Theorem 9.5 shows that the curvature condition (10.53) is in general necessary. In fact, if $P(h)=h D_{x_{1}}$ and $u(h)=h^{-(n-1) / 2} \chi\left(x_{1}\right) \chi\left(x^{\prime} / h\right)$ then for $p=2(n+$ 1)/ $(n-1)$,

$$
\|u\|_{L^{p}} \simeq h^{(n-1)(1 / p-1 / 2)}=h^{-(n-1) /(n+1)} \neq O\left(h^{-1 / p}\right) .
$$

However for the simplest case in which (10.53) holds,

$$
p(x, \xi)=\xi_{1}-\xi_{2}^{2}-\cdots-\xi_{n}^{2}
$$

the estimate (10.55) is optimal. To see that put

$$
u(h):=h^{-(n-1) / 4} \chi_{0}\left(x_{1}\right) \exp \left(-\left|x^{\prime}\right|^{2} / 2 h\right),
$$

where $x=\left(x_{1}, x^{\prime}\right), \chi_{0} \in C_{\mathrm{c}}^{\infty}(\mathbb{R})$. Then

$$
\begin{gathered}
\left(-h^{2} \Delta_{x^{\prime}}+\left|x^{\prime}\right|^{2}\right) u(h)=(n-1) h u(h), \\
\|u(h)\|_{L^{2}} \simeq 1,\left|x^{\prime}\right|^{2 k} u(h)=O_{L^{2}}\left(h^{k}\right) . \text { Hence } \\
p^{\mathrm{w}}(x, h D) u(h)=O_{L^{2}}(h),
\end{gathered}
$$

and

$$
\|u(h)\|_{L^{p}\left(\mathbb{R}^{n}\right)} \simeq h^{(n-1)(2 / p-1) / 4}=h^{-1 / p}, \quad p=2(n+1) /(n-1) .
$$

Before proving Theorem 10.13 we prove a lemma which is a consequence of Theorem 10.12

LEMMA 10.14. In the notation of Theorem 10.13 we have

$$
\begin{equation*}
\left\|\int_{0}^{t} U(t, s) \mathbf{1}_{I}(s) f(s, x) d s\right\|_{L^{p}\left(\mathbb{R}_{t} \times \mathbb{R}_{x}^{k}\right)} \leq C h^{-1 / p} \int_{\mathbb{R}}\|f(s, x)\|_{L^{2}\left(\mathbb{R}_{x}^{k}\right)} d s \tag{10.56}
\end{equation*}
$$

Proof: 1. We apply the integral version of Minkowski's inequality:

$$
\begin{aligned}
& \left\|\int_{0}^{t} U(t, s) \mathbf{1}_{I}(s) f(s, x) d s\right\|_{L^{p}\left(\mathbb{R}_{t} \times \mathbb{R}_{x}^{k}\right)} \\
& \quad \leq C \int_{I \cap \mathbb{R}_{+}}\left\|\mathbf{1}_{[s, \infty)}(t) U(t, s) f(s, x)\right\|_{L^{p}\left(\mathbb{R}_{t} \times \mathbb{R}_{x}^{k}\right)} d s \\
& \quad \leq C \int_{I \cap \mathbb{R}_{+}}\|U(t, s) f(s, x)\|_{L^{p}\left(\mathbb{R}_{t} \times \mathbb{R}_{x}^{k}\right)} d s .
\end{aligned}
$$

2. Now we use the estimate (10.49) with $p=q=2(n+1) /(n-1)$ :

$$
\|U(t, s) f(s, x)\|_{L^{p}\left(\mathbb{R}_{t} \times \mathbb{R}_{x}^{k}\right)} \leq C h^{-1 / p}\|f(s, x)\|_{L^{2}\left(\mathbb{R}_{x}^{k}\right)},
$$

from which ( $I$ is compact) (10.56) follows.

Proof of Theorem 10.13: 1. We follow the same procedure as in the proof of Theorem 9.5. As in that case the condition (10.54) is local in phase space, that is, it implies that for any $\chi \in C_{\mathrm{c}}^{\infty}\left(T^{*} \mathbb{R}^{n}\right)$,

$$
p^{\mathrm{w}}(x, h D) \chi^{\mathrm{w}}(x, h D) u(h)=O(h),
$$

in $L^{2}$.
2. We factorize $p(x, \xi)$ as in (9.14) and we conclude that for $\chi$ with sufficiently small support,

$$
\left(h D_{x_{1}}-a\left(x, h D_{x^{\prime}}\right)\right) \chi^{\mathrm{w}}(x, h D) u(h)=O_{L^{2}}(h) .
$$

Let

$$
f\left(x_{1}, x^{\prime}, h\right)=\left(h D_{x_{1}}-a\left(x, h D_{x^{\prime}}\right)\right)\left(\chi^{\mathrm{w}} u(h)\right) .
$$

Since $\|f\|_{L^{2}}=O(h)$, we see

$$
\begin{equation*}
\int_{\mathbb{R}}\left\|f\left(x_{1}, \cdot\right)\right\|_{L^{2}\left(\mathbb{R}^{n-1}\right)} d x_{1} \leq C\|f\|_{L^{2}\left(\mathbb{R}^{n}\right)}=O(h) \tag{10.57}
\end{equation*}
$$

3. We now apply Theorem 10.12 with $t=x_{1}$ and $x$ replaced by $x^{\prime} \in \mathbb{R}^{n-1}$, that is $k=n-1$. The assumption (10.53) shows that $\partial_{\xi^{\prime}}^{2} a$ is nondegenerate in the support of $\chi$. We can choose $\psi$ and $\chi$ in the definition of $U(t)$ in the statement of Theorem 10.12 so that

$$
\chi^{\mathrm{w}}(x, h D) u\left(x_{1}, x^{\prime}, h\right)=\frac{i}{h} \int_{0}^{x_{1}} U(t, s) f\left(s, x^{\prime}\right) d s+O_{\mathcal{S}}\left(h^{\infty}\right) .
$$

Let us choose $p=q, k=n-1$ in (10.49), that is,

$$
p=q=\frac{2(n+1)}{n-1} .
$$

Then, using (10.49), (10.57), and (10.56),

$$
\begin{aligned}
\left\|\chi^{\mathrm{w}}(x, h D) u\right\|_{L^{p}} & \leq \frac{1}{h} h^{-1 / p} \int_{\mathbb{R}}\|f(s, \cdot, h)\|_{L^{2}\left(\mathbb{R}^{n-1}\right)} d s+O\left(h^{\infty}\right) \\
& =O\left(h^{-1 / p}\right)
\end{aligned}
$$

A partition of unity argument used in the proof of Theorem 9.5 concludes the proof.

As a corollary we obtain Sogge's bounds on spectral clusters on Riemannian manifolds:

THEOREM 10.15 ( $L^{p}$ bounds on eigenfuctions). Suppose that $M$ is an n-dimensional compact Riemannian manifold and let $\Delta_{g}$ be its Laplace-Beltrami operator. If

$$
0=\lambda_{0}<\lambda_{1} \leq \cdots \lambda_{j} \rightarrow \infty
$$

is the complete set of eigenvalues of $-\Delta_{g}$, and

$$
-\Delta_{g} \varphi_{j}=\lambda_{j} \varphi_{j}
$$

are the corresponding eigenfunctions, then for any $c_{j} \in \mathbb{C}, j=0,1, \cdots$,

$$
\begin{align*}
& \left\|\sum_{\mu \leq \sqrt{\lambda_{j}} \leq \mu+1} c_{j} \varphi_{j}\right\|_{L^{p}} \leq C \mu^{\sigma(p)}\left\|\sum_{\mu \leq \sqrt{\lambda_{j}} \leq \mu+1} c_{j} \varphi_{j}\right\|_{L^{2}}, \\
& \sigma(p)= \begin{cases}\frac{n-1}{2}\left(\frac{1}{2}-\frac{1}{p}\right) & \text { for } 2 \leq p \leq \frac{2(n+1)}{n-1}, \\
\frac{n-1}{2}-\frac{n}{p} & \text { for } \frac{2(n+1)}{n-1} \leq p \leq \infty .\end{cases} \tag{10.58}
\end{align*}
$$

In particular

$$
\begin{equation*}
\left\|\varphi_{j}\right\|_{L^{p}} \leq C \lambda_{j}^{\sigma(p) / 2}\left\|\varphi_{j}\right\|_{L^{2}} . \tag{10.59}
\end{equation*}
$$

Proof: 1. We argue as in the proof of Theorem 9.6 but now we need to check is the curvature assumption (10.53): at any point $\left(x_{0}, \xi_{0}\right)$ and for suitable coordinates

$$
p\left(x_{0}, \xi\right)=|\xi|^{2}-1, \quad \xi_{0}=(1,0, \cdots, 0)
$$

The hypersurface $p\left(x_{0}, \xi\right)=0$ is the unit sphere in $\mathbb{R}_{\xi}^{n}$ and it has a nondegenerate second fundamental form.
2. Complex interpolation [H1, Theorem 7.1.12] between the estimate in Theorem 9.6, the trivial $L^{2}$ estimate, and the estimate in Theorem 10.13 gives the full result.

### 10.4 MORE SYMPLECTIC GEOMETRY

To further apply the local theory of quantized symplectic transformations to the study of semiclassical operators we will need two results from symplectic geometry. The first is a stronger form of Darboux's Theorem 2.11, which we state without proof.

THEOREM 10.16 (Variant of Darboux's Theorem). Let $A$ and $B$ be two subsets of $\{1, \cdots, n\}$, and suppose that

$$
p_{j}(x, \xi) \quad(j \in A), \quad q_{k}(x, \xi) \quad(k \in B)
$$

are smooth, real-valued functions defined in a neighbourhood of $(0,0) \in$ $\mathbb{R}^{2 n}$, with linearly independent gradients at $(0,0)$.

If

$$
\begin{gather*}
\left\{q_{i}, q_{j}\right\}=0 \quad(i, j \in A), \quad\left\{p_{k}, p_{l}\right\}=0 \quad(k, l \in B),  \tag{10.60}\\
\left\{p_{k}, q_{j}\right\}=\delta_{k j} \quad(j \in A, k \in B),
\end{gather*}
$$

then there exists a symplectomorphism $\kappa$, locally defined near $(0,0)$, such that $\kappa(0,0)=(0,0)$ and

$$
\begin{equation*}
\kappa^{*} q_{j}=x_{j} \quad(j \in A), \quad \kappa^{*} p_{k}=\xi_{j} \quad(k \in B) . \tag{10.61}
\end{equation*}
$$

See Hörmander [H2, Theorem 21.1.6] for an elegant exposition.
The next result is less standard and comes from the work of Duistermaat and Sjöstrand: consult Hörmander [H2, Lemma 21.3.4] for the proof.

THEOREM 10.17 (Symplectic integrating factor). Let $p$ and $q$ be smooth, real-valued functions defined near $(0,0) \in \mathbb{R}^{2 n}$, satisfying

$$
\begin{equation*}
p(0,0)=q(0,0)=0, \quad\{p, q\}(0,0)>0 \tag{10.62}
\end{equation*}
$$

Then there exists a smooth, positive function $u$ for which

$$
\begin{equation*}
\{u p, u q\} \equiv 1 \tag{10.63}
\end{equation*}
$$

in a neighborhood of $(0,0)$.

### 10.5 NORMAL FORMS FOR OPERATORS WITH REAL SYMBOLS

Operators of real principal type. Recall that we are taking our order function to be

$$
m:=\left(1+|x|^{2}+|\xi|^{2}\right)^{\frac{k}{2}}
$$

Set $P=p^{\mathrm{w}}(x, h D ; h)$, where

$$
p(x, \xi ; h) \sim p_{0}(x, \xi)+h p_{1}(x, \xi)+\cdots+h^{N} p_{N}(x, \xi)+\cdots,
$$

for $p_{j} \in S(m)$. We assume that the real-valued principal symbol $p_{0}$ satisfies

$$
\begin{equation*}
p_{0}(0,0)=0, \quad \partial p_{0}(0,0) \neq 0 \tag{10.64}
\end{equation*}
$$

and then say that $P$ is an operator of real principal type at $(0,0)$.

THEOREM 10.18 (Normal form for real principal type operators). Suppose that $P=p^{\mathrm{w}}(x, h D ; h)$ is a semiclassical real principal type operator at $(0,0)$.

Then there exist
(i) a local canonical transformation $\kappa$ defined near $(0,0)$, such that $\kappa(0,0)=(0,0)$ and

$$
\begin{equation*}
\kappa^{*} \xi_{1}=p_{0} \tag{10.65}
\end{equation*}
$$

and
(ii) an operator $T$, quantizing $\kappa$ in the sense of Theorem 10.6, such that

$$
\begin{equation*}
T^{-1} \text { exists microlocally near }((0,0),(0,0)) \tag{10.66}
\end{equation*}
$$

and

$$
\begin{equation*}
T P T^{-1}=h D_{x_{1}} \quad \text { microlocally near }((0,0),(0,0)) \tag{10.67}
\end{equation*}
$$

INTERPRETATION. The point is that using this theorem, we can transplant various mathematical objects related to $P$ to others related to $h D_{x_{1}}$, which are much easier to study. A simple example is given by the following estimate:

$$
\|u\| \leq \frac{C}{h}\|P u\|
$$

when $u=u(h) \in \mathcal{S}$ has $\mathrm{WF}_{h}(u)$ in a small neighbourhood of $(0,0)$.
Proof. 1. Theorem 10.16 applied with $A=\emptyset$ and $B=\{1\}$, provides $\kappa$ satisfying (10.65) near $(0,0)$. Then Theorem 10.1 gives us a family of symplectic transfomations $\gamma_{t}$ for $0 \leq t \leq 1$.

Let $F(t)$ be defined using the family $\gamma_{t}$ in Theorem 10.6, and put $T_{0}=F(1)$. Then

$$
T_{0} P-h D_{x_{1}}=E \text { microlocally near }(0,0) .,
$$

for $E=\operatorname{Op}(e), e \in S^{-1}$.
2. We now look for a symbol $a \in S$ so that $a$ is elliptic at $(0,0)$ and

$$
h D_{x_{1}}+E=A h D_{x_{1}} A^{-1} \text { microlocally near }(0,0)
$$

for $A:=\operatorname{Op}(a)$. This is the same as solving

$$
\left[h D_{x_{1}}, A\right]+E A=0 .
$$

Since $P=p_{0}^{\mathrm{w}}+h p_{1}^{\mathrm{w}}+h^{2} p_{2}^{\mathrm{w}}+\cdots$, the Remark after the proof of Theorem 10.2 shows that

$$
e(x, \xi ; h)=h e_{0}(x, \xi)+h^{2} e_{1}(x, \xi)+\cdots .
$$

Hence we can find $a_{0} \in S$ such that $a_{0}(0,0) \neq 0$ and

$$
\frac{1}{i}\left\{\xi_{1}, a_{0}\right\}+e_{0} a_{0}=0
$$

near $(0,0)$.
Define $A_{0}:=\operatorname{Op}\left(a_{0}\right)$; then

$$
\left[h D_{x_{1}}, A_{0}\right]+E A_{0}=\operatorname{Op}\left(r_{0}\right)
$$

for a symbol $r_{0} \in S^{-2}$.
3. We now inductively find $A_{j}=\operatorname{Op}\left(a_{j}\right)$, for $a_{j} \in S^{-j}$, satisfying

$$
\left[h D_{x_{1}}, A_{0}+A_{1}+\cdots+A_{N}\right]+E\left(A_{0}+A_{1}+\cdots A_{N}\right)=\operatorname{Op}\left(r_{N}\right),
$$

for $r_{N} \in S^{-N-2}(1)$. We then put

$$
A \sim A_{1}+A_{2}+\cdots+A_{N}+\cdots,
$$

which is elliptic near $(0,0)$. Finally, define

$$
T:=A^{-1} T_{0} .
$$

This operator quantizes $\kappa$ in the sense of Theorem 10.6.

### 10.6 NORMAL FORMS FOR OPERATORS WITH COMPLEX SYMBOLS

Operators of complex principal type. Assume as before that $P=$ $p^{\mathrm{w}}(x, h D ; h)$ has the symbol

$$
p(x, \xi ; h) \sim p_{0}(x, \xi)+h p_{1}(x, \xi)+\cdots+h^{N} p_{N}(x, \xi)+\cdots
$$

with $p_{j} \in S(m)$. We now allow $p(x, \xi)$ to be complex-valued, and still say that $P$ is principal type at $(0,0)$ if

$$
p_{0}(0,0)=0, \quad \partial p_{0}(0,0) \neq 0
$$

Discussion. If $\partial\left(\operatorname{Re} p_{0}\right)$ and $\partial\left(\operatorname{Im} p_{0}\right)$ are linearly independent, then the submanifold of $\mathbb{R}^{2 n}$ where $P$ is not elliptic has codimension two -
as opposed to codimension one in the real-valued case. The symplectic form restricted to that submanifold is non-degenerate if

$$
\left\{\operatorname{Re} p_{0}, \operatorname{Im} p_{0}\right\} \neq 0
$$

Under this assumption a combination of Theorems 10.16 and 10.17 shows that there exists a canonical transformation $\kappa$, defined near $(0,0)$, and a smooth positive function $u$ such that

$$
\kappa^{*}\left(\xi_{1} \pm i x_{1}\right)=u p_{0} .
$$

That is, after a multiplication by a function we obtain the symbol of the creation or annihilation operator for the harmonic oscillator in the $\left(x_{1}, \xi_{1}\right)$ variables. (Recall the discussion of the harmonic oscillator in Section 6.1.)

THEOREM 10.19 (Normal form for the complex symplectic case). Suppose that $P=p^{\mathrm{w}}(x, h D ; h)$ is a semiclassical principal type operator at $(0,0)$, with principal symbol $p_{0}$ satisfying

$$
\begin{equation*}
p_{0}(0,0)=0, \quad \pm\left\{\operatorname{Re} p_{0}, \operatorname{Im} p_{0}\right\}(0,0)>0 . \tag{10.68}
\end{equation*}
$$

Then there exist
(i) a local canonical transformation $\kappa$ defined near $(0,0)$ and a smooth function $u$ such that $\kappa(0,0)=(0,0), u(0,0)>0$, and

$$
\kappa^{*}\left(\xi_{1} \pm i x_{1}\right)=u p_{0} ;
$$

and (ii) an operator $T$, quantizing $\kappa$ in the sense of Theorem 10.6, and a pseudodifferential operator A, elliptic at $(0,0)$, such that

$$
\begin{equation*}
T^{-1} \text { exists microlocally near }((0,0),(0,0)) \tag{10.69}
\end{equation*}
$$

and
(10.70) $\quad T P T^{-1}=A\left(h D_{x_{1}} \pm i x_{1}\right) \quad$ microlocally near $((0,0),(0,0))$.

INTERPRETATION. We can transplant mathematical objects related to $P$ to others related to $A\left(h D_{x_{1}} \pm i x_{1}\right)$, which are clearly much easier to study.

Proof. 1. We start as in the proof of Theorem 10.18. To simplify the notation, let us assume

$$
\left\{\operatorname{Re} p_{0}, \operatorname{Im} p_{0}\right\}>0 .
$$

As noted above, using Theorems 10.16 and 10.17 we can find a smooth function $u$, with $u(0,0)>0$, and a local canonical transformation $\kappa$ such that $\kappa(0,0)=(0,0)$ and $\kappa^{*}\left(\xi_{1}+i x_{1}\right)=u p_{0}$.

Quantizing as before, we obtain an operator $T_{0}$ satisfying

$$
\begin{equation*}
T_{0} P=Q\left(h D_{x_{1}}+i x_{1}+E\right) T_{0}, \tag{10.71}
\end{equation*}
$$

where $Q=\operatorname{Op}(q)$ for a function $q$ satisfying

$$
\gamma^{*} q=1 / u
$$

and $E=\operatorname{Op}(e)$ for some $e \in S^{-1}$.
2. We now need to find pseudodifferential operators $B$ and $C$, elliptic at $(0,0)$, and such that

$$
\begin{equation*}
\left(h D_{x_{1}}+i x_{1}+E\right) B \equiv C\left(h D_{x_{1}}+i x_{1}\right), \tag{10.72}
\end{equation*}
$$

microlocally near $(0,0)$. As in the proof of Theorem 10.18, we have

$$
E=\operatorname{Op}(e), \quad e=h e_{0}(x, \xi)+h^{2} e_{1}(x, \xi)+\cdots
$$

We will find the symbols of $B$ and $C$ by computing successive terms in their expansions:

$$
\begin{aligned}
& b \sim b_{0}+h b_{1}+\cdots+h^{N} b_{N}+\cdots, \\
& c \sim c_{0}+h c_{1}+\cdots+h^{N} c_{N}+\cdots .
\end{aligned}
$$

3. Let us rewrite (10.72) as

$$
\left(h D_{x_{1}}+i x_{1}+E\right) B-C\left(h D_{x_{1}}+i x_{1}\right)=\operatorname{Op}(r),
$$

for

$$
r(x, \xi)=r_{0}(x, \xi)+h r_{1}(x, \xi)+\cdots+h^{N} r_{N}(x, \xi)+\cdots,
$$

with

$$
\begin{aligned}
r_{0} & =\left(\xi_{1}+i x_{1}\right)\left(b_{0}-c_{0}\right), \\
r_{1} & =\left(\xi_{1}+i x_{1}\right)\left(b_{1}-c_{1}\right)+e_{0} b_{0}+\left\{\xi_{1}+i x_{1}, b_{0}\right\} / 2 i-\left\{c_{0}, \xi_{1}+i x_{1}\right\} / 2 i .
\end{aligned}
$$

Here we used composition formula in Weyl calculus (see Theorem 4.6).
We want to choose $b$ and $c$ so that $r_{j} \equiv 0$ for all $j$. For $r_{0}=0$ we simply need $b_{0}=c_{0}$. Then to obtain $r_{1}=0$ we have to solve

$$
-i\left(\partial_{x_{1}}-i \partial_{\xi_{1}}\right) b_{0}+e_{0} b_{0}+\left(\xi_{1}+i x_{1}\right)\left(b_{1}-c_{1}\right)=0
$$

4. We first find $b_{0}$ such that

$$
\left\{\begin{aligned}
-i\left(\partial_{x_{1}}-i \partial_{\xi_{1}}\right) b_{0}+e_{0} b_{0} & =O\left(x_{1}^{\infty}\right) \\
\left.b_{0}\right|_{x_{1}=0} & =1 ;
\end{aligned}\right.
$$

that is, the left hand side vanishes to infinite order at $x_{1}=0$, and $b_{0}=1$ there. The derivatives $\left.\partial_{x_{1}}^{k} e_{0}\right|_{x_{1}=0}$ determine $\left.\partial_{x_{1}}^{k} b_{0}\right|_{x_{1}=0}$. Then Borel's Theorem 4.11 produces a smooth function $b_{0}$ with these prescribed derivatives.
5. With $b_{0}=c_{0}$ chosen that way we see that

$$
t_{1}:=\left(-i\left(\partial_{x_{1}}-i \partial_{\xi_{1}}\right) b_{0}+e_{0} b_{0}\right) /\left(\xi_{1}+i x_{1}\right)
$$

is a smooth function: the numerator vanishes to infinite order on the zero set of the denominator. If we put

$$
\begin{equation*}
c_{1}=b_{1}+t_{1} \tag{10.73}
\end{equation*}
$$

then $r_{1}=0$.
6. Now, using (10.73) the same calculation as before, we see that

$$
r_{3}=\left(\xi_{1}+i x_{1}\right)\left(b_{2}-c_{2}\right)+e_{0} b_{1}-i\left\{\xi_{1}+i x_{1}, b_{1}\right\}+\tilde{r}_{3},
$$

where $\tilde{r}_{3}$ depends only on $b_{0}=c_{0}, t_{1}$, and $e$. Hence $\tilde{r}_{3}$ is already determined. We proceed as in Step 4 and first solve

$$
\left\{\begin{aligned}
-i\left(\partial_{x_{1}}-i \partial_{\xi_{1}}\right) b_{1}+e_{0} b_{1}+\tilde{r}_{3} & =O\left(x_{1}^{\infty}\right) \\
\left.b_{1}\right|_{x=1} & =0
\end{aligned}\right.
$$

This determines $b_{1}$ (and hence $c_{1}$ ). We continue in the same way to determine $b_{2}$ (and hence $c_{2}$ ). An iteration of the argument completes the construction of $b$ and $c$, for which (10.72) holds microlocally near $(0,0)$.
7. Finally, we put $T=B^{-1} T_{0}$, where $B^{-1}$ is the microlocal inverse of $B$ near $(0,0)$, and $A=B^{-1} Q C$, to obtain the statement of the theorem.

### 10.7 SEMICLASSICAL PSEUDOSPECTRA

We present in this last section an application to the so-called semiclassical pseudospectrum. Recall from Chapter 6 that if $P=P(h)=$ $-h^{2} \Delta+V(x)$ and $V$ is real-valued, satisfying
(10.74) $V \in S\left(\langle x\rangle^{m}\right), \quad\left|\xi^{2}+V(x)\right| \geq\left(1+|\xi|^{2}+|x|^{m}\right) / C$ for $|x| \geq C$,
then the spectrum of $P$ is discrete. (We deduced this from the meromorphy of the resolvent of $P, R(z)=(P-z)^{-1}$.)
Quasimodes. Because of the Spectral Theorem, which is applicable as $V$ is real, we also know that approximate location of eigenvalues is implied by the existence of approximate eigenfunctions, called quasimodes. Indeed suppose that

$$
\begin{equation*}
\|(P-z(h)) u(h)\|=O\left(h^{\infty}\right),\|u(h)\|=1 . \tag{10.75}
\end{equation*}
$$

Then there exist $E(h)$ and $v(h)$ such that

$$
\begin{equation*}
(P-E(h)) v(h)=0,\|v(h)\|=1,|E(h)-z(h)|=O\left(h^{\infty}\right) \tag{10.76}
\end{equation*}
$$

In other words, if we can solve (10.75), then the approximate eigenvalue $z(h)$ is in fact close to a true eigenvalue $E(h)$ (although $u(h)$ need not be close to a true eigenfunction $v(h)$.)
Nonnormal operators. But it is well known that this is not the case for nonnormal operators $P$, for which the commutator $\left[P^{*}, P\right]$ does not vanish. Now if If $p=|\xi|^{2}+V(x)$, then the symbol of this commutator is

$$
\begin{equation*}
\frac{1}{i}\{\bar{p}, p\}=2\{\operatorname{Re} p, \operatorname{Im} p\} \tag{10.77}
\end{equation*}
$$

and when this is nonzero we are in the situation discussed in Theorem 10.19. This discussion leads us to

THEOREM 10.20 (Quasimodes). Suppose that $P=-h^{2} \Delta+V(x)$ and that

$$
\begin{equation*}
z_{0}=\xi_{0}^{2}+V\left(x_{0}\right), \quad \operatorname{Im}\left\langle\xi_{0}, \partial V\left(x_{0}\right)\right\rangle \neq 0 \tag{10.78}
\end{equation*}
$$

Then there exists a family of functions $u(h) \in C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{n}\right)$ such that

$$
\begin{equation*}
\left\|\left(P-z_{0}\right) u(h)\right\|_{L^{2}}=O\left(h^{\infty}\right), \quad\|u(h)\|_{L^{2}}=1 \tag{10.79}
\end{equation*}
$$

Moreover, we can choose $u(h)$ so that

$$
\begin{equation*}
\mathrm{WF}_{h}(u(h))=\left\{\left(x_{0}, \xi_{0}\right)\right\}, \quad \operatorname{Im}\left\langle\xi_{0}, \partial V\left(x_{0}\right)\right\rangle<0 \tag{10.80}
\end{equation*}
$$

Proof. We first replace $V$ by a compactly supported potential agreeing with $V$ near $x_{0}$. Our function $u(h)$ will be constructed with support near $x_{0}$.

By changing the sign of $\xi_{0}$ if necessary, but without changing $z_{0}$, we can assume that

$$
\{\operatorname{Re} p, \operatorname{Im} p\}\left(x_{0}, \xi_{0}\right)=2 \operatorname{Im}\left\langle\xi_{0}, \partial V\left(x_{0}\right)\right\rangle<0
$$

According Theorem 10.19, $P-z_{0}$ is microlocally conjugate to

$$
A\left(h D_{x_{1}}-i x_{1}\right) \quad \text { near }\left(\left(x_{0}, \xi_{0}\right),(0,0)\right) .
$$

Let

$$
u_{0}(x, h):=\exp \left(-|x|^{2} / 2 h\right)
$$

so that

$$
\left(h D_{x_{1}}-i x_{1}\right) u_{0}(h)=0, \quad \mathrm{WF}_{h}\left(u_{0}(h)\right)=\{(0,0)\}
$$

Following the notation of Theorem 10.19, we define $u(h):=T^{-1} u_{0}(h)$. Then $\mathrm{WF}_{h}(u(h))=\left\{\left(x_{0}, \xi_{0}\right)\right\}$ and

$$
\left(P-z_{0}\right) u(h) \equiv T^{-1} A\left(h D_{x_{1}}-i x_{1}\right) T\left(T^{-1} u_{0}\right) \equiv 0
$$

REMARK. If $p(x, \xi)=|\xi|^{2}+V(x)$, the potential $V$ satisfies (10.74), and

$$
\left\{p(x, \xi):(x, \xi) \in \mathbb{R}^{2 n}\right\} \neq \mathbb{C}
$$

then the operator $P$ still has a discrete spectrum. This follows from the proof of Theorem 6.7, once we have a point at which $P-z$ is elliptic. Such a point is produced if there exists $z$ not in the set of values of $p(x, \xi)$. However, the hypothesis of Theorem 10.20 holds in a dense open subset of the interior of the closure of the range of $p$.
EXAMPLE. It is also clear that more general operators can be considered. As a simple one-dimensional example, take

$$
P=\left(h D_{x}\right)^{2}+i h D_{x}+x^{2}
$$

with

$$
p(x, \xi)=\xi^{2}+i \xi+x^{2}, \quad\{\operatorname{Re} p, \operatorname{Im} p\}=-2 x
$$

Hence there is a quasimode corresponding to any point in the interior of the range of $p$, namely $\left\{z: \operatorname{Re} z \geq(\operatorname{Im} z)^{2}\right\}$. On the other hand, since

$$
e^{x / 2 h} P e^{-x / 2 h}=(h D)^{2}+x^{2}+\frac{1}{4},
$$

$P$ has the discrete spectrum $\{1 / 4+n h: n \in \mathbb{N}\}$. Since the spectrum lies inside an open set of quasimodes, it is unlikely to have any true physical meaning.

## Appendix A. Notation

## A. 1 BASIC NOTATION.

```
\(\mathbb{R}_{+}=(0, \infty)\)
\(\mathbb{R}^{n}=n\)-dimensional Euclidean space
\(x, y\) denote typical points in \(\mathbb{R}^{n}: x=\left(x_{1}, \ldots, x_{n}\right), y=\left(y_{1}, \ldots, y_{n}\right)\)
\(\mathbb{R}^{2 n}=\mathbb{R}^{n} \times \mathbb{R}^{n}\)
\(z=(x, \xi), w=(y, \eta)\) denote typical points in \(\mathbb{R}^{n} \times \mathbb{R}^{n}:\)
\(z=\left(x_{1}, \ldots, x_{n}, \xi_{1}, \ldots, \xi_{n}\right), w=\left(y_{1}, \ldots, y_{n}, \eta_{1}, \ldots, \eta_{n}\right)\)
\(\mathbb{T}^{n}=n\)-dimensional flat torus \(=\mathbb{R}^{n} / \mathbb{Z}^{n}\)
\(\mathbb{C}=\) complex plane
\(\mathbb{C}^{n}=\) n-dimensional complex space
\(\langle x, y\rangle=\sum_{i=1}^{n} x_{i} \bar{y}_{i}=\) inner product on \(\mathbb{C}^{n}\)
\(|x|=\langle x, x\rangle^{1 / 2}\)
\(\langle x\rangle=\left(1+|x|^{2}\right)^{1 / 2}\)
\(\mathbb{M}^{m \times n}=m \times n\)-matrices
\(\mathbb{S}^{n}=n \times n\) real symmetric matrices
\(A^{T}=\) transpose of the matrix \(A\)
```

$I$ denotes both the identity matrix and the identity mapping.
$J=\left(\begin{array}{cc}O & I \\ -I & O\end{array}\right)$
$\sigma(z, w)=\langle J z, w\rangle=$ symplectic inner product
$\# S=$ cardinality of the set $S$
$|E|=$ Lebesgue measure of the set $E \subseteq \mathbb{R}^{n}$

## A. 2 FUNCTIONS, DIFFERENTIATION.

The support of a function is denoted "spt", and a subscript " $c$ " on a space of functions means those with compact support.

- Partial derivatives:

$$
\partial_{x_{j}}:=\frac{\partial}{\partial x_{j}}, \quad D_{x_{j}}:=\frac{1}{i} \frac{\partial}{\partial x_{j}}
$$

- Multiindex notation: A multiindex is a vector $\alpha=\left(\alpha_{1}, \ldots, \alpha_{n}\right)$, the entries of which are nonnegative integers. The size of $\alpha$ is

$$
|\alpha|:=\alpha_{1}+\cdots+\alpha_{n} .
$$

We then write for $x \in \mathbb{R}^{n}$ :

$$
x^{\alpha}:=x_{1}{ }^{\alpha_{1}} \ldots x_{n}{ }^{\alpha_{n}},
$$

where $x=\left(x_{1}, \ldots, x_{n}\right)$.
Also

$$
\partial^{\alpha}:=\partial_{x_{1}}^{\alpha_{1}} \ldots \partial_{x_{n}}^{\alpha_{n}}
$$

and

$$
D^{\alpha}:=\frac{1}{i^{|\alpha|}} \partial_{x_{1}}^{\alpha_{1}} \ldots \partial_{x_{n}}^{\alpha_{n}}
$$

(WARNING: Our use of the symbols " $D$ " and " $D$ " differs from that in the PDE textbook [E].)

If $\varphi: \mathbb{R}^{n} \rightarrow \mathbb{R}$, then we write

$$
\partial \varphi:=\left(\varphi_{x_{1}}, \ldots, \varphi_{x_{n}}\right)=\text { gradient }
$$

and

$$
\partial^{2} \varphi:=\left(\begin{array}{lll}
\varphi_{x_{1} x_{1}} & \cdots & \varphi_{x_{1} x_{n}} \\
& \ddots & \\
\varphi_{x_{n} x_{1}} & \cdots & \varphi_{x_{n} x_{n}}
\end{array}\right)=\text { Hessian matrix }
$$

Also

$$
D \varphi:=\frac{1}{i} \partial \varphi .
$$

If $\varphi$ depends on both the variables $x, y \in \mathbb{R}^{n}$, we put

$$
\partial_{x}^{2} \varphi:=\left(\begin{array}{ccc}
\varphi_{x_{1} x_{1}} & \ldots & \varphi_{x_{1} x_{n}} \\
& \ddots & \\
\varphi_{x_{n} x_{1}} & \ldots & \varphi_{x_{n} x_{n}}
\end{array}\right)
$$

and

$$
\partial_{x, y}^{2} \varphi:=\left(\begin{array}{ccc}
\varphi_{x_{1} y_{1}} & \ldots & \varphi_{x_{1} y_{n}} \\
& \ddots & \\
\varphi_{x_{n} y_{1}} & \cdots & \varphi_{x_{n} y_{n}}
\end{array}\right)
$$

- Jacobians: Let

$$
x \mapsto y=y(x)
$$

be a diffeomorphism, $y=\left(y^{1}, \ldots, y^{n}\right)$. The Jacobian matrix is

$$
\partial y=\partial_{x} y:=\left(\begin{array}{ccc}
\frac{\partial y^{1}}{\partial x_{1}} & \cdots & \frac{\partial y^{1}}{\partial x_{n}} \\
\frac{\partial y^{n}}{\partial x_{1}} & \cdots & \frac{\partial y^{n}}{\partial x_{n}}
\end{array}\right)_{n \times n} .
$$

- Poisson bracket: If $f, g: \mathbb{R}^{n} \rightarrow \mathbb{R}$ are $C^{1}$ functions,

$$
\{f, g\}:=\left\langle\partial_{\xi} f, \partial_{x} g\right\rangle-\left\langle\partial_{x} f, \partial_{\xi} g\right\rangle=\sum_{j=1}^{n} \frac{\partial f}{\partial \xi_{j}} \frac{\partial g}{\partial x_{j}}-\frac{\partial f}{\partial x_{j}} \frac{\partial g}{\partial \xi_{j}}
$$

- The Schwartz space is

$$
\begin{aligned}
\mathcal{S}=\mathcal{S}\left(\mathbb{R}^{n}\right) & := \\
& \left\{\varphi \in C^{\infty}\left(\mathbb{R}^{n}\right)\left|\sup _{\mathbb{R}^{n}}\right| x^{\alpha} \partial^{\beta} \varphi \mid<\infty \text { for all multiindices } \alpha, \beta\right\} .
\end{aligned}
$$

We say

$$
\varphi_{j} \rightarrow \varphi \quad \text { in } \mathcal{S}
$$

provided

$$
\sup _{\mathbb{R}^{n}}\left|x^{\alpha} D^{\beta}\left(\varphi_{j}-\varphi\right)\right| \rightarrow 0
$$

for all multiindices $\alpha, \beta$
We write $\mathcal{S}^{\prime}=\mathcal{S}^{\prime}\left(\mathbb{R}^{n}\right)$ for the space of tempered distributions, which is the dual of $\mathcal{S}=\mathcal{S}\left(\mathbb{R}^{n}\right)$. That is, $u \in \mathcal{S}^{\prime}$ provided $u: \mathcal{S} \rightarrow \mathbb{C}$ is linear and $\varphi_{j} \rightarrow \varphi$ in $\mathcal{S}$ implies $u\left(\varphi_{j}\right) \rightarrow u(\varphi)$.

We say

$$
u_{j} \rightarrow u \quad \text { in } \mathcal{S}^{\prime}
$$

provided

$$
u_{j}(\varphi) \rightarrow u(\varphi) \quad \text { for all } \varphi \in \mathcal{S}
$$

## A. 3 ELEMENTARY OPERATORS.

Multiplication operator: $M_{\lambda} f(x)=\lambda f(x)$
Translation operator: $T_{\xi} f(x)=f(x-\xi)$
Reflection operator: $R f(x):=f(-x)$

## A. 4 OPERATORS.

$A^{*}=$ adjoint of the operator $A$
$[A, B]=A B-B A=$ commutator of A and B
$\sigma(A)=$ symbol of the pseudodifferential operator A
$\operatorname{spec}(A)=$ spectrum of A .
$\operatorname{tr}(A)=$ trace of A .
We say that the operator $B$ is of trace class if

$$
\operatorname{tr}(B):=\sum \sqrt{\lambda_{j}}<\infty
$$

where the $\lambda_{j} \geq 0$ are the eigenvalues of the symmetric matrix $B^{*} B$.

- If $A: X \rightarrow Y$ is a bounded linear operator, we define the operator norm

$$
\|A\|:=\sup \left\{\|A u\|_{Y} \mid\|u\|_{X} \leq 1\right\}
$$

We will often write this norm as

$$
\|A\|_{X \rightarrow Y}
$$

when we want to emphasize the spaces between which $A$ maps.
The space of bounded linear operators from $X$ to $Y$ is denoted by $L(X, Y)$; and the space of bounded linear operators from $X$ to itself is denoted $L(X)$.

## A. 5 ESTIMATES.

- We write

$$
f=O\left(h^{\infty}\right) \quad \text { as } h \rightarrow 0
$$

if for each positive integer $N$ there exists a constant $C_{N}$ such that

$$
|f| \leq C_{N} h^{N} \quad \text { for all } 0<h \leq 1
$$

- If we want to specify boundedness in the space $X$, we write

$$
f=O_{X}\left(h^{N}\right)
$$

to mean

$$
\|f\|_{X}=O\left(h^{N}\right)
$$

- If $A$ is a bounded linear operator between the spaces $X, Y$, we will often write

$$
A=O_{X \rightarrow Y}\left(h^{N}\right)
$$

to mean

$$
\|A\|_{X \rightarrow Y}=O\left(h^{N}\right)
$$

## A.6. SYMBOL CLASSES.

We record from Chapter 4 the various definitions of classes for symbols $a=a(x, \xi, h)$.

- Given an order function $m$ on $\mathbb{R}^{2 n}$, we define the corresponding class of symbols:

$$
\begin{aligned}
S(m):= & \left\{a \in C^{\infty} \mid \text { for each multiindex } \alpha\right. \\
& \text { there exists a constant } \left.C_{\alpha} \text { so that }\left|\partial^{\alpha} a\right| \leq C_{\alpha} m\right\} .
\end{aligned}
$$

- We as well define

$$
S^{k}(m):=\left\{a \in C^{\infty}| | \partial^{\alpha} a \mid \leq C_{\alpha} h^{-k} m \text { for all multiindices } \alpha\right\}
$$

and

$$
S_{\delta}^{k}(m):=\left\{a \in C^{\infty}| | \partial^{\alpha} a \mid \leq C_{\alpha} h^{-\delta|\alpha|-k} m \text { for all multiindices } \alpha\right\} .
$$

The index $k$ indicates how singular is the symbol $a$ as $h \rightarrow 0$; the index $\delta$ allows for increasing singularity of the derivatives of $a$.

- Write also

$$
S^{-\infty}(m):=\left\{a \in C^{\infty} \mid \text { for each } \alpha \text { and } N,\left|\partial^{\alpha} a\right| \leq C_{\alpha, N} h^{N} m\right\}
$$

So if $a$ is a symbol in $S^{-\infty}(m)$, then $a$ and all of its derivatives are $O\left(h^{\infty}\right)$ as $h \rightarrow 0$.

- If the order function is the constant function $m \equiv 1$, we will usually not write it:

$$
S^{k}:=S^{k}(1), S_{\delta}^{k}=S_{\delta}^{k}(1)
$$

- We will also omit zero superscripts:

$$
\begin{aligned}
S:=S^{0}= & S^{0}(1) \\
& =\left\{a \in C^{\infty}\left(\mathbb{R}^{2 n}\right)| | \partial^{\alpha} a \mid \leq C_{\alpha} \text { for all multiindices } \alpha\right\} .
\end{aligned}
$$

- We will sometimes write

$$
a=O_{S}\left(h^{N}\right),
$$

to mean that for all $\alpha$

$$
\left|\partial^{\alpha} a\right| \leq C_{\alpha} h^{N} .
$$

We use similar notation for other spaces with seminorms.

## A. 7 PSEUDODIFFERENTIAL OPERATORS.

We cross reference the following terminology from Appendix E. Let $M$ denote a manifold.

- A linear operator $A: C^{\infty}(M) \rightarrow C^{\infty}(M)$ is called a pseudodifferential operator if there exist integers $m, k$ such that for each coordinate patch $U_{\gamma}$, and there exists a symbol $a_{\gamma} \in S^{m, k}$ such that for any $\varphi, \psi \in C_{\mathrm{c}}^{\infty}\left(U_{\gamma}\right)$

$$
\varphi A(\psi u)=\varphi \gamma^{*} a_{\gamma}^{\mathrm{w}}(x, h D)\left(\gamma^{-1}\right)^{*}(\psi u)
$$

for each $u \in C^{\infty}(M)$.

- We write

$$
A \in \Psi^{m, k}(M)
$$

and also put

$$
\Psi^{k}(M):=\Psi^{0, k}(M), \quad \Psi(M):=\Psi^{0,0}(M)
$$

## Appendix B. Differential forms

In this section we provide a minimalist review of differential forms on $\mathbb{R}^{N}$. For more a detailed and fully rigorou)description of differential forms on manifolds we refer to [W, Chapter 2].

## NOTATION.

(i) If $x=\left(x_{1}, \ldots, x_{n}\right), \xi=\left(\xi_{1}, \ldots, \xi_{n}\right)$, then $d x_{j}, d \xi_{j} \in\left(\mathbb{R}^{2 n}\right)^{*}$ satisfy

$$
\begin{aligned}
d x_{j}(u) & =d x_{j}(x, \xi)=x_{j} \\
d \xi_{j}(u) & =d \xi_{j}(x, \xi)=\xi_{j} .
\end{aligned}
$$

(ii) If $\alpha, \beta \in\left(\mathbb{R}^{2 n}\right)^{*}$, then

$$
(\alpha \wedge \beta)(u, v):=\alpha(u) \beta(v)-\alpha(v) \beta(u)
$$

for $u, v \in \mathbb{R}^{2 n}$. More generally, for $\alpha_{j} \in\left(\mathbb{R}^{2 n}\right)^{*}, j=1, \cdots, m \leq 2 n$, and $u=\left(u_{1}, \cdots, u_{m}\right)$, an $m$-tuple of $u_{k} \in \mathbb{R}^{2 n}$,

$$
\begin{equation*}
\left(\alpha_{1} \wedge \cdots \wedge \alpha_{m}\right)(u)=\operatorname{det}\left(\left[\alpha_{j}\left(u_{k}\right)\right]_{1 \leq j, k \leq 2 n}\right) \tag{B.1}
\end{equation*}
$$

(iii) If $f: \mathbb{R}^{n} \rightarrow \mathbb{R}$, the differential of $f$, is the 1 -form

$$
d f=\sum_{j=1}^{n} \frac{\partial f}{\partial x_{i}} d x_{i} .
$$

(iv) An $m$-form on $\mathbb{R}^{n}$ is given by

$$
w=\sum_{i_{1}<i_{2}<\cdots<i_{m}} f_{i_{1} \cdots i_{m}}(x) d x_{i_{1}} \wedge \cdots d x_{i_{m}}, \quad f_{i_{1} \cdots i_{m}} \in C^{\infty}\left(\mathbb{R}^{n}\right) .
$$

Its action at $x$ on $m$-tuples of vectors is given using (ii).
(v) The differential of $m$-form is defined by induction using (iii) and $d(f g)=d f \wedge g+f d g$, where $f$ is a function and $g$ is an $(m-1)$-form. It satisfies $d^{2}=0$.

THEOREM B. 1 (Alternative definition of $\boldsymbol{d}$ ). Suppose $w$ is a differential 2-form, and $u \in C^{\infty}\left(\mathbb{R}^{n}, \mathbb{R}^{3}\right)$, $u=\left(u_{1}, u_{3}, u_{3}\right)$ is a 3-tuple of vectorfields. Then

$$
\begin{align*}
d w(u)= & u_{1}\left(w\left(u_{2}, u_{3}\right)\right)+u_{2}\left(w\left(u_{3}, u_{1}\right)\right)+u_{3}\left(w\left(u_{1}, u_{2}\right)\right) \\
& -w\left(\left[u_{1}, u_{2}\right], u_{3}\right)-w\left(\left[u_{2}, u_{3}\right], u_{1}\right)-w\left(\left[u_{3}, u_{1}\right], u_{2}\right) . \tag{B.2}
\end{align*}
$$

1. Both sides of (B.2) are linear in $w$ and trilinear in $u$.
2. When, say, $u_{1}$ is multiplied by $f \in C^{\infty}\left(\mathbb{R}^{n}\right)$, then

$$
d w\left(f u_{1}, u_{2}, u_{3}\right)=f d w(u)
$$

and the right hand side of (B.2) is equal to

$$
\begin{aligned}
& f u_{1}\left(w\left(u_{2}, u_{3}\right)\right)+u_{2}\left(f w\left(u_{3}, u_{1}\right)\right)+u_{3}\left(f w\left(u_{1}, u_{2}\right)\right) \\
& -w\left(\left[f u_{1}, u_{2}\right], u_{3}\right)-w\left(\left[u_{2}, u_{3}\right], f u_{1}\right)-w\left(\left[u_{3}, f u_{1}\right], u_{2}\right),
\end{aligned}
$$

and this is equal to the right hand side of (B.2) multiplied by $f$. In fact,

$$
\left[f u_{1}, u_{2}\right]=f\left[u_{1}, u_{2}\right]-\left(u_{2} f\right) u_{1}, \quad\left[u_{3}, f u_{1}\right]=f\left[u_{3}, u_{1}\right]+\left(u_{3} f\right) u_{1},
$$

and

$$
\begin{aligned}
& u_{2}\left(f w\left(u_{3}, u_{1}\right)\right)=f u_{2}\left(w\left(u_{3}, u_{1}\right)\right)+\left(u_{2} f\right) w\left(u_{3}, u_{1}\right), \\
& u_{3}\left(f w\left(u_{1}, u_{2}\right)\right)=f u_{3}\left(w\left(u_{1}, u_{2}\right)\right)+\left(u_{3} f\right) w\left(u_{1}, u_{2}\right) .
\end{aligned}
$$

3. Hence we only need to check this identity for $u$ constant and for $w=w_{1} d w_{2} \wedge d w_{3}$, where $w_{1} \in C^{\infty}$, and $w_{2}, w_{3}$ are coordinate functions (that is are among $x_{1}, \cdots x_{n}$ ). Then

$$
d w(u)=\operatorname{det}\left(\left[u_{j} w_{i}\right]_{1 \leq i, j \leq n}\right),
$$

and the right hand side of (B.2) is given by (remember that now $u_{j} w_{i}$, $i=2,3$ are constants) by the expansion of this determinant with respect to the first row, $\left(u_{1} w_{1}, u_{2} w_{1}, u_{3} w_{1}\right)$.

DEFINITION. If $\eta$ is a differential m-form and $V$ a vector field, then the contraction of $\eta$ by $V$, denoted

$$
V\lrcorner \eta,
$$

is the $(m-1)$-form defined by

$$
(V\lrcorner \eta)(u)=\eta(V, u),
$$

where $u$ is an $(m-1)$-tuple of vectorfields. We use the consistent convention that for 0 -forms, that is for functions, $V\lrcorner f=0$.

We note the following property of contraction which can be deduced from (B.1): if $v$ is a $k$-form and $w$ is an $m$-form, then

$$
\begin{equation*}
\left.V\lrcorner(v \wedge w)=(V\lrcorner v) \wedge w+(-1)^{k} v \wedge(V\lrcorner w\right) . \tag{B.3}
\end{equation*}
$$

DEFINITIONS. Let $\kappa: \mathbb{R}^{n} \rightarrow \mathbb{R}^{n}$ be a smooth mapping.
(i) If $V$ is a vector field on $\mathbb{R}^{n}$, the push-forward is

$$
\kappa_{*} V=\partial \kappa(V)
$$

(ii) If $\eta$ is a 1 -form on $\mathbb{R}^{n}$, the pull-back is

$$
\left(\kappa^{*} \eta\right)(u)=\eta\left(\kappa_{*} u\right) .
$$

THEOREM B. 2 (Differentials and pull-backs). Let $w$ be a differential m-form. We have

$$
\begin{equation*}
d\left(\kappa^{*} w\right)=\kappa^{*}(d w) . \tag{B.4}
\end{equation*}
$$

Proof. 1. We first prove this for functions: $d\left(\kappa^{*} f\right)=d(\kappa(f))=$ $\sum_{j=1}^{n} \frac{\partial y_{i}}{\partial x_{j}} \frac{\partial f}{\partial y_{i}} d x_{j}$. Furthermore,

$$
\kappa^{*}(d f)=\kappa^{*}\left(\sum_{i=1}^{n} \frac{\partial f}{\partial y_{i}} d y_{i}\right)=\sum_{i=1}^{n} \frac{\partial f}{\partial y_{i}} \kappa^{*}\left(d y_{i}\right) .
$$

2. The proof now follows by induction on the order of the differential form: any $m$-form can be written as a linear combination of forms $f d g$ where $f$ is a function, and $g$ is $(m-1)$-form.

DEFINITION. If $V$ is a vector field generating the flow $\varphi_{t}$, then the Lie derivative of $w$ is

$$
\mathcal{L}_{V} w:=\left.\frac{d}{d t}\left(\left(\varphi_{t}\right)^{*} w\right)\right|_{t=0}
$$

Here $w$ denotes a function, a vector field or a form. We recall that $\varphi_{t}$ is generated by a time independent vectorfield, $V, \varphi_{t}=\exp (t V)$, means that

$$
(d / d t) \varphi_{t}(m)=V\left(\varphi_{t}(m)\right), \quad \varphi_{0}(m)=m .
$$

EXAMPLE S. (i) If $f$ is a function,

$$
\mathcal{L}_{V} f=V(f) .
$$

(ii) If $W$ is a vector field

$$
\mathcal{L}_{V} W=[V, W] .
$$

Since for differential forms, $w, d\left(\phi_{t}\right)^{*} w=\phi_{t}^{*}(d w)$, we see that $\mathcal{L}_{V}$ commutes with $d$ :

$$
\begin{equation*}
d\left(\mathcal{L}_{V} w\right)=\mathcal{L}_{V}(d w) . \tag{B.5}
\end{equation*}
$$

We also note that $\mathcal{L}_{V}$ is a derivation: for a function $f \in C^{\infty}$ and a differential form $w$,

$$
\begin{equation*}
\mathcal{L}_{V}(f w)=\left(\mathcal{L}_{V} f\right) w+f \mathcal{L}_{V} w \tag{B.6}
\end{equation*}
$$

THEOREM B. 3 (Cartan's formula). If $w$ is a differential form,

$$
\begin{equation*}
\left.\left.\mathcal{L}_{V} w=d(V\lrcorner w\right)+(V\lrcorner d w\right) . \tag{B.7}
\end{equation*}
$$

Proof. 1. We proceed by induction on the order of differential forms. For 0-forms, that is for functions, we have

$$
\left.\left.\left.\mathcal{L}_{V} f=V f=V\right\lrcorner d f=d(V\lrcorner f\right)+(V\lrcorner d f\right),
$$

since by our convention $V\lrcorner f=0$.
2. Any $m$-form is a linear combinations of forms $f d g$ where $f$ is a function and $g$ in an ( $m-1$ )-form. Then, using (B.5), (B.6), $d^{2}=0$, and the induction hypothesis,

$$
\begin{align*}
\mathcal{L}_{V}(f d g) & =\left(\mathcal{L}_{V} f\right) d g+f \mathcal{L}_{V} d g \\
& =(V f) d g+f d\left(\mathcal{L}_{V} g\right) \\
& =(V f) d g+f d(d(V\lrcorner g)+V\lrcorner d g)  \tag{B.8}\\
& =(V f) d g+f d(V\lrcorner d g) .
\end{align*}
$$

3. The right hand side of (B.7) for $w=f d g$ is equal to

$$
\begin{align*}
& d(V\lrcorner(f d g))+V\lrcorner(d(f d g))= \\
& \quad f(V\lrcorner d g)+d f \wedge(V\lrcorner d g)+V\lrcorner(d f \wedge d g) \tag{B.9}
\end{align*}
$$

Now we can use (B.3) with $w=d f, v=d g, k=1$, to obtain

$$
V\lrcorner(d f \wedge d g)=(V f) d g-d f \wedge(V\lrcorner d g) .
$$

Inserting this in (B.9) and the comparison with (B.8) gives (B.7) for $w=f d g$ and hence for all differential $m$-forms.

THEOREM B. 4 (Poincaré's Lemma). If $\alpha$ is a $k$-form defined in the open ball $U=B^{0}(0, R)$ and if

$$
d \alpha=0,
$$

then there exists a $(k-1)$ form $\omega$ in $U$ such that

$$
d \omega=\alpha
$$

Proof. 1. Let $\Omega^{k}(U)$ denote the space of $k$-forms on $U$. We will build a linear mapping

$$
H: \Omega^{k}(U) \rightarrow \Omega^{k-1}(U)
$$

such that

$$
\begin{equation*}
d \circ H+H \circ d=I \tag{B.10}
\end{equation*}
$$

Then

$$
d(H \alpha)+H d \alpha=\alpha
$$

and so $d \omega=\alpha$ for $\omega:=H \alpha$.
2. Define $A: \Omega^{k}(U) \rightarrow \Omega^{k}(U)$ by

$$
A\left(f d x_{i_{1}} \wedge \cdots \wedge d x_{i_{k}}\right)=\left(\int_{0}^{1} t^{k-1} f(t p) d t\right) d x_{i_{1}} \wedge \cdots \wedge d x_{i_{k}}
$$

Set

$$
X:=\left\langle x, \partial_{x}\right\rangle=\sum_{j=1}^{n} x_{j} \frac{\partial}{\partial x_{j}} .
$$

We claim

$$
\begin{equation*}
A \mathcal{L}_{X}=I \quad \text { on } \Omega^{k}(U) . \tag{B.11}
\end{equation*}
$$

and

$$
\begin{equation*}
d \circ A=A \circ d \tag{B.12}
\end{equation*}
$$

Assuming these assertions, define

$$
H:=A \circ X\lrcorner .
$$

By Cartan's formula, Theorem B.3,

$$
\left.\left.\mathcal{L}_{X}=d \circ(X\lrcorner\right)+X\right\lrcorner \circ d .
$$

Thus

$$
\begin{aligned}
I=A \mathcal{L}_{X} & =A \circ d \circ(X\lrcorner)+A \circ X\lrcorner \circ d \\
& =d(A \circ X\lrcorner)+(A \circ X\lrcorner) \circ d \\
& =d \circ H+H \circ d ;
\end{aligned}
$$

and this proves (B.10).
3. To prove (B.11), we compute

$$
\begin{aligned}
A \mathcal{L}_{X}\left(f d x_{i_{1}}\right. & \left.\wedge \cdots \wedge d x_{i_{2}}\right) \\
= & A\left[\left(k f+\sum_{j=1}^{n} x_{j} \frac{\partial f}{\partial x_{j}}\right)\left(d x_{i_{1}} \wedge \cdots \wedge d x_{i_{k}}\right)\right] \\
= & \int_{0}^{1} k t^{k-1} f(t p)+\sum_{j=1}^{n} t^{k-1} x_{j} \frac{\partial f}{\partial x_{j}}(t p) \\
& \quad d t d x_{i_{1}} \wedge \cdots \wedge d x_{i_{k}} \\
= & \int_{0}^{1} \frac{d}{d t}\left(t^{k} f(t p)\right) d t d x_{i_{1}} \wedge \cdots \wedge d x_{i_{k}} \\
= & f d x_{i_{1}} \wedge \cdots \wedge d x_{i_{k}} .
\end{aligned}
$$

4. To verify (B.12), note

$$
\begin{aligned}
A \circ d\left(f d x_{i_{1}}\right. & \left.\wedge \cdots \wedge d x_{i_{k}}\right) \\
& =A\left(\sum_{j=1}^{n} \frac{\partial f}{\partial x_{j}} d x_{j} \wedge d x_{i_{1}} \wedge \cdots \wedge d x_{i_{k}}\right) \\
& =\left(\int_{0}^{1} t^{k-1} \sum_{j=1}^{n} \frac{\partial f}{\partial x_{j}}(t p) d x_{j} d t\right) d x_{i_{1}} \wedge \cdots \wedge d x_{i_{k}} \\
& =d\left(\left(\int_{0}^{1} t^{k-1} f(t p) d t\right) d x_{i_{1}} \wedge \cdots \wedge d x_{i_{k}}\right) \\
& =d \circ A\left(f d x_{i_{1}} \wedge \cdots \wedge d x_{i_{k}}\right) .
\end{aligned}
$$

## Appendix C. Functional analysis

THEOREM C. 1 (Schwartz Kernel Theorem). Let $A: \mathcal{S} \rightarrow \mathcal{S}^{\prime}$ be a continuous linear operator.

Then there exists a distribution $K_{A} \in \mathcal{S}^{\prime}\left(\mathbb{R}^{n} \times \mathbb{R}^{n}\right)$ such that

$$
\begin{equation*}
A u(x)=\int_{\mathbb{R}^{n}} K_{A}(x, y) u(y) d y \tag{C.1}
\end{equation*}
$$

for all $u \in \mathcal{S}$.
We call $K_{A}$ the kernel of $A$.

THEOREM C. 2 (Inverse Function Theorem). Let $X, Y$ denote Banach spaces and assume

$$
f: X \rightarrow Y
$$

is $C^{1}$. Select a point $x_{0} \in X$ and write $y_{0}:=f\left(x_{0}\right)$.
(i) (Right inverse) If there exists $A \in L(Y, X)$ such that

$$
\partial f\left(x_{0}\right) A=I,
$$

then there exists $g \in C^{1}(Y, X)$ such that

$$
f \circ g=I \quad \text { near } y_{0} .
$$

(ii) (Left inverse) If there exists $B \in L(Y, X)$ such that

$$
B \partial f\left(x_{0}\right)=I
$$

then there exists $g \in C^{1}(Y, X)$ such that

$$
g \circ f=I \quad \text { near } x_{0} .
$$

THEOREM C. 3 (Approximate inverses). Let $X, Y$ be Banach spaces and suppose $A: X \rightarrow Y$ is a bounded linear operator. Suppose there exist bounded linear operators $B_{1}, B_{2}: Y \rightarrow X$ such that

$$
\begin{cases}A B_{1}=I+R_{1} & \text { on } Y  \tag{C.2}\\ B_{2} A=I+R_{2} & \text { on } X\end{cases}
$$

where

$$
\left\|R_{1}\right\|<1,\left\|R_{2}\right\|<1
$$

Then $A$ is invertible.

Proof. The operator $I+R_{1}$ is invertible, with

$$
\left(I+R_{1}\right)^{-1}=\sum_{k=0}^{\infty}(-1)^{k} R_{1}^{k},
$$

this series converging since $\left\|R_{1}\right\|<1$. Hence

$$
A C_{1}=I \quad \text { for } C_{1}:=B_{1}\left(I+R_{1}\right)^{-1} .
$$

Likewise,

$$
\left(I+R_{2}\right)^{-1}=\sum_{k=0}^{\infty}(-1)^{k} R_{2}^{k}
$$

and

$$
C_{2} A=I \quad \text { for } C_{2}:=\left(I+R_{2}\right)^{-1} B_{2} .
$$

So $A$ has a left and a right inverse, and is consequently invertible, with $A^{-1}=C_{1}=C_{2}$.

THEOREM C. 4 (Norms of powers of operators). Let $A \in L\left(H_{1}, H_{2}\right)$, where $H_{1}, H_{2}$ are Hilbert spaces.
(i) Then

$$
\|A\|=\sup _{\|u\|,\|v\|=1}|\langle A u, v\rangle|,\|A\|=\left\|A^{*}\right\|,\|A\|^{2}=\left\|A^{*} A\right\| .
$$

(ii) If $A$ is self-adjoint, $\|A\|^{m}=\left\|A^{m}\right\|$ for all $m \in \mathbb{N}$.

Proof. 1. We may assume $\|A\|>0$. Note that

$$
|\langle A u, v\rangle| \leq\|A u\|\|v\| \leq\|A\|\|u\|\|v\|=\|A\|
$$

for any two unit vectors $u, v$. Thus $\sup _{\|u\|,\|v\|=1}|\langle A u, v\rangle| \leq\|A\|$.
Now if $u \notin \operatorname{ker}(A)$, we can put $v=A u /\|A u\|$. Consequently,

$$
\sup _{\|u\|\| \| \|=1}|\langle A u, v\rangle| \geq \sup _{\substack{u \notin \operatorname{ker}(\mathcal{A}) \\\|u\|=1}} \frac{1}{\|A u\|}|\langle A u, A u\rangle|=\sup _{\substack{u \notin \operatorname{ker}(A) \\\|u\|=1}}\|A u\|=\|A\| ;
$$

and therefore

$$
\begin{aligned}
\|A\|^{2} & =\sup _{\|u\|=1}\|A u\|^{2}=\sup _{\|u\|=1}\left|\left\langle A^{*} A u, u\right\rangle\right| \\
& \leq \sup _{\|u\|,\|v\|=1}\left|\left\langle A^{*} A u, v\right\rangle\right|=\left\|A^{*} A\right\| .
\end{aligned}
$$

Also, for any $u, v$ with norm one we have

$$
\left|\left\langle A^{*} A u, v\right\rangle\right|=|\langle A u, A v\rangle| \leq\|A u\|\|A v\| \leq\|A\|^{2} .
$$

Taking the supremum over $u, v$ gives us the inequality $\left\|A^{*} A\right\| \leq\|A\|^{2}$.
2. A simple induction now yields $\|A\|^{2^{k}}=\left\|A^{2^{k}}\right\|$ for all natural numbers $k$. For a general $m$, find an $n$ such that $m+n$ is a power of 2. Then $\left\|A^{n+m}\right\|=\|A\|^{n+m}$, and so

$$
\|A\|^{m}\|A\|^{n}=\|A\|^{m+n}=\left\|A^{m+n}\right\|=\left\|A^{m} A^{n}\right\| \leq\|A\|^{m}\|A\|^{n} .
$$

Therefore the inequality signs above must be equalities, and this implies $\left\|A^{m}\right\|=\|A\|^{m}$.

THEOREM C. 5 (Cotlar-Stein Theorem). Let $H_{1}, H_{2}$ be Hilbert spaces and $A_{j} \in L\left(H_{1}, H_{2}\right)$ for $j=1, \ldots$ Assume

$$
\sup _{j} \sum_{k=1}^{\infty}\left\|A_{j}^{*} A_{k}\right\|^{1 / 2} \leq C, \quad \sup _{j} \sum_{k=1}^{\infty}\left\|A_{j} A_{k}^{*}\right\|^{1 / 2} \leq C .
$$

Then the series $A:=\sum_{j=1}^{\infty} A_{j}$ converges in strong operator topology and

$$
\|A\| \leq C
$$

Proof. 1. Let us first assume that $A_{j}=0$ for $j>J$ so that $A$ is well defined. Since $A^{*} A$ is self-adjoint, the previous theorem implies

$$
\|A\|^{2 m}=\left\|\left(A^{*} A\right)^{m}\right\| .
$$

In addition,

$$
\left(A^{*} A\right)^{m}=\sum_{j_{1}, \ldots, j_{2 m}=1}^{\infty} A_{j_{1}}^{*} A_{j_{2}} \ldots A_{j_{2 m-1}}^{*} A_{j_{2 m}}=: \sum_{j_{1}, \ldots, j_{2 m}} a_{j_{1}, \ldots, j_{2 m}} .
$$

Now

$$
\left\|a_{j_{1}, \ldots, j_{2 m}}\right\| \leq\left\|A_{j_{1}}^{*} A_{j_{2}}\right\|\left\|A_{j_{3}}^{*} A_{j_{4}}\right\| \ldots\left\|A_{j_{2 m-1}}^{*} A_{j_{2 m}}\right\|,
$$

and also

$$
\left\|a_{j_{1}, \ldots, j_{2 m}}\right\| \leq\left\|A_{j_{1}}\right\|\left\|A_{j_{2}} A_{j_{3}}^{*}\right\| \ldots\left\|A_{j_{2 m-2}} A_{j_{2 m-1}}^{*}\right\|\left\|A_{j_{2 m}}\right\| .
$$

Multiply these estimates and take square roots:

$$
\left\|a_{j_{1}, \ldots, j_{2 m}}\right\| \leq C\left\|A_{j_{1}}^{*} A_{j_{2}}\right\|^{1 / 2}\left\|A_{j_{2}} A_{j_{m}}^{*}\right\|^{1 / 2} \ldots\left\|A_{j_{2 m-1}}^{*} A_{j_{2 m}}\right\|^{1 / 2}
$$

Consequently,

$$
\begin{aligned}
\|A\|^{2 m} & =\left\|\left(A^{*} A\right)^{m}\right\| \leq \sum_{j_{1}, \ldots, j_{2 m}=1}^{\infty}\left\|a_{j_{1}, \ldots, j_{2 m}}\right\| \\
& \leq C \sum_{j_{1}, \ldots, j_{2 m}=1}^{\infty}\left\|A_{j_{1}} A_{j_{2}}^{*}\right\|^{1 / 2} \ldots\left\|A_{j_{2 m-1}}^{*} A_{j_{2 m}}\right\|^{1 / 2} \\
& \leq J C C^{2 m}
\end{aligned}
$$

where the $J$ factor came from having $2 m$ sums and only $2 m-1$ factors in the summands.

Hence

$$
\|A\| \leq J^{\frac{1}{2 m}} C^{\frac{2 m+1}{2 m}} \rightarrow C \quad \text { as } m \rightarrow \infty
$$

2. To consider the general case, take $u \in E$, and suppose $u=A_{k}^{*} v$ for some $k$. Then

$$
\begin{aligned}
\left\|\sum_{j=1}^{\infty} A_{j} u\right\| & =\left\|\sum_{j=1}^{\infty} A_{j} A_{k}^{*} v\right\| \\
& \leq \sum_{j=1}^{\infty}\left\|A_{j} A_{k}^{*}\right\|^{1 / 2}\left\|A_{j} A_{k}^{*}\right\|^{1 / 2}\|v\| \\
& \leq C^{2}\|v\| .
\end{aligned}
$$

Thus $\sum_{j=1}^{\infty} A_{j} u$ converges for $u \in \Sigma:=\operatorname{span}\left\{A_{k}^{*}(E) \mid k=1, \ldots, n\right\}$ and so also for $u \in \bar{\Sigma}$. If $u$ is orthogonal to $\bar{\Sigma}$, then $u \in \operatorname{ker}\left(A_{k}\right)$ for all $k$; in which case $\sum_{j=1}^{\infty} A_{j} u=0$.

Henceforth $H$ denotes a complex Hilbert space, with inner product $\langle\cdot, \cdot\rangle$.

THEOREM C. 6 (Spectrum of self-adjoint operators). Suppose $A: H \rightarrow H$ is a bounded self-adjoint operator.
(i) Then $(A-\lambda)^{-1}$ exists and is a bounded linear operator on $H$ for $\lambda \in \mathbb{C}-\operatorname{spec}(A)$, where $\operatorname{spec}(A) \subset \mathbb{R}$ is the spectrum of $A$.
(ii) If $\operatorname{spec}(A) \subseteq[a, \infty)$, then

$$
\begin{equation*}
\langle A u, u\rangle \geq a\|u\|^{2} \quad(u \in A) . \tag{C.3}
\end{equation*}
$$

THEOREM C. 7 (Maximin and minimax principles). Suppose that $A: H \rightarrow H$ is self-adjoint and semibounded, meaning $A \geq-c_{0}$. Assume also that $\left(A+2 c_{0}\right)^{-1}: H \rightarrow H$ is a compact operator.

Then the spectrum of $A$ is discrete: $\lambda_{1} \leq \lambda_{2} \leq \lambda_{3} \cdots$; and furthermore
(i)

$$
\begin{equation*}
\lambda_{j}=\max _{\substack{V \subset H \\ \operatorname{codim} V<j}} \min _{\substack{v \in V \\ v \neq 0}} \frac{\langle A v, v\rangle}{\|v\|^{2}} \tag{C.4}
\end{equation*}
$$

(ii)

$$
\begin{equation*}
\lambda_{j}=\min _{\substack{V \subset H \\ \operatorname{dim} V \leq j}} \max _{\substack{v \in V \\ v \neq 0}} \frac{\langle A v, v\rangle}{\|v\|^{2}} \tag{C.5}
\end{equation*}
$$

In these formulas, $V$ denotes a linear subspace of $H$.

DEFINITIONS. (i) Let $Q: H \rightarrow H$ be a bounded linear operator. We define the rank of $Q$ to be the dimension of the range $Q(H)$.
(ii)If $A$ is an operator with real and discrete spectrum, we set

$$
N(\lambda):=\#\left\{\lambda_{j} \mid \lambda_{j} \leq \lambda\right\}
$$

to count the number of eigenvalues less than or equal to $\lambda$.
THEOREM C. 8 (Estimating $\boldsymbol{N}(\boldsymbol{\lambda})$ ). Let $A$ satisfy the assumptions of Theorem C.7.
(i) If

$$
\left\{\begin{array}{l}
\text { for each } \delta>0, \text { there exists an operator } Q  \tag{C.6}\\
\text { with rank } Q \leq k, \text { such that } \\
\langle A u, u\rangle \geq(\lambda-\delta)\|u\|^{2}-\langle Q u, u\rangle \text { for } u \in H
\end{array}\right.
$$

then

$$
N(\lambda) \leq k .
$$

(ii) If

$$
\left\{\begin{array}{l}
\text { for each } \delta>0, \text { there exists a subspace } V  \tag{C.7}\\
\text { with } \operatorname{dim} V \geq k, \text { such that } \\
\langle A u, u\rangle \leq(\lambda+\delta)\|u\|^{2} \text { for } u \in V
\end{array}\right.
$$

then

$$
N(\lambda) \geq k
$$

Proof. 1. Set $W:=Q(H)^{T}$. Thus codim $W=\operatorname{rank} \mathrm{Q} \leq k$. Therefore the maximin formula (C.4) implies

$$
\begin{aligned}
\lambda_{k} & =\max _{\substack{V \subseteq H \\
\operatorname{codim} V<k}} \min _{\substack{v \in V \\
v \neq 0}} \frac{\langle A v, v\rangle}{\|v\|^{2}} \geq \min _{\substack{v \in W \\
v \neq 0}} \frac{\langle A v, v\rangle}{\|v\|^{2}} \\
& =\min _{\substack{v \in W \\
v \neq 0}}\left(\lambda-\delta-\frac{\langle Q v, v\rangle}{\|v\|^{2}}\right)=\lambda-\delta,
\end{aligned}
$$

since $\langle Q v, v\rangle=0$ if $v \in Q(H)^{T}$. Hence $\lambda \leq \lambda_{k}+\delta$. This is valid for all $\delta>0$, and so

$$
N(\lambda)=\max \left\{j \mid \lambda_{j} \leq \lambda\right\} \leq k
$$

This proves assertion (i).
2. The minimax formula (C.5) directly implies that

$$
\lambda_{k} \leq \max _{\substack{v \in V \\ v \neq 0}} \frac{\langle A v, v\rangle}{\|v\|^{2}} \leq \lambda+\delta .
$$

Hence $\lambda_{k} \leq \lambda+\delta$. This is valid for all $\delta>0$, and so

$$
N(\lambda)=\max \left\{j \mid \lambda_{j} \leq \lambda\right\} \geq k .
$$

This is assertion (ii).

THEOREM C. 9 (Lidskii's Theorem). Suppose that $B$ is an operator of trace class on $L^{2}\left(M, \Omega^{\frac{1}{2}}(M)\right)$, given by the integral kernel

$$
K \in C^{\infty}\left(M \times M ; \Omega^{\frac{1}{2}}(M \times M)\right)
$$

Then $K_{\Delta}$, the restriction to the diagonal $\Delta:=\{(m, m): m \in M\}$, has a well-defined density; and

$$
\begin{equation*}
\operatorname{tr} B=\int_{\Delta} K_{\Delta} . \tag{C.8}
\end{equation*}
$$

We will also use the following general result of Keel-Tao [?]:
THEOREM C. 10 (Abstract Strichartz estimates). Let ( $X, \mathcal{M}, \mu$ ) be a $\sigma$-finite measure space, and let $U \in L^{\infty}\left(\mathbb{R}, L\left(L^{2}(X)\right)\right.$ satisfy

$$
\begin{gather*}
\|U(t)\|_{\mathcal{B}\left(L^{2}(X)\right)} \leq A, \quad t \in \mathbb{R} \\
\left\|U(t) U(s)^{*} f\right\|_{L^{\infty}(X, \mu)} \leq A h^{-\mu}|t-s|^{-\sigma}\|f\|_{L^{1}(X, \mu)}, \quad t, s \in \mathbb{R} \tag{C.9}
\end{gather*}
$$

where $A, \sigma, \mu>0$ are fixed.
The for every pair $p, q$ satisfying

$$
\frac{2}{p}+\frac{2 \sigma}{q}=\sigma, \quad 2 \leq p \leq \infty, \quad 1 \leq q \leq \infty, \quad(p, q) \neq(2, \infty)
$$

we have

$$
\begin{equation*}
\left(\int_{\mathbb{R}}\|U(t) f\|_{L^{q}(X, \mu)}^{p} d t\right)^{\frac{1}{p}} \leq B h^{-\frac{\mu}{p \sigma}}\|f\|_{L^{2}(X, \mu)} . \tag{C.10}
\end{equation*}
$$

We should stress that in the application to bounds on approximate solution (Section 10.3) we only use the "interior" exponent $p=q$ which does not require the full power of $[?]$ - see $[\mathrm{S}]$. For the reader's convenience we present the proof of that case.

Proof of the case $p=q$ : 1. A rescaling in time easily reduces the estimate to the case $h=1$.
2. The estimate we want reads

$$
\begin{equation*}
\|U(t) f\|_{L^{p}\left(\mathbb{R}_{t} \times X\right)} \leq B\|f\|_{L^{2}(X)}, \quad \frac{1}{p}=\frac{\sigma}{2(1+\sigma)} \tag{C.11}
\end{equation*}
$$

Let $p^{\prime}$ denote the exponent dual to $p: 1 / p+1 / p^{\prime}=1$. Then, since $L^{p^{\prime}}$ is dual to $L^{p}$, (C.11) is equivalent to

$$
\int_{\mathbb{R} \times X} U(t) f(x) G(t, x) d \mu(x) d t \leq\|f\|_{L^{2}(X)}\|G\|_{L^{p^{\prime}}(\mathbb{R} \times X)}
$$

for all $G \in L^{p^{\prime}}(\mathbb{R} \times X)$, and that in turn means that

$$
\left\|\int_{\mathbb{R}} U(t)^{*} G(t) d t\right\|_{L^{2}(X)} \leq C\|G\|_{L^{p^{\prime}}(\mathbb{R} \times X)} .
$$

In other words,

$$
\begin{align*}
& \left|\int_{\mathbb{R}} \int_{\mathbb{R}}\left\langle U(t)^{*} G(t), U(s)^{*} F(s)\right\rangle d t d s\right|  \tag{C.12}\\
& \quad \leq C\|G\|_{L^{p^{\prime}}(\mathbb{R} \times X)}\|F\|_{L^{p^{\prime}}(\mathbb{R} \times X)} .
\end{align*}
$$

3. We now apply the Riesz-Thorin interpolation theorem (see for instance [H1, Theorem 7.1.12]) to $U(t) U(s)^{*}$ with fixed $t, s \in \mathbb{R}$. The two fixed time estimates provided by the hypothesis (C.9) give:

$$
\left\|U(t) U(s)^{*}\right\|_{\mathcal{B}\left(L^{p^{\prime}}, L^{p}\right)} \leq A^{3-2 / p^{\prime}}|t-s|^{-\sigma\left(2 / p^{\prime}-1\right)}, \quad 1 \leq p^{\prime} \leq 2
$$

and in particular,

$$
\begin{align*}
& \left|\left\langle U(t)^{*} G(t), U(s)^{*} F(s)\right\rangle\right| \\
& \quad \leq A^{3-2 / p^{\prime}}|t-s|^{-\sigma\left(2 / p^{\prime}-1\right)}\|G(t)\|_{L^{p^{\prime}}(X)}\|F(s)\|_{L^{p^{\prime}}(X)} . \tag{C.13}
\end{align*}
$$

4. Finally we invoke the Hardy-Littlewood-Sobolev inequality which says that if $K_{a}(t)=|t|^{-1 / a}$ and $1<a<\infty$ then

$$
\begin{gather*}
\left\|K_{a} * u\right\|_{L^{r}(\mathbb{R})} \leq C\|u\|_{L^{p^{\prime}}(\mathbb{R})} \\
\frac{1}{p}+\frac{1}{r}=\frac{1}{a}, \quad 1<p^{\prime}<r \tag{C.14}
\end{gather*}
$$

see [H1, Theorem 4.5.3]. To obtain (C.12) from (C.13) we apply (C.14) with

$$
\frac{1}{a}=\sigma\left(\frac{2}{p^{\prime}}-1\right), \quad \frac{1}{p}+\frac{1}{r}=\frac{1}{a}, \quad p=r
$$

which has a unique solution

$$
p=\frac{2(1+\sigma)}{\sigma} .
$$

This completes the proof.

## Appendix D. Fredholm theory

In this appendix we will describe the role of the Schur complement formula in spectral theory, in particular in analytic Fredholm theory. Our presentation follows [S-Z2].

## D. 1 Grushin problems

Linear algebra. The Schur complement formula states for two-bytwo systems of matrices that if

$$
\left(\begin{array}{ll}
P & R_{-} \\
R_{+} & R_{0}
\end{array}\right)^{-1}=\left(\begin{array}{ll}
E & E_{+} \\
E_{-} & E_{0}
\end{array}\right)
$$

then $P$ is invertible if and only if $E_{0}$ is invertible, with

$$
\begin{equation*}
P^{-1}=E-E_{+} E_{0}^{-1} E_{-}, \quad E_{0}^{-1}=R_{0}-R_{+} P^{-1} R_{-} \tag{D.1}
\end{equation*}
$$

Generalization. We can generalize to problems of the form

$$
\left(\begin{array}{ll}
P & R_{-}  \tag{D.2}\\
R_{+} & O
\end{array}\right)\binom{u}{u_{-}}=\binom{v}{v_{+}}
$$

where

$$
P: X_{1} \rightarrow X_{2}, \quad R_{+}: X_{1} \rightarrow X_{+}, \quad R_{-}: X_{-} \rightarrow X_{2}
$$

for appropriate Banach spaces $X_{1}, X_{2}, X_{+}, X_{-}$. We call (D.2) a Grushin problem. (In practice, we start with an operator $P$ and build a Grushin problem by choosing $R_{ \pm}$, in which case it is normally sufficient to take $R_{0}=0$.)

If the Grushin problem (D.2) is invertible, we call it well-posed and we write its inverse as follows:

$$
\binom{u}{u_{-}}=\left(\begin{array}{ll}
E & E_{+}  \tag{D.3}\\
E_{-} & E_{0}
\end{array}\right)\binom{v}{v_{+}}
$$

for operators

$$
E: X_{2} \rightarrow X_{1}, \quad E_{0}: X_{+} \rightarrow X_{-}, \quad E_{+}: X_{+} \rightarrow X_{1}, \quad E_{-}: X_{2} \rightarrow X_{-} .
$$

LEMMA D. 1 (The operators in a Grushin problem). If (D.2) is well-posed, then the operators $R_{+}, E_{-}$are surjective, and the operators $E_{+}, R_{-}$are injective.

## D. 2 Fredholm operators

DEFINITIONS. (i) A bounded linear operator $P: X_{1} \rightarrow X_{2}$ is called a Fredholm operator if the kernel of $P$,

$$
\operatorname{ker} P:=\left\{u \in X_{1} \mid P u=0\right\}
$$

and the cokernel of $P$,

$$
\text { coker } P:=X_{2} / \overline{P X_{1}}, \text { where } P X_{1}:=\left\{P u \mid u \in X_{1}\right\}
$$

are both finite dimensional.
(ii) The index of a Fredholm operator is

$$
\text { ind } P:=\operatorname{dim} \operatorname{ker} P-\operatorname{dim} \operatorname{coker} P \text {. }
$$

EXAMPLE. Many important Fredholm operators have the form

$$
\begin{equation*}
P=I+K \tag{D.4}
\end{equation*}
$$

where $K$ a compact operator mapping a Banach space $X$ to itself.
Theorem D. 3 below shows that the index does not change under continuous deformations of Fredholm operators (with respect to operator norm topology). Hence for operators of the form (D.4) the index is 0 :

$$
\text { ind } P=\operatorname{ind}(I+t K)=\operatorname{ind} I=0 \quad(0 \leq t \leq 1) .
$$

The connection between Grushin problems and Fredholm operators is this:

THEOREM D. 2 (Grushin problem for Fredholm operators).
(i) Suppose that $P: X_{1} \rightarrow X_{2}$ is a Fredholm operator.

Then there exist finite dimensional spaces $X_{ \pm}$and operators $R_{-}$: $X_{-} \rightarrow X_{2}, R_{+}: X_{1} \rightarrow X_{+}$, for which the Grushin problem (D.2) is well posed. In particular, $P X_{1} \subset X_{2}$ is closed.
(ii) Conversely, suppose that that for some choice of spaces $X_{ \pm}$and operators $R_{ \pm}$, the Grushin problem (D.2) is well posed.

Then $P: X_{1} \rightarrow X_{2}$ is a Fredholm operator if and only if $E_{0}: X_{+} \rightarrow$ $X_{-}$is a Fredholm operator; in which case

$$
\begin{equation*}
\text { ind } P=\operatorname{ind} E_{0} . \tag{D.5}
\end{equation*}
$$

Assertion (ii) is particularly useful when the spaces $X_{ \pm}$are finite dimensional.

Proof. 1. Assume $P: X_{1} \rightarrow X_{2}$ is Fredholm. Let $n_{+}:=\operatorname{dim} \operatorname{ker} P$ and $n_{-}:=\operatorname{dim}$ coker $P$, and write $X_{+}:=\mathbb{C}^{n_{+}}, X_{-}:=\mathbb{C}^{n_{-}}$. Select then linear operators

$$
R_{-}: X_{-} \rightarrow X_{2}, \quad R_{+}: X_{1} \rightarrow X_{+},
$$

of maximal rank such that

$$
R_{-} X_{-} \cap P X_{1}=\{0\}, \quad \operatorname{ker}\left(\left.R_{+}\right|_{\operatorname{ker} P}\right)=\{0\} .
$$

Then the operator

$$
\left(\begin{array}{cc}
P & R_{-} \\
R_{+} & O
\end{array}\right)
$$

has a trivial kernel and is onto. Hence it is invertible, and by the Open Mapping Theorem the inverse is continuous.

In particular, consider $P$ acting on the quotient space $X_{1} /$ ker $P$, which is a Banach space since ker $P$ is closed. We have $n_{+}=0$, and

$$
P X_{1}=P\left(X_{1} / \operatorname{ker} P\right)=\left(\begin{array}{ll}
P & R_{-}
\end{array}\right)\binom{X_{1} / \operatorname{ker} P}{\{0\}}
$$

is a closed subspace.
2. Conversely, suppose that Grushin problem (D.2) is well-posed. According to Lemma D.1, the operators $R_{+}, E_{-}$are surjective, and the operators $E_{+}, R_{-}$are injective. We take $u_{-}=0$. Then

$$
\left\{\begin{array}{l}
\text { the equation } P u=v \text { is equivalent to }  \tag{D.6}\\
u=E v+E_{+} v_{+}, 0=E_{-} v+E_{0} v_{+} .
\end{array}\right.
$$

This means that

$$
E_{-}: \operatorname{Im} P \rightarrow \operatorname{Im} E_{0},
$$

and so we can define the induced map

$$
E^{\#}: X_{2} / \operatorname{Im} P \rightarrow X_{-} / \operatorname{Im} E_{0}
$$

Since $E_{-}$is surjective, so is $E^{\#}$. Also, ker $E^{\#}=\{0\}$. This follows since if $E_{-} v \in \operatorname{Im} E_{0}$, we can use (D.6) to deduce that $v \in \operatorname{Im} P$. Hence $E^{\#}$ is a bijection of the cokernels $X_{2} \operatorname{Im} P$ and $X_{-} / \operatorname{Im} E_{0}$.
3. Next, we claim that

$$
E_{+}: \operatorname{ker} E_{0} \rightarrow \operatorname{ker} P
$$

is a bijection. Indeed, if $u \in \operatorname{ker} P$, then $u=E_{+} v_{+}$and $E_{0} v_{+}=0$. Therefore $E_{+}$is onto; and this is all we need check, since $E_{+}$injective.

We conclude that

$$
\operatorname{dim} \operatorname{ker} P=\operatorname{dim} \operatorname{ker} E_{0}, \quad \operatorname{dim} \operatorname{coker} P=\operatorname{dim} \operatorname{coker} E_{0} .
$$

In particular, the indices of $P$ and $E_{0}$ are equal.
THEOREM D. 3 (Invariance of the index under deformations). The set of Fredholm operators is open in $L\left(X_{1}, X_{2}\right)$, and the index is constant in each component.

Proof. When $P$ is a Fredholm operator, we can use Theorem D. 2 to obtain $E_{0}: \mathbb{C}^{n_{+}} \rightarrow \mathbb{C}^{n_{-}}$, with

$$
\begin{equation*}
\text { ind } E_{0}=n_{+}-n_{-} . \tag{D.7}
\end{equation*}
$$

by the Rank-Nullity Theorem of linear algebra. The Grushin problem remains well-posed (with the same operators $R_{ \pm}$) if $P$ is replaced by $P^{\prime}$, provided $\left\|P-P^{\prime}\right\|<\epsilon$ for some sufficiently small $\epsilon>0$. Hence the set of Fredholm operators is open.

Using (D.7) we see that the index of $P^{\prime}$ is the same as the index of $P$. Consequently it remains constant in each connected component of the set of Fredholm operators.

We refer to Hörmander [H2, Sect.19.1] for a comprehensive introduction to Fredholm operators

## D. 3 Meromorphic continuation of operators.

The Grushin problem framework provides an elegant proof of the following standard result:

THEOREM D. 4 (Analytic Fredholm Theory). Suppose $\Omega \subset \mathbb{C}$ is a connected open set and $\{A(z)\}_{z \in \Omega}$ is a family of Fredholm operators depending holomorphically on $z$.

Then if $A\left(z_{0}\right)^{-1}$ exists at some point $z_{0} \in \Omega$, the mapping $z \mapsto A(z)^{-1}$ is a meromorphic family of operators on $\Omega$.

Proof. 1. Fix $z_{1} \in \Omega$. We form a Grushin problem for $P=A\left(z_{1}\right)$, as described in the proof of Theorem D.2. The same operators $R_{ \pm}^{z_{1}}$ also provide a well-posed Grushin problem for $P=A(z)$ for $z$ in some sufficiently small neighborhood $V\left(z_{1}\right)$ of $z_{1}$.

According to Theorem D. 3

$$
\operatorname{ind} A(z)=\operatorname{ind} A\left(z_{0}\right)=0
$$

Consequently

$$
n_{+}=n_{-}=n,
$$

and $E_{0}^{z_{1}}(z)$ is an $n \times n$ matrix with holomorphic coefficients. The invertibility of $E_{0}^{z_{1}}(z)$ is equivalent to the invertibility of $A(z)$.
2. This shows that there exists a locally finite covering $\left\{\Omega_{j}\right\}$ of $\Omega$, and a family of functions $f_{j}$, holomorphic in $\Omega_{j}$, such that if $z \in \Omega_{j}$, then $A(z)$ is invertible precisely when

$$
f_{j}(z) \neq 0 .
$$

Indeed, we can define $f_{j}:=\operatorname{det} E_{0}^{z}$, where $E_{0}^{z}$ exists for $z \in \Omega_{j}$ by the construction in Step 1. Since $\Omega$ is connected and since $A\left(z_{0}\right)$ is invertible for at least one $z_{0} \in \Omega$, none of $f_{j}$ 's is identically zero.

So det $E_{0}(z)$ a non-trivial holomorphic function in $V\left(z_{1}\right)$; and consequently $E_{0}(z)^{-1}$ is a meromorphic family of matrices. Applying (D.1), we conclude that

$$
A(z)^{-1}=E(z)-E_{+}(z) E_{-+}(z)^{-1} E_{-}(z)
$$

is a meromorphic family of operators in the neighborhood $V\left(z_{1}\right)$. Since $z_{1}$ was arbitrary, $A(z)^{-1}$ is in fact meromorphic in all of $\Omega$.

## Appendix E. Symbol calculus on manifolds

E. 1 Definitions. For reader's convenience we provide here some basic definitions.

DEFINITION. An $n$-dimensional manifold $M$ is a Hausdorff topological space with a countable basis, each point of which has a neighbourhood homeomorphic to some open set in $\mathbb{R}^{n}$.

We say that $M$ is a smooth (or $C^{\infty}$ ) manifold if there exists a family $\mathcal{F}$ of homeomorphisms between open sets:

$$
\gamma: U_{\gamma} \longrightarrow V_{\gamma}, \quad U_{\gamma} \subset M, \quad V_{\gamma} \subset \mathbb{R}^{n}
$$

satisfying the following properties:
(i)(Smooth overlaps) If $\gamma_{1}, \gamma_{2} \in \mathcal{F}$ then

$$
\gamma_{2} \circ \gamma_{1}^{-1} \in C^{\infty}\left(V_{\gamma_{2}} \cap V_{\gamma_{1}} ; V_{\gamma_{1}} \cap V_{\gamma_{2}}\right) .
$$

(ii)(Covering) The open sets $U_{\gamma}$ cover $M$ :

$$
\bigcup_{\gamma \in \mathcal{F}} U_{\gamma}=M
$$

(iii) (Maximality) Let $\lambda$ be a homeomorphism of an open set $U_{\lambda} \subset M$ onto an open set $V_{\lambda} \subset \mathbb{R}^{n}$. If for all $\gamma \in \mathcal{F}$,

$$
\gamma \circ \lambda^{-1} \in C^{\infty}\left(V_{\lambda} \cap V_{\gamma} ; V_{\lambda} \cap V_{\gamma}\right)
$$

then $\lambda \in \mathcal{F}$.

We call $\left\{\left(\gamma, U_{\gamma}\right) \mid \gamma \in \mathcal{F}\right\}$ an atlas for $M$. The open set $U_{\gamma} \subset M$ is a coordinate patch.

DEFINITION. A $C^{\infty}$ complex vector bundle over $M$ with fiber dimension $N$ consists of
(i) a $C^{\infty}$ manifold $V$,
(ii) a $C^{\infty} \operatorname{map} \pi: V \rightarrow M$, defining the fibers $V_{x}:=\pi^{-1}(\{x\})$ for $x \in M$, and
(iii) local isomorphisms

$$
\begin{gather*}
V \supset \pi^{-1}(Y) \stackrel{\psi}{\longrightarrow} Y \times \mathbb{C}^{N}  \tag{E.1}\\
\psi\left(V_{x}\right)=\{x\} \times \mathbb{C}^{N},\left.\quad \psi\right|_{V_{x}} \in G L(N, \mathbb{C}),
\end{gather*}
$$

where $G L(N, \mathbb{C})$ is the group of invertible linear transformations on $\mathbb{C}^{N}$ 。

REMARKS. (i) We can choose a covering $\left\{X_{i}\right\}_{i \in I}$ of $M$ such that for each index $i$ there exists

$$
\psi_{i}: \pi^{-1}\left(X_{i}\right) \rightarrow X_{i} \times \mathbb{C}^{N}
$$

with the properties listed in (iii) in the definition of a vector bundle.
Then

$$
g_{i j}:=\psi_{i} \circ \psi_{j}^{-1} \in C^{\infty}\left(X_{i} \cap X_{j} ; G L(N, \mathbb{C})\right)
$$

These maps are the transition matrices.
(ii) It is important to observe that we can recover the vector bundle $V$ from the transition matrices. To see this, suppose that we are given functions $g_{i j}$ satisfying the identities

$$
\begin{cases}g_{i j}(x) \circ g_{j i}(x)=I & \text { for } x \in X_{i} \cap X_{j} \\ g_{i j}(x) \circ g_{j k}(x) \circ g_{k j}(x)=I, & \text { for } x \in X_{i} \cap X_{j} \cap X_{k}\end{cases}
$$

Now form the set $V^{\prime} \subset I \times M \times \mathbb{C}^{N}$, with the equivalence relation $(i, x, t) \sim\left(i^{\prime}, x^{\prime}, t^{\prime}\right)$ if and only if $x=x^{\prime}$ and $t^{\prime}=g_{i^{\prime} i}(x) t$. Then

$$
V=V^{\prime} / \sim .
$$

DEFINITION. A section of the vector bundle $V$ is a smooth map

$$
u: M \rightarrow V
$$

such that

$$
\pi \circ u(x)=x \quad(x \in M)
$$

We write

$$
u \in \mathbb{C}^{\infty}(M, V)
$$

EXAMPLE 1: Tangent bundle. Let $M$ be a $C^{\infty}$ manifold and let $N$ be the dimension of $M$. We define the tangent bundle of $M$, denoted

$$
T(M)
$$

by defining the transition functions

$$
g_{\gamma_{i} \gamma_{j}}(x):=\partial\left(\gamma_{i} \circ \gamma_{j}^{-1}\right)(x) \in G L(n, \mathbb{R})
$$

for $x \in U_{\gamma_{i}} \cap U_{\gamma_{j}}$. Its sections $C^{\infty}(M, T(M))$ are the smooth vectorfields on $M$.

EXAMPLE 2: Cotangent bundle. For any vector bundle we can define its dual,

$$
V^{*}:=\bigcup_{x \in X}\left(V_{x}\right)^{*}
$$

since we can take

$$
g_{\gamma_{i} \gamma_{j}}=\left(g_{\gamma_{i} \gamma_{j}}^{*}\right)^{-1} .
$$

If $V=T(M)$, we obtain the cotangent bundle, denoted

$$
T^{*}(M) .
$$

Its sections $C^{\infty}\left(M, T^{*}(M)\right)$ are the differential one-forms on $M$.
EXAMPLE 3: S-density bundles. Let $M$ be an $n$-dimensional manifold and let $\left(U_{\gamma}, \gamma_{\alpha}\right)$ form a set of coordinate patches of $X$.

We define the $s$-density bundle over $X$, denoted

$$
\Omega^{s}(M),
$$

by choosing the following transition functions:

$$
g_{\gamma_{i} \gamma_{j}}(x):=\left|\operatorname{det} \partial\left(\gamma_{i} \circ \gamma_{j}^{-1}\right)\right|^{s} \circ \gamma_{j}(x),
$$

for $x \in U_{\gamma_{i}} \cap U_{\gamma_{j}}$.
This is a line bundle over $M$, that is, a bundle with with fibers of complex dimension one.

## E. 2 Pseudodifferential operators on manifolds.

Pseudodifferential operators. In this section $M$ denotes a smooth, $n$-dimensional compact Riemannian manifold without boundary. As above, we have $\left\{\left(\gamma, U_{\gamma}\right) \mid \gamma \in \mathcal{F}\right\}$ for the atlas of $M$, where each $\gamma$ is a smooth diffeomorphism of the coordinate patch $U_{\gamma} \subset M$ onto an open subset $V_{\kappa} \subset \mathbb{R}^{n}$.

NOTATION. Recall from $\S 9.3$ that a class symbols for which we have invariance under coordinate chages is given by

$$
S^{m, k}=\left\{a \in C^{\infty}\left(\mathbb{R}^{2 n}\right):\left|\partial_{x}^{\alpha} \partial_{\xi}^{\beta} a\right| \leq C_{\alpha \beta} h^{-k}\langle\xi\rangle^{m-|\beta|}\right\} .
$$

The index $k$ records how singular the symbol $a$ is as $h \rightarrow 0$, and $m$ controls the growth rate as $|\xi| \rightarrow \infty$.
DEFINITION. A linear operator

$$
A: C^{\infty}(M) \rightarrow C^{\infty}(M)
$$

is called a pseudodifferential operator if there exist integers $m, k$ such that for each coordinate patch $U_{\gamma}$, there exists a symbol $a_{\gamma} \in S^{m, k}$ such that for any $\varphi, \psi \in C_{\mathrm{c}}^{\infty}\left(U_{\gamma}\right)$ and for each $u \in C^{\infty}(M)$

$$
\begin{equation*}
\varphi A(\psi u)=\varphi \gamma^{*} a_{\gamma}^{\mathrm{w}}(x, h D)\left(\gamma^{-1}\right)^{*}(\psi u) . \tag{E.2}
\end{equation*}
$$

NOTATION. (i) In this case, we write

$$
A \in \Psi^{m, k}(M)
$$

and sometimes call $A$ a quantum observable.
(ii) To simplify notation, we also put

$$
\Psi^{k}(M):=\Psi^{0, k}(M), \quad \Psi(M):=\Psi^{0,0}(M)
$$

The symbol of a pseudodifferential operator. Our goal is to associate with a pseudodifferential operator $A$ a symbol $a$ defined on $T^{*} M$, the cotangent space of $M$.

The first lemma is a direct consequence of Lemmma 9.14:
LEMMA E. 1 (More on disjoint support). Let $b \in S^{m, k}$ and suppose $\varphi, \psi \in C_{\mathrm{c}}^{\infty}\left(\mathbb{R}^{n}\right)$. If

$$
\begin{equation*}
\operatorname{spt}(\varphi) \cap \operatorname{spt}(\psi)=\emptyset \tag{E.3}
\end{equation*}
$$

then

$$
\begin{equation*}
\left\|\varphi b^{\mathrm{w}}(x, h D) \psi\right\|_{H^{-N} \rightarrow H^{N}}=O\left(h^{\infty}\right) \tag{E.4}
\end{equation*}
$$

for all $N$.

## THEOREM E. 2 (Symbol of a pseudodifferential operator).

There exist linear maps

$$
\begin{equation*}
\sigma: \Psi^{m, k}(M) \rightarrow S^{m, k} / S^{m, k-1}\left(T^{*} M\right) \tag{E.5}
\end{equation*}
$$

and

$$
\begin{equation*}
\text { Op }: S^{m, k}\left(T^{*} M\right) \rightarrow \Psi^{m, k}(M) \tag{E.6}
\end{equation*}
$$

such that

$$
\begin{equation*}
\sigma\left(A_{1} A_{2}\right)=\sigma\left(A_{1}\right) \sigma\left(A_{2}\right) \tag{E.7}
\end{equation*}
$$

and

$$
\begin{equation*}
\sigma(\mathrm{Op}(a))=[a] \in S^{m, k} / S^{m, k-1}\left(T^{*} M\right) \tag{E.8}
\end{equation*}
$$

We call $a=\sigma(A)$ the symbol of the pseudodifferential operator $A$.
REMARK. In the identity (E.8) " $a d$ " denotes the equivalence class of $a$ in $S^{m, k} / S^{m, k-1}\left(T^{*} M\right)$. This means that

$$
[a]=[\hat{a}] \quad \text { if and only if } \quad a-\hat{a} \in S^{m, k-1}\left(T^{*} M\right) .
$$

The symbol is therefore uniquely defined in $S^{m, k}$, up to a lower order term which is less singular as $h \rightarrow 0$.

Proof. 1. Let $U$ be an open subset of $\mathbb{R}$. Suppose that $B: C_{\mathrm{c}}^{\infty}(U) \rightarrow$ $C^{\infty}(U)$ and that for all $\varphi, \psi \in C^{\infty}$ the mapping $u \longmapsto \varphi B \psi u$ belongs to $\Psi^{m, k}\left(\mathbb{R}^{n}\right)$, for all $u \in \mathcal{S}$.

We claim that there then exists a symbol $a \in S_{\mathrm{loc}}^{k}\left(U,\langle\xi\rangle^{m}\right)$ such that

$$
\begin{equation*}
B=a(x, D)+B_{0} \tag{E.9}
\end{equation*}
$$

where for all $m$

$$
\begin{equation*}
B_{0}: H_{\mathrm{c}}^{-m}(U) \rightarrow H_{\mathrm{loc}}^{m}(U) \quad \text { is } O\left(h^{\infty}\right) . \tag{E.10}
\end{equation*}
$$

To see this, first choose a locally finite partition of unity $\left\{\psi_{j}\right\}_{j \in J} \subset$ $C_{\mathrm{c}}^{\infty}(U)$ :

$$
\sum_{j \in J} \psi_{j}(x) \equiv 1 \quad(x \in U)
$$

Then

$$
\psi_{j} B \psi_{k}=a_{j k}^{\mathrm{w}}(x, h D)
$$

where $a_{j k} \in S^{k}(\langle\xi\rangle)$ and $a_{j k}(x, \xi)=0$ if $x \notin \operatorname{spt} \psi_{j}$. Now put

$$
a:=\sum_{j, k}^{\prime} a_{j k}(x, \xi) \in S_{\mathrm{loc}}^{k}\left(\langle\xi\rangle^{m}\right),
$$

where we are sum over those indices $j, k$ 's for which $\operatorname{spt} \psi_{j} \cap \operatorname{spt} \psi_{k} \neq \emptyset$. This sum is consequently locally finite.
2. We must next verify (E.10) for

$$
B_{0}:=B-a(x, h D)=\sum_{j, k}^{\prime \prime} \psi_{j} B \psi_{k},
$$

the sum over $j, k$ 's for which

$$
\operatorname{spt} \psi_{j} \cap \operatorname{spt} \psi_{k}=\emptyset
$$

Let $K_{B}(x, y)$ be the Schwartz kernel of $B$. Then the Schwartz kernel of $B_{0}$ is

$$
\begin{equation*}
K_{B_{0}}(x, y)=\sum_{j, k}^{\prime \prime} \psi_{j}(x) K_{B}(x, y) \psi_{k}(y), \tag{E.11}
\end{equation*}
$$

with the sum locally finite in $U \times U$. The operators $\psi_{j} B \psi_{k}$ satisfy the assumptions of Lemma 7.4, and hence have the desired mapping property. Because of the local finiteness of (E.11) we get the global mapping property from $H_{\text {loc }}^{-m}$ to $H_{\text {loc }}^{m}$.
3. For each coordinate chart $\left(\gamma, U_{\gamma}\right)$, where $\gamma: U_{\gamma} \rightarrow V_{\gamma}$, we can now use (E.9) with $X=V_{\gamma}$ and $B=\left(\gamma^{-1}\right)^{*} A \gamma^{*}$, to define $a_{\gamma} \in T^{*}\left(U_{\gamma}\right)$.

The second part of Theorem 9.12 shows that if $U_{\gamma_{1}} \cap U_{\gamma_{2}} \neq \emptyset$, then

$$
\begin{equation*}
\left.\left(a_{\gamma_{1}}-a_{\gamma_{2}}\right)\right|_{U_{\kappa_{1}} \cap U_{\gamma_{2}}} \in S^{k-1}\left(T^{*}\left(U_{\gamma_{1}} \cap U_{\gamma_{2}}\right),\langle\xi\rangle^{m}\right) \tag{E.12}
\end{equation*}
$$

Suppose now that we choose a covering of $M$ by coordinate charts, $\left\{U_{\alpha}\right\}_{\alpha \in J}$, and a locally finite partition of unity $\left\{\varphi_{\alpha}\right\}_{\alpha \in J}$ :

$$
\operatorname{spt} \varphi_{\alpha} \subset U_{\gamma}, \quad \sum_{\alpha \in J} \varphi_{j}(x) \equiv 1
$$

and define

$$
a:=\sum_{\alpha \in J} \varphi_{\alpha} a_{\alpha} .
$$

We see from (E.12) that $a \in S^{k}\left(T^{*} M,\langle\xi\rangle^{m}\right)$ is invariantly defined up to terms in $S^{k-1}\left(T^{*} M,\langle\xi\rangle^{m}\right)$. We consequently can define

$$
\sigma(A):=[a] \in S^{k}\left(T^{*} M,\langle\xi\rangle^{m}\right) / S^{k-1}\left(T^{*} M,\langle\xi\rangle^{m}\right)
$$

4. It remains to show the existence of

$$
\text { Op : } S^{k}\left(T^{*} M,\langle\xi\rangle^{m}\right) \longrightarrow \Psi^{m, k}(M), \quad \sigma(\mathrm{Op}(a))=[a]
$$

Suppose that for our covering of $M$ by coordinate charts, $\left\{U_{\alpha}\right\}_{\alpha \in J}$, we choose $\left\{\psi_{\alpha}\right\}_{\alpha \in J}$ such that

$$
\operatorname{spt} \psi_{\alpha} \subset U_{\gamma}, \quad \sum_{\alpha \in J} \psi_{j}^{2}(x) \equiv 1,
$$

a sum which is locally finite. Define

$$
A:=\sum_{\alpha \in J} \psi_{\alpha} \gamma_{\alpha}^{*} \operatorname{Op}\left(\tilde{a}_{\alpha}\right)\left(\gamma_{\alpha}^{-1}\right)^{*} \psi_{\alpha},
$$

where $\tilde{a}_{\alpha}(x, \xi):=a\left(\gamma_{\alpha}^{-1}(x),\left(\partial \gamma(x)^{T}\right)^{-1} \xi\right)$. Theorem 9.12 demonstrates that $\sigma(A)$ equals $[a]$.

Pseudodifferential operators acting on half-densities. We now apply the full strength of Theorem 9.12 by making the pseudodifferential operators act on half-densities.

DEFINITION. A linear operator

$$
A: C^{\infty}\left(M, \Omega^{\frac{1}{2}}(M)\right) \rightarrow C^{\infty}\left(M, \Omega^{\frac{1}{2}}(M)\right)
$$

is called a pseudodifferential operator on half-densities if there exist integers $m, k$ such that for each coordinate patch $U_{\alpha}$, and there exists a symbol $a_{\alpha} \in S^{k}\left(\langle\xi\rangle^{m}\right)$ such that for any $\varphi, \psi \in C_{\mathrm{c}}^{\infty}\left(U_{\gamma}\right)$

$$
\begin{equation*}
\varphi A(\psi u)=\varphi \gamma_{\alpha}^{*} a_{\alpha}^{\mathrm{w}}(x, h D)\left(\gamma_{\alpha}^{-1}\right)^{*}(\psi u) \tag{E.13}
\end{equation*}
$$

for each $u \in C^{\infty}\left(M, \Omega^{\frac{1}{2}}(M)\right)$.

NOTATION. In this case, we write

$$
A \in \Psi^{m, k}\left(M, \Omega^{\frac{1}{2}}(M)\right) .
$$

By adapting the proof of Theorem E. 2 to the case of half-densities using the first part of Theorem 9.12 we obtain

THEOREM E. 3 (Symbol on half-densities). There exist linear maps

$$
\begin{equation*}
\sigma: \Psi^{m, k}\left(M, \Omega^{1 / 2}(M)\right) \rightarrow S^{m, k} / S^{m, k-2}\left(T^{*} M\right) \tag{E.14}
\end{equation*}
$$

and

$$
\begin{equation*}
\text { Op } \left.: S^{m, k}\left(T^{*} M\right) \rightarrow \Psi^{m, k}\left(M, \Omega^{1 / 2}(M)\right)\right) \tag{E.15}
\end{equation*}
$$

such that

$$
\begin{equation*}
\sigma\left(A_{1} A_{2}\right)=\sigma\left(A_{1}\right) \sigma\left(A_{2}\right) \tag{E.16}
\end{equation*}
$$

and

$$
\begin{equation*}
\sigma(\mathrm{Op}(a))=[a] \in S^{m, k} / S^{m, k-2}\left(T^{*} M\right) . \tag{E.17}
\end{equation*}
$$

## E. 3 PDE on manifolds.

We revisit in this last section some of our theory from Chapters 5-7, replacing the flat spaces $\mathbb{R}^{n}$ and $\mathbb{T}^{n}$ by an arbitrary compact Riemannian manifold $(M, g)$, for the metric

$$
g:=\sum_{i, j=1}^{n} g_{i j} d x_{i} d x_{j} .
$$

Write

$$
\left(\left(g^{i j}\right)\right):=\left(\left(g_{i j}\right)\right)^{-1}, \quad \bar{g}:=\operatorname{det}\left(\left(g_{i j}\right)\right) .
$$

## E.3.1 Notation.

Tangent, cotangent bundles. We can use the metric to build an identification of the tangent and cotangent bundles of $M$. We identify

$$
\xi \in T_{x}^{*} M \quad \text { with } \quad X \in T_{x} M
$$

written $\xi \sim X$, provided

$$
\xi(Y)=g_{x}(Y, X)
$$

for all $Y \in T_{x} M$.

Flows. Under the identification $X \sim \xi$, the flow of $H_{p}$ on $T^{*} M$, generated by the symbol

$$
\begin{equation*}
p:=|\xi|_{g}^{2}=\sum_{i, j=1}^{n} g^{i j} \xi_{i} \xi_{j}=\sum_{i, j=1}^{n} g_{i j} X_{i} X_{j}=g(X, X) \tag{E.18}
\end{equation*}
$$

is the geodesic flow on $T M$.
Laplace-Beltrami operator. The Laplace-Beltrami operator $\Delta_{g}$ on $M$ is defined in local coordinates by

$$
\begin{equation*}
\Delta_{g}:=\frac{1}{\sqrt{\bar{g}}} \sum_{i, j=1}^{n} \frac{\partial}{\partial x_{j}}\left(g^{i j} \sqrt{\bar{g}} \frac{\partial}{\partial x_{j}}\right) . \tag{E.19}
\end{equation*}
$$

The function $p$ defined by (E.18) is the symbol of the LaplaceBeltrami operator $-h^{2} \Delta_{g}$.

PDE on manifolds. Given then a potential $V \in C^{\infty}(M)$, we can define the Schrödinger operator

$$
\begin{equation*}
P(h):=-h^{2} \Delta_{g}+V(x) . \tag{E.20}
\end{equation*}
$$

The flat wave equation from Chapter 5 is replaced by an equation involving the Laplace-Beltrami operator:

$$
\begin{equation*}
\left(\partial_{t}^{2}+a(x) \partial_{t}-\Delta_{g}\right) u=0 \tag{E.21}
\end{equation*}
$$

The unknown $u$ is a function of $x \in M$ and $t \in \mathbb{R}$.
Half-densities. Half-densities on $M$ can be identified with functions using the Riemannian density:

$$
u=u(x)|d x|^{\frac{1}{2}}=\tilde{u}(x)\left(g^{\frac{1}{2}} d x\right)^{\frac{1}{2}}
$$

E.3.2 Damped wave equation on manifolds. We consider this initial-value problem for the wave equation:

$$
\left\{\begin{array}{cl}
\left(\partial_{t}^{2}+a(x) \partial_{t}-\Delta\right) u=0 & \text { on } M \times \mathbb{R}  \tag{E.22}\\
u=0, u_{t}=f & \text { on } M \times\{t=0\}
\end{array}\right.
$$

where $a \geq 0$; and, as in Chapter 6, define the energy of a solution at time $t$ to be

$$
E(t):=\frac{1}{2} \int_{M}\left(\partial_{t} u\right)^{2}+\left|\partial_{x} u\right|^{2} d x
$$

It is then straightforward to adapt the proofs in $\S 5.3$ to establish

THEOREM E. 4 (Exponential decay on manifold). Suppose $u$ solves the wave equation with damping (E.21), with the initial conditions

Assume also that there exists a time $T>0$ such that each geodesic of length greater than or equal to $T$ intersects the set $\{a>0\}$.

Then there exist constants $C, \beta>0$ such that

$$
\begin{equation*}
E(t) \leq C e^{-\beta t}\|f\|_{L^{2}} \tag{E.23}
\end{equation*}
$$

for all times $t \geq 0$.
E.3.3 Weyl's Law for compact manifolds. More work is needed to generalize Weyl's Law from Chapter 6 to manifolds. We will prove it using a different approach, based on the Spectral Theorem.

First, we need to check that the spectrum is discrete and that follows from the compactness of the resolvent:

LEMMA E. 5 (Resolvent on manifold). If $P$ is defined by (E.20), then

$$
(P+i)^{-1}=O(1): L^{2}(M) \rightarrow H_{h}^{2}(M),
$$

where the semiclassical Sobolev spaces are defined as in §7.1.

We prove this by the same method as that for Lemma 7.1.
Eigenvalues and eigenfunctions. According to Riesz's Theorem on the discreteness of the spectrum of a compact operator, we conclude that the spectrum of $(P+i)^{-1}$ is discrete, with an accumulation point at 0 .

Hence we can write

$$
\begin{equation*}
P(h)=\sum_{j=1}^{\infty} E_{j}(h) u_{j}(h) \otimes \overline{u_{j}(h)}, \tag{E.24}
\end{equation*}
$$

where $\left\{u_{j}(h)\right\}_{j=1}^{\infty}$ is an orthonormal set of all eigenfunctions of $P(h)$ :

$$
P(h) u_{j}(h)=E_{j}(h) u_{j}(h), \quad\left\langle u_{k}(h), u_{l}(h)\right\rangle=\delta_{l k},
$$

and

$$
E_{j}(h) \rightarrow \infty .
$$

THEOREM E. 6 (Functional calculus). Suppose that $f$ is a holomorphic function, such that for $|\operatorname{Im} z| \leq 2$ and any $N$ :

$$
f(z)=O\left(\langle z\rangle^{-N}\right)
$$

Define

$$
\begin{align*}
& f(P):= \\
& \frac{1}{2 \pi i} \int_{\mathbb{R}}(t-i-P)^{-1} f(t-i)-(t+i-P)^{-1} f(t+i) d t . \tag{E.25}
\end{align*}
$$

Then $f(P) \in \Psi^{-\infty}(M)$, with

$$
\sigma(f(P))=f\left(|\xi|_{g}^{2}+V(x)\right)
$$

Furthermore,

$$
\begin{equation*}
f(P)=\sum_{j=1}^{\infty} f\left(E_{j}(h)\right) u_{j}(h) \otimes \overline{u_{j}(h)} \tag{E.26}
\end{equation*}
$$

in $L^{2}$.

Proof. 1. The statement (E.26) follows from (E.24), which shows that

$$
(P-z)^{-1}=\sum_{j=1}^{\infty} \frac{u_{j} \otimes \overline{u_{j}}}{E_{j}(h)-z} .
$$

Since $f$ decays rapidly as $t \rightarrow \infty$, we can compute residues in (E.25) to conclude that

$$
\begin{aligned}
& f\left(E_{j}(h)\right)= \\
& \frac{1}{2 \pi i} \int_{\mathbb{R}}\left(\left(t-i-E_{j}(h)\right)^{-1} f(t-i)-\left(t+i-E_{j}(h)\right)^{-1} f(t+i)\right) d t .
\end{aligned}
$$

2. We now use Beals's Theorem 9.8, to deduce that $f(P)$ is a pseudodifferential operator. As discussed in Appendix E all we need to show is that for $\varphi, \psi \in C_{\mathrm{c}}^{\infty}(M)$, with supports in arbitrary coordinate patches, $\varphi f(P) \psi$ is a pseudodifferential operator. As described there it can be considered as an operator on $\mathbb{R}^{n}$ and, by Theorem 9.8 , it suffices to check that for any linear $l_{j}(x, \xi)$ we have

$$
\left\|\operatorname{ad}_{l_{1}(x, h D)} \circ \cdot \circ \operatorname{ad}_{l_{N}(x, h D)} f(P)\right\|_{L^{2} \rightarrow L^{2}}=O\left(h^{N}\right)
$$

To show this, note that according to Lemma E.5,

$$
\|(P-t \pm i)^{-1}\left(\operatorname{ad}_{L_{1}} \circ \cdots \circ \operatorname{ad}_{L_{k}} P(P-t \pm i)^{-1} \|_{L^{2} \rightarrow L^{2}}=O\left(h^{k}\right)\right.
$$

where $L_{j} \in \Psi^{0,0}(M)$. Now for a linear function $l$ on $\mathbb{R}^{2 n}$,

$$
\begin{aligned}
& \operatorname{ad}_{l(x, h D)}\left(\varphi(P-t \pm i)^{-1} \psi\right)= \\
& \quad-(P-t \pm i)^{-1}\left(\operatorname{ad}_{L} P\right)(P-t \pm i)^{-1}+O_{L^{2} \rightarrow L^{2}}(h)
\end{aligned}
$$

where $L \in \Psi^{0,0}(M)$. The rapid decay of $f$ gives

$$
\begin{gathered}
\left\|\operatorname{ad}_{L} \int_{\mathbb{R}} f(t)(t \pm i-P)^{-1} d t\right\| \leq \\
\int_{\mathbb{R}}|f(t)|\left\|(P-t \pm i)^{-1} \operatorname{ad}_{L} P(P-t \pm i)^{-1}\right\|_{L^{2} \rightarrow L^{2}} d t=O(h),
\end{gathered}
$$

and this argument can be easily iterated.
3. Since

$$
\left.\mathrm{Op}\left(|\xi|_{g}^{2}+V(x)-t \pm i\right)^{-1}\right)(P-t \pm i)=I+O_{L^{2} \rightarrow L^{2}}(h)
$$

it follows that

$$
\left.\mathrm{Op}\left(|\xi|_{g}^{2}+V(x)-t \pm i\right)^{-1}\right)=(P-t \pm i)^{-1}+O_{L^{2} \rightarrow L^{2}}(h) .
$$

Hence the symbol of $(P+t \pm i)^{-1}$ (which we already know is a pseudodifferential operator) is given by $\left(|\xi|_{g}^{2}+V(x)-t \pm i\right)^{-1}$.

A residue calculation now shows us that

$$
f(P)=\operatorname{Op}\left(f\left(|\xi|_{g}^{2}+V(x)-t \pm i\right)\right)+O_{L^{2} \rightarrow L^{2}}(h) ;
$$

that is, the symbol of $f(P)$ is $f\left(|\xi|_{g}^{2}+V(x)\right)$.

## THEOREM E. 7 (Weyl's asymptotics on compact manifolds).

 For any $a<b$, we have$$
\begin{align*}
& \#\{E(h) \mid a \leq E(h) \leq b\}= \\
& \quad \frac{1}{(2 \pi h)^{n}}\left(\operatorname{Vol}_{T^{*} M}\left\{a \leq|\xi|_{g}^{2}+V(x) \leq b\right\}+o(1)\right) \tag{E.27}
\end{align*}
$$

as $h \rightarrow 0$.

Proof. 1. Let $f_{1}, f_{2}$ be two functions satisfying the assumptions of Theorem E. 6 such that for real $x$

$$
\begin{equation*}
f_{1}(x) \leq \mathbf{1}_{[a, b]}(x) \leq f_{2}(x), \tag{E.28}
\end{equation*}
$$

where $\mathbf{1}_{[a, b]}(x)$ is the characteristic function of the interval $[a, b]$.
It follows that

$$
\operatorname{tr} f_{1}(P) \leq \#\{E(h) \mid a \leq E(h) \leq b\} \leq \operatorname{tr} f_{2}(P)
$$

2. Theorem C. 9 now shows that for $j=1,2$

$$
\operatorname{tr} f_{j}(P)=\frac{1}{(2 \pi h)^{n}}\left(\int_{T^{*} M} f_{j}\left(|\xi|_{g}^{2}+V(x)\right) d x d \xi+O(h)\right) .
$$

We note that since $f_{j}(P) \in \Psi^{-\infty}(M)$, the errors in the symbolic computations are all $O\left(h\langle\xi\rangle^{-\infty}\right)$, and hence can be integrated.
3. The final step is to construct $f_{1}^{\epsilon}$ and $f_{2}^{\epsilon}$ satisfying the hypotheses of Theorem E. 6 and (E.28), and such that for $j=1,2$, we have

$$
\int_{T^{*} M} f_{j}^{\epsilon}\left(|\xi|_{g}^{2}+V(x)\right) d x d \xi \rightarrow \operatorname{Vol}_{T^{*} M}\left\{a \leq|\xi|_{g}^{2}+V(x) \leq b\right\}
$$

as $\epsilon \rightarrow 0$. This is done as follows. Define
$\chi_{1}^{\epsilon}:=(1-\epsilon) \mathbf{1}_{[a+\epsilon, b-\epsilon]}-\epsilon\left(\mathbf{1}_{[a-\epsilon, a+\epsilon]}+\mathbf{1}_{[b-\epsilon, b+\epsilon]}\right), \quad \chi_{2}^{\epsilon}:=(1+\epsilon) \mathbf{1}_{[a-\epsilon, b+\epsilon]}$, and then put

$$
f_{j}^{\epsilon}(z):=\frac{1}{2 \pi \epsilon} \int_{\mathbb{R}} \chi_{j}^{\epsilon}(x) \exp \left(-\frac{(x-z)^{2}}{2 \epsilon^{2}}\right) d x .
$$

We easily check that all the assumptions are satisfied.
REMARKS. (i) If $V \equiv 0$, we recover the leading term in the usual Weyl asymptotics of the Laplacian on a compact manifold: let $0=$ $\lambda_{0}<\lambda_{1} \leq \lambda_{2} \leq \cdots \leq \lambda_{j} \rightarrow \infty$ be the complete set of eigenvalues of $-\Delta_{g}$ on $M$. Then

$$
\begin{equation*}
\#\left\{j: \lambda_{j} \leq r\right\} \sim \frac{\operatorname{Vol}\left(B_{\mathbb{R}^{n}}(0,1)\right)}{(2 \pi)^{n}} \operatorname{Vol}(M) r^{n / 2}, \quad r \rightarrow \infty \tag{E.29}
\end{equation*}
$$

In fact, we can take $a=0, b=1$, and $h=1 / \sqrt{r}$, and apply Theorem E.7: the eigenvalues $-\Delta_{g}$ are just rescaled eigenvalues of $-h^{2} \Delta_{g}$ and the $\operatorname{Vol}\left(B_{\mathbb{R}^{n}}(0,1)\right)$ term comes from integrating out the $\xi$ variables.

We note also that (E.29) implies that

$$
\begin{equation*}
j^{2 / n} / C_{M} \leq \lambda_{j} \leq C_{M} j^{2 / n} \tag{E.30}
\end{equation*}
$$

(ii) Also, upon rescaling and applying Theorem C.8, we obtain estimates for counting all the eigenvalues of $P(h)=-h^{2} \Delta_{g}+V(x)$. Let $E_{0}(h)<E_{1}(h) \leq \cdots \leq E_{j}(h) \rightarrow \infty$ be all the eigenvalues of the self-adjoint operator $P(h)$. Then for $r>1$,

$$
\begin{equation*}
\#\left\{j: E_{j}(h) \leq r\right\} \leq C_{M, V} h^{-n} r^{n / 2} \tag{E.31}
\end{equation*}
$$

This crude estimate will be useful in $\S 9.3$.

## Sources and further reading

Chapter 1: The book of Griffiths [G] provides a nice elementary introduction to quantum mechanics. For a modern physical perspective, consult Heller-Tomsovic [H-T] or Stöckmann [St].

Chapter 2: The proof of Theorem 2.11 is from Moser [Mo]; see also Cannas da Silva [CdS]. A PDE oriented introduction to symplectic geometry may be found in Hörmander [H3, Chapter 21].

Chapter 3: Good references are Friedlander and M. Joshi [F-J] and Hörmander [H1]. The PDE example in $\S 3.1$ is from [H1, Section 7.6].

Chapter 4: The presentation of semiclassical calculus is based uponM. Dimassi-Sjöstrand [D-S, Chapter 7]. See also Martinez [M], in particular for the Fefferman-Cordoba proof of the sharp Gårding inequality. The argument presented here followed the proof of [D-S, Theorem 7.12].

Chapter 5: Semiclassical defect measures were introduced independently in Gérard [Ge] and Lions-Paul [L-P]; see also Tartar [T]. The basic results presented here come from [Ge]. Theorem 5.8 comes from [R-T], but the proof here follows [L] and uses also some ideas of Morawetz.

Chapter 6: The proof of Weyl asymptotics is a semiclassical version of the classical Dirichlet-Neumann bracketting proof for the bounded domains.

Chapter 7: Estimates in the classically forbidden region in $\S 7.1$ are known as Agmon or Lithner-Agmon estimates. They play a crucial role in the analysis of spectra of multiple well potential and of the Witten complex: see [D-S, Chapter 6] for an introduction and references. Here we followed an argument of [N], but see also [?, Proposition 3.2]. The presentation of Carleman estimates in $\S 7.2$ is based on discussions with N. Burq and D. Tataru.

Chapter 8: The Quantum Ergodicity Theorem 8.4 is from a 1974 paper of Shnirelman, and it is sometimes referred to as Shnirelman's Theorem. The first complete proof, in a different setting, was provided by Zelditch. We have followed his more recent proof, as presented in [?]. The same proof applied with finer spectral asymptotics gives a stronger semiclassical version, first presented in [?].
Chapter 9: For $h=1, \Psi^{m, k}$ form the class of Kohn-Nirenberg pseudodifferential operators: see [H2, §18.1] or [G-S] for a thorough presentation. Much can be said about the properties of semiclassical wave front sets and we refer to Alexandrova [A] for a recent discussion.

The proof of symbol invariance is from the appendix to [?]. The semiclassical wavefront set is an analog of the usual wavefront set in microlocal analysis - see [H2] and is closely related to the frequency set introduced in [?]. [?] presented the semiclassical pointwise bounds reproduced here.

Our presentation of Beals's Theorem follows [D-S, Chapter 8], where it was based on [?]. Theorem 9.9, in a much greater generality, was proved in [?]. The self-contained proof in the simple case considered here comes from the appendix to [?].

Section 9.4 a special of a general result in [?, Théoreme 6.4]. See [?], [?] for examples of conjugation techniques, and [M] for a slightly different perspective.
Chapter 10: The definition of quantization of symplectomorphisms using deformation follows the Heisenberg picture of quantum mechan$i c s$. The proof of Theorem 10.2 comes from [Ch, Section 3] where a stronger version of the result is also given. The construction of $U(t)$ borrows from the essentially standard presentation in [?, Section 7]. For the discussion of the Maslov index see [?] and [?]. Fourier integral operators which are closely related to our discussion of quantization and of propagators are discussed in detail in [?] and [H2, Chapter 25].

Semiclassical Strichartz estimates for $P=-h^{2} \Delta_{g}-1$ appeared explicitely in [?] who used them to prove existence results for non-linear Schrödinger equations on two and three dimensional compact manifolds. We refer to that paper for pointers to the vast literature on Strichartz estimates and their applications. The adaptation of Sogge's $L^{p}$ estimates to the semiclassical setting comes from [?] and was inspired by discussions with N. Burq, H. Koch, C.D. Sogge, and D. Tataru, see [?] and [S].

The proofs for the theorems cited in §10.4 are in [H2, Theorem 21.1.6] and [H2, Theorem 21.1.6]. Theorem 10.18 is a semiclassical analog of the standard $C^{\infty}$ result of Duistermaat-Hörmander [H2, Proposition 26.1.3']. Theorem 10.19 is a semiclassical adaptation of a microlocal result of Duistermaat-Sjöstrand [H2, Proposition 26.3.1].

Theorem 10.20 was proved in one dimension in [?]. See also [?] for more on quasimodes and pseudospectra and for further references.
Appendices: Ilan Hirshberg provided us with Theorem C. 4 and its proof.

## References

[A] I. Alexandrova, Semi-classical wavefront set and Fourier integral operators, Can. J. Math., to appear.
[B-C] J.-M. Bony and J.-Y. Chemin, Espaces fonctionnels associés au calcul de Weyl-Hörmander, Bull. Soc. math. France, 122(1994), 77-118.
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