# ASYMMETRIC TRANSPORT IN CONTINUOUS TOPOLOGICAL INSULATORS 

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In memory of my beloved grandmother, Tova Rosin

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#### Abstract

This thesis concerns continuous models for 2-dimensional topological insulators. Such systems are characterized by asymmetric transport along a 1-dimensional curve representing the interface between two insulating materials. The asymmetric transport is quantified by an interface conductivity. In two distinct settings, we derive tractable analytic formulas for this interface conductivity and provide a large class of perturbations under which it is stable. Our theory applies to models of twisted bilayer graphene, low-energy superconductors, and relativistic electrons (possibly subject to a magnetic field) described by an appropriate Dirac operator, among others.

Our second main focus is numerical approximations of the above systems. We define a modified interface conductivity on a box with periodic boundary conditions, and show that it is stable and converges rapidly to its infinite-space analogue as the size of the box goes to infinity. We illustrate with several examples that one can restrict a topological insulator to a large and discrete torus to obtain accurate numerical evaluations of the conductivity. Numerical techniques that do not require periodic truncation are also implemented and analyzed. We derive a novel integral equation for the time-harmonic Klein-Gordon equation with appropriate jump conditions along a one-dimensional interface. We implement a fast multipole and sweeping-accelerated iterative algorithm for solving the integral equations, and demonstrate numerically the fast convergence of this method. Several numerical examples of solutions and scattering effects illustrate our theory.


## CHAPTER 1

## INTRODUCTION

When two insulating materials are brought next to each other, the resulting system may admit edge modes and exhibit a robust and asymmetric transport along the interface. Mathematically, this phenomenon is most often explained by a topological characterization of the two insulators and/or the new material they form. Examples and applications are found in various fields, such as solid state physics, quantum computing, and the geophysical sciences [17, 33, 76, 86, 87]. The focus of this thesis is two-fold: first, to provide a rigorous mathematical analysis of a physical observable that quantifies the above behavior; and second, to perform and analyze numerical approximations of the observable as well as surface waves corresponding to these systems.

In Sections 2 and 3, the physical systems are modeled by (single-particle) Hamiltonians $H$ acting on $\mathcal{H}:=L^{2}\left(\mathbb{R}^{2}\right) \otimes \mathbb{C}^{n}$ for $n \in \mathbb{N}_{+}$. We assume that $H$ is a self-adjoint pseudodifferential operator from its domain of definition $\mathcal{D}(H) \subset \mathcal{H}$ to $\mathcal{H}$. We label the spatial coordinates $(x, y) \in \mathbb{R}^{2}$ and the corresponding dual variables $(\xi, \zeta) \in \mathbb{R}^{2}$. In the simplest case, the coefficients of (the symbol of) $H$ are constant in $(x, y)$ for $|y| \geq y_{0}$. Thus $H$ fits the above description with $|y| \geq y_{0}$ corresponding to the insulating domains, call them $H_{ \pm}$, and the interface of the two materials contained in $|y|<y_{0}$. Note that a Hamiltonian is called insulating (with respect to an energy interval) if it has a spectral gap (in that interval).

More specifically, in Section 2 below, we will fix $m>0$ and $-\infty<E_{1}<E_{2}<\infty$ and consider a large class of Hamiltonians $H$ including those that satisfy
(H1) Let $H=\operatorname{Op}(\sigma)$ with $\sigma \in E S_{1,0}^{m}$. Suppose there exist symbols $\sigma_{ \pm} \in E S_{1,0}^{m}$ independent of $(x, y)$ with no spectrum in the open interval $\left(E_{1}, E_{2}\right)$, such that $\sigma=\sigma_{+}$whenever $y \geq y_{0}$ and $\sigma=\sigma_{-}$whenever $y \leq-y_{0}$, for some $y_{0}>0$.

We refer to Appendix A. 1 for notation and definitions. In particular, the symbol class
$E S_{1,0}^{m}$ is defined below (A.1.4). Moreover, $\sigma$ is the symbol of a Hamiltonian $H=\operatorname{Op}(\sigma)$ written in Weyl quantization (A.1.1) (with $h=1$ ), while $m$ is the order of the operator. In most applications, $\sigma(x, y, \xi, \zeta)$ is the matrix-valued symbol of a differential operator of order $m$.

An important hypothesis is the ellipticity condition in the definition of $E S_{1,0}^{m}$, stating that the singular values of the symbol grow like $|(\zeta, \zeta)|^{m}$ at infinity uniformly in the spatial variables. The second main assumption is that the symbol be independent of position for $|y| \geq y_{0}$. The slab $-y_{0}<y<y_{0}$ models the transition from the bulk symbol $\sigma_{-}$to the bulk symbol $\sigma_{+}$. The above ellipticity condition implies that $H$ is self-adjoint with domain of definition the standard Hilbert space $\mathcal{H}^{m}$ in (A.1.3) [20, 56].

Note that we will also use more general hypotheses in Section 2 (allowing for more complicated geometries; see (H0), (H0') and (H1') below), though we now focus on (H1) for ease of exposition.

We now define the following physical observable $[7,10,17,37,40,41,77]$

$$
\begin{equation*}
\sigma_{I}(H, P, \varphi):=\operatorname{Tr} i[H, P] \varphi^{\prime}(H) \tag{1.0.1}
\end{equation*}
$$

where $\varphi \in \mathfrak{S}\left(0,1 ; E_{1}, E_{2}\right)$ and $P=P(x) \in \mathfrak{S}(0,1)$. Here, we denote by $\mathfrak{S}\left(c_{1}, c_{2} ; \lambda_{1}, \lambda_{2}\right)$ the set of switch functions, i.e., smooth real-valued functions $f$ on $\mathbb{R}$ such that $f(x)=c_{1}$ for $x \leq \lambda_{1}$ and $f(x)=c_{2}$ for $x \geq \lambda_{2}$. The union over $\lambda_{1}<\lambda_{2}$ is denoted by $\mathfrak{S}\left(c_{1}, c_{2}\right)$.

The conductivity $\sigma_{I}:=\sigma_{I}(H, P, \varphi)$ is the main object of interest in Section 2. It is a physical observable that can be interpreted as the rate of signal crossing $x_{0} \in \operatorname{supp}\left(P^{\prime}\right)$ per unit time. Indeed, if $\psi(t)=e^{-i t H} \psi_{0}$ solves the Schrödinger equation, then $\frac{d}{d t}\langle P\rangle_{t}=$ $\frac{d}{d t}\langle\psi(t), P \psi(t)\rangle=\operatorname{Tr} i[H, P]\left[\psi \psi^{*}\right](t)$. Replacing the density $\psi \psi^{*}(t)$ by the stationary mixed state density $\varphi^{\prime}(H)$ provides the above interpretation to $\sigma_{I}$. It therefore characterizes asymmetric transport along the interface.

The explicit computation of $\sigma_{I}$ as the above trace is tractable for Dirac operators [7] and
more general operators with sufficiently simple spectral decomposition [10]. In a variety of settings however, $\sigma_{I}$ has been shown to equal a difference of bulk topological invariants:

$$
2 \pi \sigma_{I}(H, P, \varphi)=I\left(H_{+}\right)-I\left(H_{-}\right)
$$

which are often significantly simpler to compute. Such a result is called a bulk-edge or bulkinterface correspondence $[7,10,37,40,42,49,68]$. The above conductivity has also been related to the (necessarily quantized) index of a Fredholm operator in various contexts; see [7, 10]. In Section 2.4, we extend these results to prove that

$$
\begin{equation*}
2 \pi \sigma_{I}(H, P, \varphi)=\operatorname{Index}(\bar{P} U(H) \bar{P}) \tag{1.0.2}
\end{equation*}
$$

for a large class of Hamiltonians including (H1). When $H$ satisfies (H1), the above holds if $\bar{P}=\bar{P}(x)$ is the Heavyside step function, with $P(x)=P \in \mathfrak{S}(0,1)$ smooth as before. Note that (1.0.2) implies that $\sigma_{I}$ is immune to "continuous" perturbations (to be defined in Section 2.4 below), as the right-hand side must be integer-valued. The derivation of (1.0.2) involves a powerful result from [6], which relates the Fredholm index to various traces.

First, however, we directly prove the stability of $\sigma_{I}$ without using index theory; see Section 2.1. In Section 2.2, we prove a bulk-interface correspondence for all Hamiltonians satisfying (H1) by deriving an accessible Fedosov-Hörmander formula for $\sigma_{I}$. We next show the usefulness of such a correspondence and compute the conductivity explicitly from the Fedosov-Hörmander formula for a number of models coming from solid state physics [10, 17, 86] and (properly regularized) equatorial waves [10, 33, 48, 79]; see also [15] for an application to compute topological invariants of replica models for Floquet topological insulators. Section 2.4 then generalizes the theory from Sections $2.1-2.3$ to junction (rather than interface) models, with applications to materials such as twisted bilayer graphene.

The theory in Section 2 applies to a number of systems of differential equations that
appear as low-energy models for many topological insulators and superconductors. However, it does not apply to the magnetic Schrödinger equations that model the integer quantum Hall effect [5, 6, 16], or to partial-differential models with micro-structures [37, 44].

In Section 3, we calculate $\sigma_{I}$ and prove its stability in the context of magnetic Dirac equations. Such equations are used in models of graphene subject to external electric and magnetic fields. The magnetic Dirac Hamiltonian is not elliptic and thus the analysis from Section 2 does not apply. Instead, $\sigma_{I}$ is related to a spectral flow (see (3.0.5) below), which we calculate explicitly.

The magnetic Dirac Hamiltonian implements three domain walls that induce asymmetric transport. These domain walls correspond to transitions in the magnetic field, electric potential, and a "mass term" across two insulating materials. The existing literature has analyzed separately the roles of a bounded mass term [7, 10] and magnetic field [29] in generating asymmetric transport for Dirac models. To our knowledge, this work is the first to combine the two effects.

The magnetic Dirac Hamiltonian, $H$, is given by (3.0.2). Its square satisfies $H^{2}=$ $D_{x}^{2}+\left(D_{y}-A_{2}(x)\right)^{2}+V(x) D_{x} \sigma_{1}+V(x)\left(D_{y}-A_{2}(x)\right) \sigma_{2}+R$ for $R$ bounded. The leading-order terms $D_{x}^{2}+\left(D_{y}-A_{2}(x)\right)^{2}$ are precicely the Iwatsuka Hamiltonian analyzed in [35], where there are many results analogous to the ones in this paper. Bulk invariants for magnetic Schrödinger operators are proposed and analyzed in $[6,16]$. The distinguishing features of our setting are the additional domain walls $m$ (the mass term) and $V$ (the electric potential), the lack of definiteness of $H$ (the spectrum of $H$ is not bounded above or below), and the fact that $H$ is a first-order matrix valued differential operator (instead of a second-order scalar one). As we will see below, the lack of definiteness perhaps provides the biggest challenge; it will be easy to estimate the absolute value of spectral branches of $H$, but obtaining the sign of these branches will require more care.

In [28], the interface conductivity is calculated for magnetic Schrödinger Hamiltonians
with constant magnetic field and confining potentials. The interface conductivity is induced by the potentials much like an "edge conductivity" would be generated by "hard wall" Dirichlet boundary conditions. A bulk-edge correspondence (involving the edge conductivity) for magnetic Dirac models is proven in [29].

A main result in Section 3 is to derive an explicit expression for $2 \pi \sigma_{I}$ by means of a spectral flow; see Theorems 3.1.1 and 3.2.5 below. We show that $2 \pi \sigma_{I}$ is quantized and (for positive energies and constant $m$ and $V$ ) decreases in uniform increments as the strength of the magnetic field increases. This behavior of the interface conductivity resembles the integer quantum Hall effect. Theorem 3.1.1 is a bulk-interface correspondence in that it relates the spectral flow of $H$ to a difference of bulk quantities. The terms in the difference are not expressed as bulk invariants, such as for instance Chern numbers as introduced in [16, 29]. This issue, addressed for non-magnetic Dirac Hamiltonians in [7, 10], is not considered further here. For existing results on the bulk-interface correspondence involving a spectral flow, see also [7, 47].

Section 4 concerns numerical approximations of topological insulators. We consider Hamiltonians $H$ satisfying (H1), and restrict them to a box of size $L$ with periodic boundary conditions. We define a modified interface conductivity (see (4.0.3) below) on the periodic domain, and show that it converges to its infinite space analogue super-algebraically in $L^{-1}$ as $L \rightarrow \infty$. We prove stability of the modified conductivity in the same limit, and demonstrate our theoretical results with numerical examples.

The approximate interface conductivity (4.0.3), which is easily estimated computationally, serves as the main spectral object quantifying the topological nature of the asymmetric transport between insulators in a periodic setting. Several works have addressed the computation of bulk topological indices [62, 63, 64, 67, 78] , based on the notion of "spectral localizer index", typically determined by the signs of eigenvalues of an appropriate finite dimensional
matrix. The numerical index was used recently in [65] to determine numerically the interface along which wave packets may propagate. For an analysis of the propagation of wave packets along curved interfaces in a Dirac model, see [12]. Eigenvalues and edge states for discrete and continuous models are also computed numerically in [83] using a method that avoids artificial Dirichlet boundary conditions by utilizing the resolvent of the Hamiltonian.

In Section 5, we derive novel integral equations for the time-harmonic Klein-Gordon equation in the presence of an interface. Solutions are surface waves that are exponentially localized to the vicinity of the interface, and propagate outward to infinity. The original PDE is inspired by Dirac equations, which are ubiquitous in the analysis of topological insulators (see Section 5.1.2). We analyze the mathematical properties (including well-posedness) of our integral equations in Section 5.2, and present a numerical method for solving them in Section 5.3. Examples of solutions and scattering effects are illustrated in Section 5.4.

Similar models to this one arise, and have been studied in a variety of contexts. For work on the related topic of "leaky quantum graphs" and derivation of similar integral equations, we refer to [43] and references therein. Under the assumption that $\Gamma$ is a compact perturbation of the flat interface, [43] for instance derives asymptotic expansions for generalized eigenfunctions of the Helmholtz operator and obtains expressions for reflection coefficients in a corresponding scattering theory. Note that our setting places no restriction on the angle between the asymptotic branches of $\Gamma$. The point spectrum of elliptic second-order partial differential operators with "singular interactions" along a compact interface has also been analyzed, with an integral formulation in [54]. A Galerkin method for solving it is then proposed there.

Surface waves and plasmon waves also arise naturally in other physical contexts such as the solution of Maxwell's equation in a dielectric medium where the ratio of permittivities approach a negative real number, see $[69,85]$ and the references therein and $[52,51,53]$ for numerical methods. Finally, similar surface-wave preconditioners also referred to as on-
surface radiation conditions have been used in other contexts for solving high-frequency scattering problems in acoustics, electromagnetics, and elasticity, see $[2,3,4,26,25,32,59]$ and the references therein. The on-surface radiation conditions are typically used to improve the performance of iterative solvers in complicated geometries and the high-frequency regime and not for the resolution of surface waves inherent to the governing equations.

## CHAPTER 2

## ELLIPTIC INTERFACE MODELS

### 2.1 Stability of physical observable

The purpose of this section is to show that $\sigma_{I}(H, P, \varphi)$ introduced in (1.0.1) is well defined, and prove its stability under perturbations of $H, P$ and $\varphi$. The following results are obtained using the theory of pseudo-differential operators (see appendix A) and stability properties of the trace. Similar results are derived in $[10,11]$ by showing that $\sigma_{I}$ is the index of a Fredholm operator, which is known to be invariant under continuous deformations [56]. Here, we do not relate $\sigma_{I}$ to any index and instead prove its stability directly, as is done in other contexts in [37, 40, 49, 68].

While the results in this section are significant in their own right, they will also be used to prove Theorem 2.2.1 below (one of the main results of the paper). The following stability properties of $\sigma_{I}$ show that the latter can be computed in the semi-classical limit (see Theorem 2.1.11). This allows for a Fedosov-Hörmander formula (2.2.7) for the conductivity, which is the subject of section 2.2.

To prove that $\sigma_{I}$ is stable, it will be useful to define the following class of pseudodifferential operators, where we fix $\tilde{E}_{1}<\tilde{E}_{2}$, and let $\Phi \in \mathcal{C}_{c}^{\infty}(\mathbb{R})$ such that $\Phi \equiv 1$ in [ $\left.\tilde{E}_{1}, \tilde{E}_{2}\right]$.
(H0) Let $H=\operatorname{Op}(\sigma)$, such that $\sigma \in E S_{1,0}^{m}$ and $\Phi(H) \in \operatorname{Op}\left(S\left(\langle y, \xi, \zeta\rangle^{-\infty}\right)\right)$.

By Proposition A.3.1, any $H$ satisfying (H0) is self-adjoint with domain of definition $\mathcal{D}(H)=\mathcal{H}^{m}$. Moreover, $(z-H)^{-1} \in \operatorname{Op}\left(S\left(\langle\xi, \zeta\rangle^{-m}\right)\right)$ with symbolic bounds that blow up algebraically as $\Im z \rightarrow 0$. For any $\psi \in \mathcal{C}_{c}^{\infty}(\mathbb{R})$ and $s \in \mathbb{N}$, we have $\psi(H)=(i-$ $H)^{-s} \phi_{s}(H)$ for some $\phi_{s} \in \mathcal{C}_{c}^{\infty}(\mathbb{R})$. Using the Helffer-Sjöstrand formula and composition calculus, this implies that $\psi(H) \in \operatorname{Op}\left(S\left(\langle\xi, \zeta\rangle^{-\infty}\right)\right)$ for every $\psi \in \mathcal{C}_{c}^{\infty}(\mathbb{R})$. If, in particular,
$\psi \in \mathcal{C}_{c}^{\infty}\left(\tilde{E}_{1}, \tilde{E}_{2}\right)$, then (H0) and the composition calculus imply that $\psi(H)=\psi(H) \Phi(H) \in$ $\operatorname{Op}\left(S\left(\langle y, \xi, \zeta\rangle^{-\infty}\right)\right)$.

Note that if $\Phi \in \mathcal{C}_{c}^{\infty}\left(E_{1}, E_{2}\right)$, then any operator $H$ satisfying (H1) must also satisfy (H0); see Proposition 2.1.10 below. However, all of the results before Proposition 2.1.10 will assume only that $H$ satisfies (H0).

We first show that $\sigma_{I}(H, P, \varphi)$ is well defined.
Lemma 2.1.1. Suppose $H$ satisfies (H0), and let $P(x)=P \in \mathfrak{S}(0,1)$ and $\varphi \in \mathfrak{S}\left(0,1 ; \tilde{E}_{1}, \tilde{E}_{2}\right)$. Then $[H, P] \varphi^{\prime}(H)$ is trace-class. If $\psi \in \mathcal{C}_{c}^{\infty}\left(\tilde{E}_{1}, \tilde{E}_{2}\right)$, then $q(H)[\psi(H), P]$ is trace-class for any polynomial $q$.

Proof. Recall that $\varphi^{\prime} \in \mathcal{C}_{c}^{\infty}\left(\tilde{E}_{1}, \tilde{E}_{2}\right)$, and hence $\varphi^{\prime}(H) \in \operatorname{Op}\left(S\left(\langle y, \xi, \zeta\rangle^{-\infty}\right)\right)$. Since $[H, P]=$ $(1-P) H P-P H(1-P) \in \operatorname{Op}\left(S\left(\langle x\rangle^{-\infty}\langle y, \xi, \zeta\rangle^{m}\right)\right)$, the composition calculus implies that $[H, P] \varphi^{\prime}(H)$ is trace-class.

Now if $\psi \in \mathcal{C}_{c}^{\infty}\left(\tilde{E}_{1}, \tilde{E}_{2}\right)$, then $\psi(H) \in \operatorname{Op}\left(S\left(\langle y, \xi, \zeta\rangle^{-\infty}\right)\right)$, and thus $[\psi(H), P]=(1-$ P) $\psi(H) P-P \psi(H)(1-P) \in \operatorname{Op}\left(S\left(\langle x, y, \xi, \zeta\rangle^{-\infty}\right)\right)$ by the composition calculus. Since $q(H) \in \operatorname{Op}\left(S\left(\langle\xi, \zeta\rangle^{k m}\right)\right)$ for some $k \in \mathbb{N}$, it follows that

$$
q(H)[\psi(H), P] \in \operatorname{Op}\left(S\left(\langle x, y, \xi, \zeta\rangle^{-\infty}\right)\right)
$$

is trace-class.

Our next result shows that the trace in (1.0.1) is not modified after regularization of the commutator.

Lemma 2.1.2. Let $P(x)=P \in \mathfrak{S}(0,1)$ and $\varphi \in \mathfrak{S}\left(0,1 ; \tilde{E}_{1}, \tilde{E}_{2}\right)$. Let $S$ be any open interval containing $\operatorname{supp} \varphi^{\prime}$, and let $\Psi \in \mathcal{C}_{c}^{\infty}(\mathbb{R})$ such that $\Psi(\lambda)=\lambda$ in $S$. Then given any self-adjoint $H$ for which $[H, P] \varphi^{\prime}(H)$ is trace-class, it follows that $[\Psi(H), P] \varphi^{\prime}(H)$ is trace-class, with

$$
\begin{equation*}
\sigma_{I}(H, P, \varphi)=\operatorname{Tr} i[\Psi(H), P] \varphi^{\prime}(H) \tag{2.1.1}
\end{equation*}
$$

By Lemma 2.1.1, we know that Lemma 2.1.2 applies to any $H$ satisfying (H0).

Proof. Using the Helffer-Sjöstrand formula, we write

$$
\begin{aligned}
{[\Psi(H), P] } & =\left[-\frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\Psi}(z)(z-H)^{-1} d^{2} z, P\right] \\
& =-\frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\Psi}(z)(z-H)^{-1}[H, P](z-H)^{-1} d^{2} z
\end{aligned}
$$

which implies that

$$
[\Psi(H), P] \varphi^{\prime}(H)=-\frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\Psi}(z)(z-H)^{-1}[H, P] \varphi^{\prime}(H)(z-H)^{-1} d^{2} z
$$

Since $\left\|(z-H)^{-1}\right\| \leq|\Im z|^{-1}$ with $\bar{\partial} \tilde{\Psi} \in \mathcal{C}_{c}^{\infty}(\mathbb{C})$ and $|\bar{\partial} \tilde{\Psi}|(z) \leq C|\Im z|^{2}$, the assumption that $[H, P] \varphi^{\prime}(H)$ is trace-class implies that $[\Psi(H), P] \varphi^{\prime}(H)$ must be trace-class as well.

Let $\Phi_{0}, \Phi_{00} \in \mathcal{C}_{c}^{\infty}(\mathbb{R})$ such that $\varphi^{\prime}=\varphi^{\prime} \Phi_{0}=\varphi^{\prime} \Phi_{00}$ and $\Phi_{0}=\Phi_{0} \Phi_{00}$, with $\Psi \Phi_{00}=\lambda \Phi_{00}$. Below, we use the shorthand $f:=f(H)$ for all compactly supported functions $f$. It follows from cyclicity of the trace [57] that

$$
\begin{aligned}
\operatorname{Tr}[H, P] \varphi^{\prime} & =\operatorname{Tr}[H, P] \varphi^{\prime} \Phi_{0}=\operatorname{Tr} \Phi_{0}[H, P] \varphi^{\prime}=\operatorname{Tr} \Phi_{0}\left(\Phi_{00} H P-P H \Phi_{00}\right) \varphi^{\prime} \\
& =\operatorname{Tr} \Phi_{0}\left(\Phi_{00} \Psi P-P \Psi \Phi_{00}\right) \varphi^{\prime}=\operatorname{Tr} \Phi_{0}[\Psi, P] \varphi^{\prime}=\operatorname{Tr}[\Psi, P] \varphi^{\prime}
\end{aligned}
$$

and the proof is complete.
We now show that the conductivity is independent of $\varphi \in \mathfrak{S}\left(0,1 ; \tilde{E}_{1}, \tilde{E}_{2}\right)$.

Proposition 2.1.3. Suppose $H$ satisfies (H0), and let $P(x)=P \in \mathfrak{S}(0,1)$ and $\varphi_{1}, \varphi_{2} \in$ $\mathfrak{S}\left(0,1 ; \tilde{E}_{1}, \tilde{E}_{2}\right)$. Then

$$
\sigma_{I}\left(H, P, \varphi_{1}\right)=\sigma_{I}\left(H, P, \varphi_{2}\right)
$$

Proof. It suffices to show that $\operatorname{Tr} i[H, P] \phi^{\prime}(H)=0$ for all $\phi \in \mathcal{C}_{c}^{\infty}\left(\tilde{E}_{1}, \tilde{E}_{2}\right)$. Let $\psi \in$ $\mathcal{C}_{c}^{\infty}\left(\tilde{E}_{1}, \tilde{E}_{2}\right)$ such that $\psi(\lambda)=\lambda$ for all $\lambda$ in some open interval containing $\operatorname{supp}(\phi)$. It follows that

$$
\begin{aligned}
\operatorname{Tr}[H, P] \phi^{\prime}(H) & =\operatorname{Tr}[\psi(H), P] \phi^{\prime}(\psi(H))=\operatorname{Tr}\left(-\frac{1}{\pi}[\psi(H), P] \int_{\mathbb{C}} \bar{\partial} \tilde{\phi}^{\prime}(z)(z-\psi(H))^{-1} d^{2} z\right) \\
& =\operatorname{Tr}\left(-\frac{1}{\pi}[\psi(H), P] \int_{\mathbb{C}} \bar{\partial} \tilde{\phi}(z)(z-\psi(H))^{-2} d^{2} z\right) \\
& =\operatorname{Tr}\left(-\frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\phi}(z)(z-\psi(H))^{-1}[\psi(H), P](z-\psi(H))^{-1} d^{2} z\right) \\
& =\operatorname{Tr}\left[-\frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\phi}(z)(z-\psi(H))^{-1} d^{2} z, P\right]=\operatorname{Tr}[\phi(H), P],
\end{aligned}
$$

where we have used Lemma 2.1.2 to justify the first equality, the Helffer-Sjöstrand formula for the second and last equalities, integration by parts in $\partial$ for the third equality, cyclicity of the trace for the fourth equality, and the identity $\left[(z-A)^{-1}, B\right]=(z-A)^{-1}[A, B](z-A)^{-1}$ for the fifth equality.

Now, let $P_{1}, P_{2} \in \mathfrak{S}(0,1)$ with $P_{j}=P_{j}(x)$ such that $P P_{1}=P$ and $(1-P)\left(1-P_{2}\right)=1-P$. Then

$$
\begin{aligned}
\operatorname{Tr}[\phi(H), P] & =\operatorname{Tr}((1-P) \phi(H) P-P \phi(H)(1-P))=\operatorname{Tr}(1-P) \phi(H) P-\operatorname{Tr} P \phi(H)(1-P) \\
& =\operatorname{Tr}(1-P) \phi(H) P P_{1}-\operatorname{Tr} P \phi(H)(1-P)\left(1-P_{2}\right) \\
& =\operatorname{Tr} P_{1}(1-P) \phi(H) P-\operatorname{Tr}\left(1-P_{2}\right) P \phi(H)(1-P) \\
& =\operatorname{Tr} P P_{1}(1-P) \phi(H)-\operatorname{Tr}(1-P)\left(1-P_{2}\right) P \phi(H) \\
& =\operatorname{Tr} P(1-P) \phi(H)-\operatorname{Tr}(1-P) P \phi(H)=0
\end{aligned}
$$

and the proof is complete.

We now prove the stability of the trace with respect to changes in the domain wall $P$.

Proposition 2.1.4. Suppose $H$ satisfies (H0), $\varphi \in \mathfrak{S}\left(0,1 ; \tilde{E}_{1}, \tilde{E}_{2}\right)$, and $P_{1}, P_{2} \in \mathfrak{S}(0,1)$
with $P_{j}=P_{j}(x)$. Then

$$
\sigma_{I}\left(H, P_{1}, \varphi\right)=\sigma_{I}\left(H, P_{2}, \varphi\right)
$$

Proof. Using Lemma 2.1.2, we have

$$
\sigma_{I}\left(H, P_{2}, \varphi\right)-\sigma_{I}\left(H, P_{1}, \varphi\right)=\operatorname{Tr} i\left[\Psi(H), P_{2}-P_{1}\right] \varphi^{\prime}(H) .
$$

Since $P_{2}-P_{1} \in\langle x\rangle^{-\infty}$, the assumption (H0) implies that $\left(P_{2}-P_{1}\right) \varphi^{\prime}(H)$ is trace-class. Therefore,

$$
\begin{aligned}
\operatorname{Tr} i\left[\Psi(H), P_{2}-P_{1}\right] \varphi^{\prime}(H) & =\operatorname{Tr} i \Psi(H)\left(P_{2}-P_{1}\right) \varphi^{\prime}(H)-\operatorname{Tr} i\left(P_{2}-P_{1}\right) \Psi(H) \varphi^{\prime}(H) \\
& =\operatorname{Tr} i\left(P_{2}-P_{1}\right) \varphi^{\prime}(H) \Psi(H)-\operatorname{Tr} i\left(P_{2}-P_{1}\right) \Psi(H) \varphi^{\prime}(H)=0,
\end{aligned}
$$

where we have used cyclicity of the trace to justify the second equality, and the fact that $\left[\varphi^{\prime}(H), \Psi(H)\right]=0$ for the last equality.

We have shown that if $H$ satisfies (H0), then $\sigma_{I}(H, P, \varphi)$ is independent of $P(x)=P \in$ $\mathfrak{S}(0,1)$ and $\varphi \in \mathfrak{S}\left(0,1 ; E_{1}, E_{2}\right)$. Now we want to analyze the stability of $\sigma_{I}(H, P, \varphi)$ with respect to perturbations of $H$. For $W$ a symmetric linear operator (with various additional assumptions in the results below), let

$$
H^{(\mu)}=H+\mu W \quad \text { for } \quad \mu \in[0,1] .
$$

We begin by introducing a class of appropriately decaying perturbations under which the interface conductivity is stable.

Theorem 2.1.5. Suppose $H$ satisfies (H0), $P(x)=P \in \mathfrak{S}(0,1)$ and $\varphi \in \mathfrak{S}\left(0,1 ; \tilde{E}_{1}, \tilde{E}_{2}\right)$. Assume that $W$ is symmetric with $W \in \operatorname{Op}\left(S_{1,0}^{m} \cap S\left(\langle x, y\rangle^{-\delta}\right)\right)$ for some $\delta>0$. If $H^{(1)}$
satisfies (H0), then $\sigma_{I}\left(H^{(1)}, P, \varphi\right)=\sigma_{I}(H, P, \varphi)$.
Proof. The assumption that $H^{(1)}$ satisfies (H0) implies $\sigma_{I}\left(H^{(1)}, P, \varphi\right)$ is well defined (see Lemma 2.1.1), with

$$
\begin{aligned}
& \sigma_{I}\left(H^{(1)}, P, \varphi\right)-\sigma_{I}(H, P, \varphi)= \\
& \quad \operatorname{Tr} i\left[\Psi\left(H^{(1)}\right), P\right]\left(\varphi^{\prime}\left(H^{(1)}\right)-\varphi^{\prime}(H)\right)+\operatorname{Tr} i\left[\Psi\left(H^{(1)}\right)-\Psi(H), P\right] \varphi^{\prime}(H)
\end{aligned}
$$

by Lemma 2.1.2. Using cyclicity of the trace as in the proof of Proposition 2.1.3, we find that

$$
\operatorname{Tr} i\left[\Psi\left(H^{(1)}\right)-\Psi(H), P\right] \varphi^{\prime}(H)=-\operatorname{Tr} i\left[\varphi^{\prime}(H), P\right]\left(\Psi\left(H^{(1)}\right)-\Psi(H)\right)
$$

Thus by Proposition 2.1.4, it suffices to show that $\operatorname{Tr}\left[A, P_{x_{0}}\right] B \rightarrow 0$ as $x_{0} \rightarrow \infty$ for $A \in$ $\left\{\Psi\left(H^{(1)}\right), \varphi^{\prime}(H)\right\}$ and $B \in\left\{\varphi^{\prime}\left(H^{(1)}\right)-\varphi^{\prime}(H), \Psi\left(H^{(1)}\right)-\Psi(H)\right\}$, where $P_{x_{0}}(x):=P\left(x-x_{0}\right)$.

Fix $\varepsilon>0$. For any $\phi \in \mathcal{C}_{c}^{\infty}\left(\tilde{E}_{1}, \tilde{E}_{2}\right)$, the Helffer-Sjöstrand formula implies that

$$
\phi\left(H^{(1)}\right)-\phi(H)=\frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\phi}(z)\left(z-H^{(1)}\right)^{-1} W(z-H)^{-1} d^{2} z
$$

Propositions A.2.1 and A.3.1 (together with the rapid decay of $\bar{\partial} \tilde{\phi}$ near the real axis) then imply that

$$
\begin{equation*}
\phi\left(H^{(1)}\right)-\phi(H) \in \operatorname{Op}\left(S\left(\langle x, y\rangle^{-\delta}\langle\xi, \zeta\rangle^{-m}\right)\right) \subset \operatorname{Op}\left(S\left(\langle x, y, \xi, \zeta\rangle^{-\delta}\right)\right) \tag{2.1.2}
\end{equation*}
$$

where we have assumed without loss of generality that $\delta<m$. Thus $B \in \operatorname{Op}\left(S\left(\langle x, y, \xi, \zeta\rangle^{-\delta}\right)\right)$, so there exist $b_{0} \in \mathcal{C}_{c}^{\infty}\left(\mathbb{R}^{4}\right)$ and $b_{1} \in S(1)$ as small as necessary such that $B=B_{0}+B_{1}$ with $B_{j}:=\operatorname{Op}\left(b_{j}\right)$ and $\left\|B_{1}\right\|<\varepsilon$. Writing $\left[A, P_{x_{0}}\right]=\left(1-P_{x_{0}}\right) A P_{x_{0}}-P_{x_{0}} A\left(1-P_{x_{0}}\right)$, Lemma 2.1.6 and the composition calculus imply that $\left[A, P_{x_{0}}\right] \in \operatorname{Op}\left(S\left(\left\langle x-x_{0}, y, \xi, \zeta\right\rangle^{-\infty}\right)\right)$ uniformly in
$x_{0}$. We conclude that $\left\|\left[A, P_{x_{0}}\right]\right\|_{1} \leq C$ uniformly in $x_{0}$, hence $\limsup _{x_{0} \rightarrow \infty}\left\|\left[A, P_{x_{0}}\right] B_{1}\right\|_{1}<$ $C \varepsilon$. Moreover, the decay of $b_{0}$ implies that

$$
\left\|\left[A, P_{x_{0}}\right] B_{0}\right\|_{1} \leq C \int_{\mathbb{R}^{4}}\left\langle x-x_{0}, y, \xi, \zeta\right\rangle^{-5}\langle x, y, \xi, \zeta\rangle^{-5} d x d y d \xi d \zeta \longrightarrow 0
$$

as $x_{0} \rightarrow \infty$. We have thus shown that

$$
\limsup _{x_{0} \rightarrow \infty}\left\|\left[A, P_{x_{0}}\right] B\right\|_{1} \leq \limsup _{x_{0} \rightarrow \infty}\left\|\left[A, P_{x_{0}}\right] B_{0}\right\|_{1}+\limsup _{x_{0} \rightarrow \infty}\left\|\left[A, P_{x_{0}}\right] B_{1}\right\|_{1}<C \varepsilon .
$$

Since $\varepsilon$ was arbitrary, the proof is complete.
Next, we introduce a class of relatively compact perturbations $W$ for which the symbols of compactly supported functionals of $H^{(1)}$ decay rapidly in $\langle y, \xi, \zeta\rangle$.

Lemma 2.1.6. Suppose $H$ satisfies (H0), and let $W$ be a symmetric pseudo-differential operator such that $W \in \operatorname{Op}\left(S_{1,0}^{m} \cap S\left(\langle\xi, \zeta\rangle^{m-\delta}\langle x, y\rangle^{-\delta}\right)\right)$ for some $\delta>0$. Then $H^{(1)} \in$ $\operatorname{Op}\left(E S_{1,0}^{m}\right)$ is self-adjoint with domain of definition $\mathcal{H}^{m}$, and $\phi\left(H^{(1)}\right) \in \operatorname{Op}\left(S\left(\langle y, \xi, \zeta\rangle^{-\infty}\right)\right)$ for any $\phi \in \mathcal{C}_{c}^{\infty}\left(\tilde{E}_{1}, \tilde{E}_{2}\right)$.

Proof. That $H^{(1)} \in \operatorname{Op}\left(E S_{1,0}^{m}\right)$ follows immediately from ellipticity of $H$ and the decay of $W$. By Proposition A.3.1, this means $H^{(1)}$ is self-adjoint with domain of definition $\mathcal{H}^{m}$. Moreover, $\left(z-H^{(1)}\right)^{-1} \in \operatorname{Op}\left(S\left(\langle\xi, \zeta\rangle^{-m}\right)\right)$ for all $\Im z \neq 0$, with symbolic bounds blowing up at worst algebraically as $\Im z \rightarrow 0$ (see Proposition A.3.1).

Let $\phi, \phi_{0} \in \mathcal{C}_{c}^{\infty}\left(\tilde{E}_{1}, \tilde{E}_{2}\right)$ such that $\phi \phi_{0}=\phi$, and define $\Theta:=\phi\left(H^{(1)}\right)-\phi(H)$ and $\Theta_{0}:=\phi_{0}\left(H^{(1)}\right)-\phi_{0}(H)$. Then $\Theta=\Theta \Theta_{0}+\phi(H) \Theta_{0}+\Theta \phi_{0}(H)$, and hence

$$
\Theta\left(1-\Theta_{0}\right)=\phi(H) \Theta_{0}+\Theta \phi_{0}(H)
$$

By (2.1.2), we have $\Theta, \Theta_{0} \in \operatorname{Op}\left(S\left(\langle x, y, \xi, \zeta\rangle^{-\delta}\right)\right)$. Thus there exist $\theta_{00} \in \mathcal{C}_{c}^{\infty}\left(\mathbb{R}^{4}\right)$ and $\theta_{01} \in S(1)$ as small as necessary such that $\Theta_{00}=\operatorname{Op}\left(\theta_{00}\right), \Theta_{01}=\operatorname{Op}\left(\theta_{01}\right)$ with $\left\|\Theta_{01}\right\|<1$
(see Proposition A.2.2), and $\Theta_{0}=\Theta_{00}+\Theta_{01}$. Since $\phi(H), \phi_{0}(H) \in \operatorname{Op}\left(S\left(\langle y, \xi, \zeta\rangle^{-\infty}\right)\right.$ ) by assumption on $H$, it follows that

$$
\Theta\left(1-\Theta_{01}\right)=\Theta\left(1-\Theta_{0}\right)+\Theta \Theta_{00} \in \operatorname{Op}\left(S\left(\langle y, \xi, \zeta\rangle^{-\infty}\right)\right)
$$

hence

$$
\Theta=\left(\Theta\left(1-\Theta_{0}\right)+\Theta \Theta_{00}\right)\left(1-\Theta_{01}\right)^{-1} \in \operatorname{Op}\left(S\left(\langle y, \xi, \zeta\rangle^{-\infty}\right)\right)
$$

We conclude that $\phi\left(H^{(1)}\right)=\phi(H)+\Theta \in \operatorname{Op}\left(S\left(\langle y, \xi, \zeta\rangle^{-\infty}\right)\right)$, and the proof is complete.

We now show that the interface conductivity is stable with respect to this class of relatively compact perturbations $W$.

Theorem 2.1.7. Suppose $H$ satisfies (H0), $P(x)=P \in \mathfrak{S}(0,1)$ and $\varphi \in \mathfrak{S}\left(0,1 ; \tilde{E}_{1}, \tilde{E}_{2}\right)$. If $W$ is symmetric with $W \in \operatorname{Op}\left(S_{1,0}^{m} \cap S\left(\langle\xi, \zeta\rangle^{m-\delta}\langle x, y\rangle^{-\delta}\right)\right)$ for some $\delta>0$, then

$$
\sigma_{I}\left(H^{(1)}, P, \varphi\right)=\sigma_{I}(H, P, \varphi)
$$

Proof. Fix $\phi \in \mathcal{C}_{c}^{\infty}\left(\tilde{E}_{1}, \tilde{E}_{2}\right)$ such that $\phi \equiv 1$ in some open interval containing $\operatorname{supp}\left(\varphi^{\prime}\right)$. Lemma 2.1.6 then implies that $H^{(1)}$ satisfies (H0), with $\Phi$ replaced by $\phi$. The result then follows from Theorem 2.1.5.

Next, we derive a stability result that no longer requires the perturbation to be relatively compact. We will instead assume that $W$ is relatively bounded with respect to $H$, and require that $W$ be "sufficiently small."

Theorem 2.1.8. Suppose $H$ satisfies (H0), $P(x)=P \in \mathfrak{S}(0,1)$ and $\varphi \in \mathfrak{S}\left(0,1 ; \tilde{E}_{1}, \tilde{E}_{2}\right)$. If $W$ is symmetric with $W \in \mathrm{Op}\left(S_{1,0}^{m}\right)$, then $\sigma_{I}\left(H^{(\mu)}, P, \varphi\right)=\sigma_{I}(H, P, \varphi)$ for all $\mu>0$ sufficiently small.

In the proof below, we will be analyzing operators $A=\operatorname{Op}(a)$ that depend on the parameter $\mu$. For $\mathfrak{m}: \mathbb{R}^{4} \rightarrow[0, \infty)$ an order function, we will write $A \in \operatorname{Op}(\mu S(\mathfrak{m}))$ to mean that $a \in S(\mathfrak{m})$ for all $\mu$, with $\left|\partial^{\alpha} a\right| \leq C_{\alpha} \mu \mathfrak{m}$ uniformly in $\mu$.

Proof. Since $H \in \operatorname{Op}\left(E S_{1,0}^{m}\right)$ and $W$ is symmetric, it follows that $H^{(\mu)} \in \operatorname{Op}\left(E S_{1,0}^{m}\right)$ is self-adjoint (with domain of definition $\mathcal{H}^{m}$ ) whenever $\mu>0$ is sufficiently small.

Let $\phi, \phi_{0} \in \mathcal{C}_{c}^{\infty}\left(E_{1}, E_{2}\right)$ such that $\phi \phi_{0}=\phi$. Following the proof of Lemma 2.1.6, we define $\Theta:=\phi\left(H^{(\mu)}\right)-\phi(H)$ and $\Theta_{0}:=\phi_{0}\left(H^{(\mu)}\right)-\phi_{0}(H)$, and verify that

$$
\Theta\left(1-\Theta_{0}\right)=\phi(H) \Theta_{0}+\Theta \phi_{0}(H)
$$

The Helffer-Sjöstrand formula implies that

$$
\Theta_{0}=-\frac{\mu}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\phi}_{0}(z)\left(z-H^{(\mu)}\right)^{-1} W(z-H)^{-1} d^{2} z
$$

By Propositions A.2.1 and A.3.2, this means $\Theta_{0} \in \operatorname{Op}\left(\mu S\left(\langle\xi, \zeta\rangle^{-m}\right)\right.$ ). It follows that (1-$\left.\Theta_{0}\right)^{-1} \in \operatorname{Op}(S(1))$ whenever $\mu>0$ is sufficiently small, hence

$$
\Theta=\left(\phi(H) \Theta_{0}+\Theta \phi_{0}(H)\right)\left(1-\Theta_{0}\right)^{-1} \in \operatorname{Op}\left(S\left(\langle y, \xi, \zeta\rangle^{-\infty}\right)\right)
$$

by (H0). We conclude that $\phi\left(H^{(\mu)}\right)=\phi(H)+\Theta \in \operatorname{Op}\left(S\left(\langle y, \xi, \zeta\rangle^{-\infty}\right)\right)$. Since $\phi$ was arbitrary, this means $\sigma_{I}\left(H^{(\mu)}, P, \varphi\right)$ is well defined whenever $\mu>0$ is small enough.

Let $\chi \in \mathfrak{S}(1,0 ; 1,2)$ be monotonically non-increasing, and define

$$
\chi_{\varepsilon}(x, y, \xi, \zeta):=\chi(\varepsilon|(x, y, \xi, \zeta)|)
$$

and $H^{(\mu, \varepsilon)}:=\operatorname{Op}\left(\sigma+\mu \chi_{\varepsilon} w\right)$ for $\varepsilon \in(0,1]$. Since $\chi_{\varepsilon} w \in S\left(\langle x, y, \xi, \zeta\rangle^{-\infty}\right)$, Theorem 2.1.7
implies that

$$
\sigma_{I}\left(H^{(\mu, \varepsilon)}, P, \varphi\right)=\sigma_{I}(H, P, \varphi), \quad \varepsilon \in(0,1]
$$

It remains to show that $\sigma_{I}\left(H^{(\mu, \varepsilon)}, P, \varphi\right)-\sigma_{I}\left(H^{(\mu)}, P, \varphi\right) \rightarrow 0$ as $\varepsilon \downarrow 0$. In order to do this, we will first need symbolic bounds that are uniform in $\mu$ and $\varepsilon$.

For any multi-index $\alpha$ and any $p>0$, we have $\left|\partial^{\alpha} \chi \varepsilon\right|(x, y, \xi, \zeta) \leq C_{\alpha, p}|\alpha|\langle\varepsilon x, \varepsilon y, \varepsilon \xi, \varepsilon \zeta\rangle^{-p}$ uniformly in $\varepsilon$. In particular, this means

$$
\left|\partial^{\alpha} \chi_{\varepsilon}\right|(x, y, \xi, \zeta) \leq C_{\alpha,|\alpha|^{\varepsilon^{|\alpha|}}\langle\varepsilon x, \varepsilon y, \varepsilon \xi, \varepsilon \zeta\rangle^{-|\alpha|} \leq C_{\alpha,|\alpha|}\langle x, y, \xi, \zeta\rangle^{-|\alpha|}, \quad \varepsilon \in(0,1] . . ~ . ~}^{\text {. }} . \quad \text {. }
$$

By definition, $\operatorname{Op}(w):=W$ satisfies $\left|\partial_{x, y}^{\beta_{1}} \partial_{\xi, \zeta}^{\beta_{2}} w\right|(x, y, \xi, \zeta) \leq C_{\beta}\langle\xi, \zeta\rangle^{m-\left|\beta_{2}\right|}$ for any multiindex $\beta=\left(\beta_{1}, \beta_{2}\right)$. It follows that

$$
\begin{equation*}
\left|\partial^{\alpha} \chi_{\varepsilon} \partial^{\beta} w\right|(x, y, \xi, \zeta) \leq C_{\alpha, \beta}\langle x, y, \xi, \zeta\rangle^{-|\alpha|}\langle\xi, \zeta\rangle^{m-\left|\beta_{2}\right|} \leq C_{\alpha, \beta}\langle\xi, \zeta\rangle^{m-\left|\alpha_{2}\right|-\left|\beta_{2}\right|} \tag{2.1.3}
\end{equation*}
$$

uniformly in $\varepsilon \in(0,1]$, hence the set $\mathcal{W}:=\left\{\chi_{\varepsilon} w: \varepsilon \in(0,1]\right\}$ satisfies the assumptions of Proposition A.3.2. It follows that $\left(z-H^{(\mu, \varepsilon)}\right)^{-1} \in \mathrm{Op}\left(S\left(\langle\xi, \zeta\rangle^{-m}\right)\right)$ with all symbolic bounds uniform in $\varepsilon \in(0,1]$ and blowing up algebraically as $\Im z \rightarrow 0$. Recall that $H^{(\mu)} \in \operatorname{Op}\left(E S_{1,0}^{m}\right)$ whenever $\mu>0$ is sufficiently small, thus the same can be said of the ( $\varepsilon$-independent) operator $\left(z-H^{(\mu)}\right)^{-1}$. With $\Theta^{(\varepsilon)}:=\phi\left(H^{(\mu, \varepsilon)}\right)-\phi\left(H^{(\mu)}\right)$ and $W_{\varepsilon}:=\operatorname{Op}\left(\left(1-\chi_{\varepsilon}\right) w\right)$, the Helffer-Sjöstrand formula and Proposition A.2.1 then imply that

$$
\begin{equation*}
\Theta^{(\varepsilon)}=\frac{\mu}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\phi}(z)\left(z-H^{(\mu, \varepsilon)}\right)^{-1} W_{\varepsilon}\left(z-H^{(\mu)}\right)^{-1} d^{2} z \in \operatorname{Op}\left(\mu S\left(\langle\xi, \zeta\rangle^{-m}\right)\right) \tag{2.1.4}
\end{equation*}
$$

uniformly in $\mu>0$ sufficiently small and $\varepsilon \in(0,1]$. Since $\phi$ is arbitrary, we also have that $\Theta_{0}^{(\varepsilon)}:=\phi_{0}\left(H^{(\mu, \varepsilon)}\right)-\phi_{0}\left(H^{(\mu)}\right) \in \operatorname{Op}\left(\mu S\left(\langle\xi, \zeta\rangle^{-m}\right)\right)$ uniformly in $\mu$ and $\varepsilon$. Thus if $\mu>0$ is small enough, then $\left(1-\Theta_{0}^{(\varepsilon)}\right)^{-1} \in \operatorname{Op}(S(1))$ uniformly in $\varepsilon \in(0,1]$. Since $\phi\left(H^{(\mu)}\right) \in$
$\operatorname{Op}\left(S\left(\langle y, \xi, \zeta\rangle^{-\infty}\right)\right)$, the familiar identity

$$
\Theta^{(\varepsilon)}=\left(\phi\left(H^{(\mu)}\right) \Theta_{0}^{(\varepsilon)}+\Theta^{(\varepsilon)} \phi_{0}\left(H^{(\mu)}\right)\right)\left(1-\Theta_{0}^{(\varepsilon)}\right)^{-1}
$$

implies that $\phi\left(H^{(\mu, \varepsilon)}\right)=\Theta^{(\varepsilon)}+\phi\left(H^{(\mu)}\right) \in \operatorname{Op}\left(S\left(\langle y, \xi, \zeta\rangle^{-\infty}\right)\right)$ uniformly in $\varepsilon \in(0,1]$.
As in the proof of Theorem 2.1.7, cyclicity of the trace implies that

$$
\begin{aligned}
& \sigma_{I}\left(H^{(\mu, \varepsilon)}, P, \varphi\right)-\sigma_{I}\left(H^{(\mu)}, P, \varphi\right)= \\
& \quad \operatorname{Tr} i\left[\Psi\left(H^{(\mu, \varepsilon)}\right), P\right]\left(\varphi^{\prime}\left(H^{(\mu, \varepsilon)}\right)-\varphi^{\prime}\left(H^{(\mu)}\right)\right)-\operatorname{Tr} i\left[\varphi^{\prime}\left(H^{(\mu)}\right), P\right]\left(\Psi\left(H^{(\mu, \varepsilon)}\right)-\Psi\left(H^{(\mu)}\right)\right) .
\end{aligned}
$$

Thus it suffices to show that $\operatorname{Tr}\left[A_{\varepsilon}, P\right] B_{\varepsilon} \rightarrow 0$ as $\varepsilon \downarrow 0$, for $A_{\varepsilon} \in\left\{\varphi^{\prime}\left(H^{(\mu)}\right), \Psi\left(H^{(\mu, \varepsilon)}\right)\right\}$ and $B_{\varepsilon} \in\left\{\varphi^{\prime}\left(H^{(\mu, \varepsilon)}\right)-\varphi^{\prime}\left(H^{(\mu)}\right), \Psi\left(H^{(\mu, \varepsilon)}\right)-\Psi\left(H^{(\mu)}\right)\right\}$. In the paragraph above, we showed that $A_{\varepsilon} \in \operatorname{Op}\left(S\left(\langle y, \xi, \zeta\rangle^{-\infty}\right)\right)$ uniformly in $\varepsilon \in(0,1]$. The composition calculus then implies that $\left[A_{\varepsilon}, P\right] \in \operatorname{Op}\left(S\left(\langle x, y, \xi, \zeta\rangle^{-\infty}\right)\right)$ uniformly in $\varepsilon \in(0,1]$. Using (2.1.4) and the paragraph above it, we see that $B_{\varepsilon} \in \operatorname{Op}\left(S\left(1-\chi_{\varepsilon}(x, y, \xi, \zeta)+\varepsilon\right)\right)$ uniformly in $\varepsilon \in(0,1]$. We thus have $\left[A_{\varepsilon}, P\right] B_{\varepsilon} \in \operatorname{Op}\left(S\left(\langle x, y, \xi, \zeta\rangle^{-5}\left(1-\chi_{\varepsilon}(x, y, \xi, \zeta)+\varepsilon\right)\right)\right)$ uniformly in $\varepsilon \in(0,1]$, meaning that (see appendix A.1)

$$
\begin{aligned}
\left\|\left[A_{\varepsilon}, P\right] B_{\varepsilon}\right\|_{1} & \leq C \int_{\mathbb{R}^{4}}\langle x, y, \xi, \zeta\rangle^{-5}\left(1-\chi_{\varepsilon}(x, y, \xi, \zeta)+\varepsilon\right) d x d y d \xi d \zeta \\
& \leq C\left(\int_{\left\{|(x, y, \xi, \zeta)| \geq \varepsilon^{-1}\right\}}\langle x, y, \xi, \zeta\rangle^{-5} d x d y d \xi d \zeta+\varepsilon \int_{\mathbb{R}^{4}}\langle x, y, \xi, \zeta\rangle^{-5} d x d y d \xi d \zeta\right) \\
& \leq C \varepsilon
\end{aligned}
$$

This completes the proof.

Theorem 2.1.8 implies the following

Corollary 2.1.9. Let $P(x)=P \in \mathfrak{S}(0,1)$ and $\varphi \in \mathfrak{S}\left(0,1 ; \tilde{E}_{1}, \tilde{E}_{2}\right)$. Suppose $W \in \operatorname{Op}\left(S_{1,0}^{m}\right)$
is symmetric such that $H^{(\mu)}$ satisfies (H0) for all $\mu \in[0,1]$. Then

$$
\sigma_{I}\left(H^{(1)}, P, \varphi\right)=\sigma_{I}(H, P, \varphi)
$$

Proof. For any $\mu \in[0,1]$, Theorem 2.1.8 implies that $\sigma_{I}\left(H^{\left(\mu^{\prime}\right)}, P, \varphi\right)$ is constant over $\mu^{\prime}$ in some small neighborhood of $\mu$. Thus $\sigma_{I}\left(H^{(\mu)}, P, \varphi\right)$ is independent of $\mu \in[0,1]$, and the result is complete.

We now turn our attention to operators $H$ satisfying (H1) in Section 1. First, we show that the latter is a stronger assumption than (H0), meaning that all the previous results obtained in this section still apply.

Proposition 2.1.10. Suppose $H=\operatorname{Op}(\sigma)$ satisfies (H1) and $\Phi \in \mathcal{C}_{c}^{\infty}\left(E_{1}, E_{2}\right)$. Define $H_{h}:=\operatorname{Op}_{h}(\sigma)$ for $h \in(0,1]$. Then $\Phi\left(H_{h}\right) \in \mathrm{Op}_{h}\left(S\left(\langle y, \xi, \zeta\rangle^{-\infty}\right)\right)$.

Proof. We observe that $H_{h}=\mathrm{Op}_{1}(\sigma(x, y, h \xi, h \zeta))$ satisfies (H1) for all $0<h \leq 1$, with $\sigma_{ \pm}(h \xi, h \zeta)$ replacing $\sigma_{ \pm}(\xi, \zeta)$. This means $H_{h}$ is self-adjoint with domain $\mathcal{H}^{m}$ that is dense in $\mathcal{H}$, and $\left(i+H_{h}\right)^{-1} \in \operatorname{Op}_{h}\left(S\left(\langle\xi, \zeta\rangle^{-m}\right)\right)$ as recalled in Proposition A.3.1 below. Writing $\Phi\left(H_{h}\right)=\left(i+H_{h}\right)^{-N} \Phi_{N}\left(H_{h}\right)$ for $\Phi_{N}$ still compactly supported, we obtain that $\Phi\left(H_{h}\right) \in$ $\mathrm{Op}_{h}\left(S\left(\langle\xi, \zeta\rangle^{-\infty}\right)\right)$.

We now prove decay in $y$. Since $\sigma_{ \pm}$are independent of $(x, y)$ and have a spectral gap in $\left(E_{1}, E_{2}\right)$, it follows that $H_{ \pm, h}:=\operatorname{Op}_{h}\left(\sigma_{ \pm}\right)$also have a spectral gap in $\left(E_{1}, E_{2}\right)$. Hence $\Phi\left(H_{ \pm, h}\right)=0$, meaning that

$$
\Phi\left(H_{h}\right)=\phi(y)\left(\Phi\left(H_{h}\right)-\Phi\left(H_{+, h}\right)\right)+(1-\phi(y))\left(\Phi\left(H_{h}\right)-\Phi\left(H_{-, h}\right)\right)
$$

where we assume $\phi \in \mathfrak{S}(0,1)$. By the Helffer-Sjöstrand Formula, we see that

$$
\Phi\left(H_{h}\right)-\Phi\left(H_{+, h}\right)=-\frac{1}{\pi} \int_{Z} \bar{\partial} \tilde{\Phi}(z)\left(z-H_{h}\right)^{-1}\left(H_{h}-H_{+, h}\right)\left(z-H_{+, h}\right)^{-1} d^{2} z
$$

with $H_{h}-H_{+, h}=\mathrm{Op}_{h}\left(\sigma-\sigma_{+}\right)$and $\sigma-\sigma_{+}$vanishing for $y>0$ sufficiently large. By Propositions A.2.1 and A.3.1 (and the rapid decay of $\bar{\partial} \tilde{\Phi}$ near the real axis), it follows that $\Phi\left(H_{h}\right)-\Phi\left(H_{+, h}\right) \in \operatorname{Op}_{h}\left(S\left(\left\langle y_{+}\right\rangle^{-\infty}\right)\right)$. Since $\phi(y)=\phi \in \operatorname{Op}_{h}\left(S\left(\left\langle y_{-}\right\rangle^{-\infty}\right)\right)$, the composition calculus implies that $\phi(y)\left(\Phi\left(H_{h}\right)-\Phi\left(H_{+, h}\right)\right) \in \mathrm{Op}_{h}\left(\langle y\rangle^{-\infty}\right)$. A parallel argument proves that $(1-\phi(y))\left(\Phi\left(H_{h}\right)-\Phi\left(H_{-, h}\right)\right) \in \mathrm{Op}_{h}\left(\langle y\rangle^{-\infty}\right)$, hence $\Phi\left(H_{h}\right) \in \mathrm{Op}_{h}\left(\langle y\rangle^{-\infty}\right)$. The result then follows from interpolation.

We now show that under (H1), the interface conductivity is stable with respect to semiclassical rescaling.

Theorem 2.1.11. Suppose $H=\mathrm{Op}(\sigma)$ satisfies (H1), and define $H_{h}:=\mathrm{Op}_{h}(\sigma)$. Let $P(x)=P \in \mathfrak{S}(0,1)$ and $\varphi \in \mathfrak{S}\left(0,1 ; E_{1}, E_{2}\right)$. Then $\sigma_{I}\left(H_{h}, P, \varphi\right)=\sigma_{I}(H, P, \varphi)$ for all $h \in(0,1]$.

Proof. Fix $h \in(0,1]$. Since $H_{h}=\operatorname{Op}\left(\sigma_{h}\right)$ with $\sigma_{h}(x, y, \xi, \zeta)=\sigma(x, y, h \xi, h \zeta)$, we know that $H_{h}$ satisfies (H1). Observe also that for any $h^{\prime} \in(0,1]$, we have $H_{h}-H_{h^{\prime}} \in \operatorname{Op}\left(S_{1,0}^{m}\right)$. For $h^{\prime} \in(0,1]$ and $\mu \in[0,1]$, define $\sigma_{h, h^{\prime}}^{(\mu)}:=\sigma_{h}+\mu\left(\sigma_{h^{\prime}}-\sigma_{h}\right)$. By continuity and ellipticity of $\sigma$, we know that whenever $\left|h^{\prime}-h\right|$ is sufficiently small, $\sigma_{h, h^{\prime}}^{(\mu)} \in E S_{1,0}^{m}$ for all $\mu \in[0,1]$. Moreover, (H1) implies that $\sigma_{h, h^{\prime}}^{(\mu)}(x, y, \xi, \zeta)=\sigma_{h, h^{\prime}, \pm}^{(\mu)}(\xi, \zeta)$ whenever $\pm y$ is sufficiently large, where

$$
\sigma_{h, h^{\prime}, \pm}^{(\mu)}(\xi, \zeta):=\sigma_{ \pm}(h \xi, h \zeta)+\mu\left(\sigma_{ \pm}\left(h^{\prime} \xi h^{\prime} \zeta\right)-\sigma_{ \pm}(h \xi, h \zeta)\right)
$$

Since $\sigma_{ \pm}$have a spectral gap in $\left(E_{1}, E_{2}\right)$, it follows (also from $\sigma_{h, h^{\prime}}^{(\mu)} \in E S_{1,0}^{m}$ and continuity) that whenever $\left|h^{\prime}-h\right|$ is sufficiently small, $\sigma_{h, h^{\prime}, \pm}^{(\mu)}$ has a spectral gap in $\left(E_{1}+\Delta, E_{2}-\Delta\right)$ for all $\mu \in[0,1]$, where $\Delta:=\left(E_{2}-E_{1}\right) / 4$. Thus we have shown that whenever $\left|h^{\prime}-h\right|$ is sufficiently small and $\mu \in[0,1]$, the operator $H_{h, h^{\prime}}^{(\mu)}:=\operatorname{Op}\left(\sigma_{h, h^{\prime}}^{(\mu)}\right)$ satisfies (H1) with $\left(E_{1}, E_{2}\right)$ replaced by $\left(E_{1}+\Delta, E_{2}-\Delta\right)$. By Theorem 2.1.8, we conclude that whenever $\left|h^{\prime}-h\right|$ is sufficiently small, $\sigma_{I}\left(H_{h, h^{\prime}}^{(\mu)}, P, \varphi_{1}\right)$ is independent of $\mu \in[0,1]$, provided $\varphi_{1} \in \mathfrak{S}\left(0,1 ; E_{1}+\Delta, E_{2}-\Delta\right)$. In particular, this means $\sigma_{I}\left(H_{h^{\prime}}, P, \varphi_{1}\right)=\sigma_{I}\left(H_{h}, P, \varphi_{1}\right)$ if $\left|h^{\prime}-h\right|$ is sufficiently small. But
since $H_{h^{\prime}}$ satisfies (H1), we know by Proposition 2.1.3 that $\sigma_{I}\left(H_{h^{\prime}}, P, \varphi\right)=\sigma_{I}\left(H_{h}, P, \varphi\right)$. We have thus shown that $\sigma_{I}\left(H_{h^{\prime}}, P, \varphi\right)$ is constant over $h^{\prime}$ in some small neighborhood of $h$. Since $h \in(0,1]$ was arbitrary, the result is complete.

Finally, we show that the interface conductivity is immune to oscillations in the $x$-variable.

Theorem 2.1.12. Suppose $H=\mathrm{Op}(\sigma)$ satisfies (H1), $P(x)=P \in \mathfrak{S}(0,1)$ and $\varphi \in$ $\mathfrak{S}\left(0,1 ; E_{1}, E_{2}\right)$. Fix $x_{0} \in \mathbb{R}$ and define $\sigma_{x_{0}}(x, y, \xi, \zeta):=\sigma\left(x_{0}, y, \xi, \zeta\right)$ and $H_{x_{0}}:=\operatorname{Op}\left(\sigma_{x_{0}}\right)$. Then $\sigma_{I}\left(H_{x_{0}}, P, \varphi\right)=\sigma_{I}(H, P, \varphi)$.

Proof. Let $\chi \in \mathcal{C}_{c}^{\infty}(-2,2)$ such that $\chi(x)=1$ whenever $x \in[-1,1]$. Define $\chi_{\varepsilon}(x):=\chi(\varepsilon x)$. For $\varepsilon \in(0,1]$, define $\sigma^{(\varepsilon)}(x, y, \xi, \zeta):=\sigma\left(x-\left(x-x_{0}\right) \chi_{\varepsilon}\left(x-x_{0}\right), y, \xi, \zeta\right)$, so that $\sigma^{(\varepsilon)}=\sigma_{x_{0}}$ whenever $\left|x-x_{0}\right| \leq \varepsilon^{-1}$ and $\sigma^{(\varepsilon)}=\sigma$ whenever $\left|x-x_{0}\right| \geq 2 \varepsilon^{-1}$. Since $H^{(\varepsilon)}:=\operatorname{Op}\left(\sigma^{(\varepsilon)}\right)$ satisfies (H1) and $\sigma^{(\varepsilon)}-\sigma$ vanishes whenever $\langle x, y\rangle$ is sufficiently large, Theorem 2.1.5 implies that $\sigma_{I}\left(H^{(\varepsilon)}, P, \varphi\right)=\sigma_{I}(H, P, \varphi)$ for all $\varepsilon \in(0,1]$.

It remains to show that $\sigma_{I}\left(H^{(\varepsilon)}, P, \varphi\right)-\sigma_{I}\left(H_{x_{0}}, P, \varphi\right) \rightarrow 0$ as $\varepsilon \downarrow 0$. To do this, we will emulate the argument used to prove Theorem 2.1.8. In particular, it suffices to show that $\operatorname{Tr}\left[A_{\varepsilon}, P\right] B_{\varepsilon} \rightarrow 0$ as $\varepsilon \downarrow 0$, for $A_{\varepsilon} \in\left\{\phi\left(H_{x_{0}}\right), \phi\left(H^{(\varepsilon)}\right)\right\}, B_{\varepsilon}=\phi\left(H^{(\varepsilon)}\right)-\phi\left(H_{x_{0}}\right)$ and $\phi \in \mathcal{C}_{c}^{\infty}\left(E_{1}, E_{2}\right)$. As before, Proposition A.3.2 and the composition calculus imply that $\left[A_{\varepsilon}, P\right] \in \operatorname{Op}\left(S\left(\langle x, y, \xi, \zeta\rangle^{-\infty}\right)\right)$ uniformly in $\varepsilon \in(0,1]$. By the Helffer-Sjöstrand formula,

$$
B_{\varepsilon}=\frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\phi}(z)\left(z-H^{(\varepsilon)}\right)^{-1}\left(H_{x_{0}}-H^{(\varepsilon)}\right)\left(z-H_{x_{0}}\right)^{-1} d^{2} z
$$

Since $\sigma_{x_{0}}-\sigma^{(\varepsilon)}$ vanishes whenever $\left|x-x_{0}\right| \leq \varepsilon^{-1}$, it follows that $\sigma_{x_{0}}-\sigma^{(\varepsilon)} \in S\left(\left(1-\chi_{\varepsilon}(x-\right.\right.$ $\left.\left.x_{0}\right)+\varepsilon\right)\langle\xi, \zeta\rangle^{m}$ ) uniformly in $\varepsilon \in(0,1]$. By Propositions A.2.1 and A.3.2 (and the rapid decay of $\bar{\partial} \tilde{\phi}$ near the real axis), we conclude that $B_{\varepsilon} \in \operatorname{Op}\left(S\left(1-\chi_{\varepsilon}\left(x-x_{0}\right)+\varepsilon\right)\right)$ uniformly
in $\varepsilon \in(0,1]$. Therefore,

$$
\begin{aligned}
\left\|\left[A_{\varepsilon}, P\right] B_{\varepsilon}\right\|_{1} & \leq C \int_{\mathbb{R}^{4}}\langle x\rangle^{-2}\langle y, \xi, \zeta\rangle^{-4}\left(1-\chi_{\varepsilon}\left(x-x_{0}\right)+\varepsilon\right) d x d y d \xi d \zeta \\
& \leq C \int_{\mathbb{R}}\langle x\rangle^{-2}\left(1-\chi_{\varepsilon}\left(x-x_{0}\right)+\varepsilon\right) d x \\
& \leq C\left(\int_{\left\{\left|x-x_{0}\right| \geq \varepsilon^{-1}\right\}}\langle x\rangle^{-2} d x+\varepsilon \int_{\mathbb{R}}\langle x\rangle^{-2} d x\right) \leq C \varepsilon
\end{aligned}
$$

and the proof is complete.

### 2.2 Bulk-interface correspondence

In this section, we derive an integral representation of the interface conductivity $\sigma_{I}$ in (1.0.1) that takes the form of a Fedosov-Hörmander formula (2.2.7) as in [10, 11]. We then simplify the evaluation of this integral under additional assumptions on the Hamiltonian.

The formula (2.2.7) may then be interpreted as in [10] as bulk-difference invariant; see Corollary 2.2.2. That the interface conductivity is given as bulk-difference invariant (which are defined in a more general setting than a difference of bulk-invariants) is usually referred to as a bulk-interface (or bulk-boundary) correspondence; see [7, 10, 37, 40, 42, 49, 68]. Establishing such a correspondence is one of the major objectives of mathematical analyses of topological insulators as it relates the quantized asymmetric transport, a physical observation, to the topological properties of two bulk insulators.

### 2.2.1 Bulk-Interface correspondence and semiclassical calculus

The aim of this section is to write the conductivity as an integral involving only the symbol $\sigma$ of the Hamiltonian. As demonstrated by sections 2.2.2 and 2.3 below, this integral can be used to easily evaluate $\sigma_{I}$ in many cases of interest. Unless one can obtain a full spectral decomposition of the Hamiltonian as in [10], the integral and its simplifications represent
the most accessible formulas for the conductivity.
The main result of this section is Theorem 2.2.1 below. The formula (2.2.1) simplifies nicely in various cases of interest (see section 2.2.2), and also implies a bulk-interface correspondence (Corollary 2.2.2) and Fedosov-Hörmander formula (Corollary 2.2.3).

Theorem 2.2.1. Suppose $H=\operatorname{Op}(\tilde{\sigma})$ satisfies (H1), $P(x)=P \in \mathfrak{S}(0,1)$ and $\varphi \in$ $\mathfrak{S}\left(0,1 ; E_{1}, E_{2}\right)$. Let $\alpha \in\left(E_{1}, E_{2}\right)$, and let $R \subset \mathbb{R}^{3}$ be bounded with a piecewise smooth boundary $\partial R$. Fix $x_{0} \in \mathbb{R}$, and define $\sigma(y, \xi, \zeta):=\tilde{\sigma}\left(x_{0}, y, \xi, \zeta\right)$. Assume $R$ contains all points $(y, \xi, \zeta)$ where $\sigma(y, \xi, \zeta)$ has an eigenvalue of $\alpha$. Then defining $z:=\alpha+i \omega$ and $\sigma_{z}:=z-\sigma$, we have that

$$
\begin{equation*}
\sigma_{I}(H, P, \varphi)=\frac{i}{16 \pi^{3}} \int_{\partial R} \int_{-\infty}^{+\infty} \Theta d \omega d \Sigma, \quad \Theta:=\operatorname{tr} \varepsilon_{i j k} \sigma_{z}^{-1} \partial_{i} \sigma_{z} \sigma_{z}^{-1} \partial_{j} \sigma_{z} \sigma_{z}^{-1} \nu_{k} \tag{2.2.1}
\end{equation*}
$$

where $\nu$ is the unit vector (outwardly) normal to $\partial R, d \Sigma=d \Sigma(y, \xi, \zeta)$ is the Euclidean measure on $\partial R$, and $\varepsilon_{i j k}$ is the anti-symmetric tensor with $\varepsilon_{123}=1$ and the variables identified by $(1,2,3)=(y, \xi, \zeta)$.

The strategy for the proof is to use an asymptotic expansion in the semiclassical parameter $h$ and apply Theorem 2.1.11 to eliminate terms that are not $O(1)$. A similar technique is used in [10, 11].

Proof. By Proposition 2.1.3, $\sigma_{I}(H, P, \varphi)$ is independent of $\varphi \in \mathfrak{S}\left(0,1 ; E_{1}, E_{2}\right)$, thus we can take $\varphi^{\prime} \in \mathcal{C}_{c}^{\infty}\left(\alpha_{0}, \alpha\right)$ for some $\alpha_{0}>E_{1}$. Moreover, by Theorem 2.1.12, we can without loss of generality assume that $\tilde{\sigma}(x, y, \xi, \zeta)=\sigma(y, \xi, \zeta)$ for all $(x, y, \xi, \zeta) \in \mathbb{R}^{4}$. Theorem 2.1.11 states that with $H_{h}:=\operatorname{Op}_{h}(\tilde{\sigma}), \sigma_{I}\left(H_{h}, P, \varphi\right)$ is independent of $h \in(0,1]$. Thus we will expand $\sigma_{I}\left(H_{h}, P, \varphi\right)$ in powers of $h$ and ignore terms that are not $O(1)$. We will use the shorthand $\sigma_{I}:=\sigma_{I}\left(H_{h}, P, \varphi\right)$.

Let $\mathrm{Op}_{h} \nu_{h}:=\varphi^{\prime}\left(H_{h}\right)$. By Proposition A.3.3, we have

$$
\begin{equation*}
\nu_{h}+\frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\varphi}^{\prime}(z) \tilde{q}_{z, h} d^{2} z \in S^{-3 / 2}\left(\langle y, \xi, \zeta\rangle^{-\infty}\right) \tag{2.2.2}
\end{equation*}
$$

where

$$
\tilde{q}_{z, h}=\sigma_{z}^{-1}+\frac{i h}{2}\left\{\sigma_{z}^{-1}, \sigma_{z}\right\}_{\zeta, y} \sigma_{z}^{-1}, \quad\{a, b\}_{\zeta, y}:=\partial_{\zeta} a \partial_{y} b-\partial_{y} a \partial_{\zeta} b
$$

and the right-hand side of (2.2.2) is defined by (A.1.2). With $\mathrm{Op}_{h}\left(\kappa_{h}\right):=\left[H_{h}, P\right]$, we have that

$$
\kappa_{h}+i h k_{1}-\frac{h^{2}}{4} k_{2} \in S^{-3}\left(\langle x\rangle^{-\infty}\langle\xi, \zeta\rangle^{m}\right), \quad k_{1}:=\partial_{\xi} \sigma P^{\prime}(x), \quad k_{2}:=\partial_{\xi \xi} \sigma P^{\prime \prime}(x)
$$

Since $\nu_{h} \in S\left(\langle y, \xi, \zeta\rangle^{-\infty}\right)$ and $\kappa_{h} \in S^{-1}\left(\langle x\rangle^{-\infty}\langle\xi, \zeta\rangle^{m}\right)$, the composition calculus implies that

$$
\begin{equation*}
\kappa_{h} \sharp_{h} \nu_{h}-\kappa_{h} \nu_{h}+\frac{i h}{2}\left\{\kappa_{h}, \nu_{h}\right\} \in S^{-3}\left(\langle x, y, \xi, \zeta\rangle^{-\infty}\right), \tag{2.2.3}
\end{equation*}
$$

where

$$
\{a, b\}:=\partial_{\xi} a \partial_{x} b+\partial_{\zeta} a \partial_{y} b-\partial_{x} a \partial_{\xi} b-\partial_{y} a \partial_{\zeta} b
$$

Note that $S^{-3}$ (rather than $S^{-2}$ ) appears on the right-hand side of (2.2.3) because $\kappa_{h}$ is $O(h)$ in $S\left(\langle x\rangle^{-\infty}\langle\xi, \zeta\rangle^{m}\right)$. Therefore,

$$
\sigma_{I}=\frac{i}{(2 \pi h)^{2}} \operatorname{tr} \int_{\mathbb{R}^{4}} \kappa_{h} \sharp H_{h} d R_{4}=\frac{i}{(2 \pi h)^{2}} \operatorname{tr} \int_{\mathbb{R}^{4}}\left(\kappa_{h} \nu_{h}-\frac{i h}{2}\left\{\kappa_{h}, \nu_{h}\right\}\right) d R_{4}+o(1)
$$

as $h \rightarrow 0$, with $d R_{4}:=d x d y d \xi d \zeta$. Since

$$
\kappa_{h} \nu_{h}=\left(i h k_{1}-\frac{h^{2}}{4} k_{2}\right) \frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\varphi}^{\prime}(z)\left(\sigma_{z}^{-1}+\frac{i h}{2}\left\{\sigma_{z}^{-1}, \sigma_{z}\right\}_{\zeta, y} \sigma_{z}^{-1}\right) d^{2} z+h^{5 / 2} a_{h}
$$

with $a_{h} \in S\left(\langle x, y, \xi, \zeta\rangle^{-\infty}\right)$, and

$$
\left\{\kappa_{h}, \nu_{h}\right\}=\left\{i h k_{1}, \frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\varphi}^{\prime}(z) \sigma_{z}^{-1} d^{2} z\right\}+h^{2} b_{h}, \quad b_{h} \in S\left(\langle x, y, \xi, \zeta\rangle^{-\infty}\right)
$$

it follows that

$$
\begin{aligned}
\sigma_{I}=\frac{i}{(2 \pi h)^{2}} \frac{1}{\pi} \operatorname{tr} \int_{\mathbb{R}^{4}} \int_{\mathbb{C}} \bar{\partial} \tilde{\varphi}^{\prime}(z)\left(i h k_{1} \sigma_{z}^{-1}\right. & -\frac{h^{2}}{2} k_{1}\left\{\sigma_{z}^{-1}, \sigma_{z}\right\}_{\zeta, y} \sigma_{z}^{-1} \\
& \left.-\frac{h^{2}}{4} k_{2} \sigma_{z}^{-1}+\frac{h^{2}}{2}\left\{k_{1}, \sigma_{z}^{-1}\right\}\right) d^{2} z d R_{4}+o(1)
\end{aligned}
$$

as $h \rightarrow 0$. Since $\sigma_{I}$ is independent of $h$, it follows that the $O\left(h^{-1}\right)$ term above vanishes, and thus

$$
\begin{equation*}
\sigma_{I}=\frac{i}{(2 \pi)^{3}} \operatorname{tr} \int_{\mathbb{R}^{4}} \int_{\mathbb{C}} \bar{\partial} \tilde{\varphi}^{\prime}(z)\left(-k_{1} \sigma_{z}^{-1}\left\{\sigma_{z}, \sigma_{z}^{-1}\right\}_{\zeta, y}-\frac{1}{2} k_{2} \sigma_{z}^{-1}+\left\{k_{1}, \sigma_{z}^{-1}\right\}\right) d^{2} z d R_{4} \tag{2.2.4}
\end{equation*}
$$

Observe that whenever $(y, \xi, \zeta) \notin R, z \mapsto \sigma_{z}^{-1}$ is holomorphic and thus the above integral over $z$ vanishes (this is verified via an integration by parts in $\bar{\partial}$ ). Since $k_{1}$ and $k_{2}$ vanish whenever $|x|$ is sufficiently large, the integration region $\mathbb{R}^{4}$ in (2.2.4) can be replaced by the volume $I \times R$, with $I \subset \mathbb{R}$ a bounded interval.

Recalling that $\partial_{x} \sigma \equiv 0$, it follows that $\left\{k_{1}, \sigma_{z}^{-1}\right\}=P^{\prime}(x)\left\{\partial_{\xi} \sigma, \sigma_{z}^{-1}\right\}_{\zeta, y}-P^{\prime \prime}(x) \partial_{\xi} \sigma \partial_{\xi} \sigma_{z}^{-1}$ in (2.2.4). Using that $\int P^{\prime}=1$ and $\int P^{\prime \prime}=0$, we apply Fubini's Theorem and integrate
(2.2.4) in $x$ to obtain

$$
\begin{equation*}
\sigma_{I}=\frac{i}{(2 \pi)^{3}} \operatorname{tr} \int_{R} \int_{\mathbb{C}} \bar{\partial} \tilde{\varphi}^{\prime}(z)\left(\partial_{\xi} \sigma_{z} \sigma_{z}^{-1}\left\{\sigma_{z}, \sigma_{z}^{-1}\right\}_{\zeta, y}-\left\{\partial_{\xi} \sigma_{z}, \sigma_{z}^{-1}\right\}_{\zeta, y}\right) d^{2} z d R_{3} \tag{2.2.5}
\end{equation*}
$$

with $d R_{3}:=d y d \xi d \zeta$. At this point, we use the identity $\{a, b\}_{\zeta, y}=\partial_{\zeta}\left(a \partial_{y} b\right)-\partial_{y}\left(a \partial_{\zeta} b\right)$ and integration by parts in $(\zeta, y)$ to write

$$
\begin{equation*}
\int_{R} \int_{\mathbb{C}} \bar{\partial}_{\varphi^{\prime}}(z)\left\{\partial_{\xi} \sigma_{z}, \sigma_{z}^{-1}\right\}_{\zeta, y} d^{2} z d R_{3}=\int_{\partial R} \int_{\mathbb{C}} \bar{\partial} \tilde{\varphi}^{\prime}(z) \partial_{\xi} \sigma_{z}\left(\partial_{y} \sigma_{z}^{-1} \nu_{\zeta}-\partial_{\zeta} \sigma_{z}^{-1} \nu_{y}\right) d^{2} z d \Sigma \tag{2.2.6}
\end{equation*}
$$

Since $z \mapsto \sigma_{z}^{-1}$ is holomorphic whenever $(y, \xi, \zeta) \in \partial R$, an integration by parts in $\bar{\partial}$ reveals that the right-hand side of (2.2.6) vanishes. Thus we are left with

$$
\sigma_{I}=\frac{i}{(2 \pi)^{3}} \operatorname{tr} \int_{R} \int_{\mathbb{C}} \bar{\partial} \tilde{\varphi}^{\prime}(z) \partial_{\xi} \sigma_{z} \sigma_{z}^{-1}\left\{\sigma_{z}, \sigma_{z}^{-1}\right\}_{\zeta, y} d^{2} z d R_{3} .
$$

Integrating by parts in $\bar{\partial}$ with $z=: \lambda+i \omega$, we see that

$$
\sigma_{I}=\left.\frac{1}{2(2 \pi)^{3}} \operatorname{tr} \int_{R} \int_{\alpha_{0}}^{\alpha} \varphi^{\prime}(\lambda) \partial_{\xi} \sigma_{z} \sigma_{z}^{-1}\left\{\sigma_{z}, \sigma_{z}^{-1}\right\}_{\zeta, y}\right|_{\omega=0^{-}} ^{\omega=0^{+}} d \lambda d R_{3} .
$$

Since $\partial_{\xi} \sigma_{z} \sigma_{z}^{-1}\left\{\sigma_{z}, \sigma_{z}^{-1}\right\}_{\zeta, y} \rightarrow 0$ as $|\omega| \rightarrow \infty$, we have

$$
\left.\partial_{\xi} \sigma_{z} \sigma_{z}^{-1}\left\{\sigma_{z}, \sigma_{z}^{-1}\right\}_{\zeta, y}\right|_{\omega=0^{-}} ^{\omega=0^{+}}=-\int_{-\infty}^{+\infty} \partial_{\omega}\left(\partial_{\xi} \sigma_{z} \sigma_{z}^{-1}\left\{\sigma_{z}, \sigma_{z}^{-1}\right\}_{\zeta, y}\right) d \omega
$$

Cyclicity of the trace and the fact that $\partial_{\omega} \sigma_{z}=i$ imply that

$$
\operatorname{tr} \partial_{\omega}\left(\partial_{\xi} \sigma_{z} \sigma_{z}^{-1}\left\{\sigma_{z}, \sigma_{z}^{-1}\right\}_{\zeta, y}\right)=-i \operatorname{tr} \varepsilon_{i j k} \partial_{k}\left(\sigma_{z}^{-1} \partial_{i} \sigma_{z} \sigma_{z}^{-1} \partial_{j} \sigma_{z} \sigma_{z}^{-1}\right)
$$

where $\varepsilon_{i j k}$ is the anti-symmetric tensor with $\varepsilon_{123}=1$, and the variables are identified by
$(1,2,3)=(y, \xi, \zeta)$. Pulling $\partial_{k}$ out of the integral over $\omega$ and integrating by parts, we get

$$
\left.\operatorname{tr} \int_{R} \partial_{\xi} \sigma_{z} \sigma_{z}^{-1}\left\{\sigma_{z}, \sigma_{z}^{-1}\right\}_{\zeta, y}\right|_{\omega=0^{-}} ^{\omega=0^{+}} d R_{3}=i \int_{\partial R} \int_{-\infty}^{+\infty} \Theta d \omega d \Sigma
$$

where we recall the definition of $\Theta$ in (2.2.1). Thus we have shown that

$$
\sigma_{I}=\frac{i}{16 \pi^{3}} \int_{\left[\alpha_{0}, \alpha\right]} \varphi^{\prime}(\lambda) \int_{\partial R} \int_{-\infty}^{+\infty} \Theta d \omega d \Sigma d \lambda
$$

Integrating by parts in $\lambda$, we obtain

$$
\sigma_{I}=\frac{i}{16 \pi^{3}} \int_{\partial R} \int_{-\infty}^{+\infty} \Theta d \omega d \Sigma
$$

with now $z=\alpha+i \omega$ in the above integrand. The fact that only the boundary term survives follows from analyticity of $\Theta$ in $z$ over the region of integration (so that $\partial_{\lambda} \Theta=-i \partial_{\omega} \Theta$ ). This completes the proof.

Corollary 2.2.2. Suppose $H=\operatorname{Op}(\sigma)$ satisfies (H1), $P(x)=P \in \mathfrak{S}(0,1)$ and $\varphi \in$ $\mathfrak{S}\left(0,1 ; E_{1}, E_{2}\right)$. Let $\alpha \in\left(E_{1}, E_{2}\right)$ and define $\tau_{ \pm, z}:=z-\sigma_{ \pm}$with $z=\alpha+i \omega$. Then

$$
2 \pi \sigma_{I}(H, P, \varphi)=\frac{i}{8 \pi^{2}}\left(I_{+}-I_{-}\right), \quad I_{ \pm}=\int_{\mathbb{R}^{3}} \operatorname{tr}\left[\tau_{ \pm, z}^{-1} \partial_{\xi} \tau_{ \pm, z}, \tau_{ \pm, z}^{-1} \partial_{\zeta} \tau_{ \pm, z}\right] \tau_{ \pm, z}^{-1} d \omega d \xi d \zeta
$$

Proof. Fix $y_{0}>0$ sufficiently large, and define $R_{M}:=\left(-y_{0}, y_{0}\right) \times(-M, M)^{2}$ for all $M>0$. Observe that $|\Theta| \leq C\langle\omega\rangle^{-3}$ and $|\Theta| \leq C\langle\xi, \zeta\rangle^{-m-2}$ uniformly in $(y, \xi, \zeta) \in \partial R_{M}$ and $M>0$ sufficiently large, hence $|\Theta| \leq C\langle\omega\rangle^{-3 / 2}\langle\xi, \zeta\rangle^{-\frac{m+2}{2}}$ by interpolation. It follows that $\int_{-\infty}^{+\infty}|\Theta| d \omega \leq C\langle\xi, \zeta\rangle^{-\frac{m+2}{2}}$. Therefore, taking $R:=R_{M}$ in Theorem 2.2.1 and sending $M \rightarrow \infty$, we see that the contributions to $\sigma_{I}$ in (2.2.1) from the sides of $\partial R_{M}$ with normal vector in the $\xi$ and $\zeta$ directions vanish. Indeed, the area of these surfaces is proportional to $M$, with the maximum of the integrand bounded by $C M^{-1-m / 2}$. Thus we are left with integrals over the sides corresponding to $y= \pm y_{0}$, over which $\sigma(y, \xi, \zeta)=\sigma_{ \pm}(\xi, \zeta)$. As a
consequence, $\sigma_{I}=\frac{i}{16 \pi^{3}}\left(I_{+}-I_{-}\right)$, and the proof is complete.
Corollary 2.2.3. Suppose $H=\mathrm{Op}(\tilde{\sigma})$ satisfies (H1), $P(x)=P \in \mathfrak{S}(0,1)$ and $\varphi \in$ $\mathfrak{S}\left(0,1 ; E_{1}, E_{2}\right)$. Let $\alpha \in\left(E_{1}, E_{2}\right)$, and let $S \subset \mathbb{R}^{4}$ be bounded with a piecewise smooth boundary $\partial S$. Fix $x_{0} \in \mathbb{R}$ and define $\sigma(y, \xi, \zeta):=\tilde{\sigma}\left(x_{0}, y, \xi, \zeta\right)$. Assume $S$ contains all points $(0, y, \xi, \zeta)$ where $\sigma(y, \xi, \zeta)$ has an eigenvalue of $\alpha$. Then for $z:=\alpha+i \omega$ and $\sigma_{z}:=z-\sigma$, we have that

$$
\begin{equation*}
2 \pi \sigma_{I}(H, P, \varphi)=\frac{1}{24 \pi^{2}} \int_{\partial S} u \cdot \nu d \Sigma_{3}, \quad u_{l}:=\operatorname{tr} \epsilon_{i j k l} \partial_{i} \sigma_{z} \sigma_{z}^{-1} \partial_{j} \sigma_{z} \sigma_{z}^{-1} \partial_{k} \sigma_{z} \sigma_{z}^{-1} \tag{2.2.7}
\end{equation*}
$$

where $\nu$ is the outward unit normal to $\partial S$, $d \Sigma_{3}=d \Sigma_{3}(\omega, y, \xi, \zeta)$ is the Euclidean measure on $\partial S$, and $\epsilon_{i j k l}$ the anti-symmetric tensor with $\epsilon_{1234}=1$ and the variables identified by $(1,2,3,4)=(\omega, y, \xi, \zeta)$.

Observe that we can write the above integral in a more geometric form:

$$
\int_{\partial S} u \cdot \nu d \Sigma_{3}=\int_{\partial S}\left(\sigma_{z}^{-1} d \sigma_{z}\right)^{3}
$$

Proof. By Theorem 2.2.1,

$$
\begin{equation*}
\sigma_{I}=\frac{i}{16 \pi^{3}} \lim _{M \rightarrow \infty} \int_{\partial R} \int_{-M}^{M} \Theta d \omega d \Sigma \tag{2.2.8}
\end{equation*}
$$

Since $|\Theta| \leq C\langle\omega\rangle^{-3}$, it follows that if $z=\alpha \pm i M$ and $i, j, k \in\{\omega, y, \xi, \zeta\}$, then

$$
\begin{equation*}
\int_{R}\left|\operatorname{tr} \partial_{i} \sigma_{z} \sigma_{z}^{-1} \partial_{j} \sigma_{z} \sigma_{z}^{-1} \partial_{k} \sigma_{z} \sigma_{z}^{-1}\right| d R_{3} \longrightarrow 0 \tag{2.2.9}
\end{equation*}
$$

as $M \rightarrow \infty$. Since $\partial_{\omega} \sigma=i$, we know that

$$
\begin{equation*}
\Theta=-i \operatorname{tr} \varepsilon_{i j k} \partial_{\omega} \sigma \sigma_{z}^{-1} \partial_{i} \sigma_{z} \sigma_{z}^{-1} \partial_{j} \sigma_{z} \sigma_{z}^{-1} \nu_{k} \tag{2.2.10}
\end{equation*}
$$

It then follows from cyclicity of the trace that

$$
\begin{equation*}
\sigma_{I}=\frac{1}{48 \pi^{3}} \lim _{M \rightarrow \infty} \int_{\partial S_{M}} u \cdot \nu d \Sigma_{3}, \tag{2.2.11}
\end{equation*}
$$

where $S_{M}:=[-M, M] \times R$. Indeed, the integral over $[-M, M] \times \partial R \subset \partial S_{M}$ is precisely the integral on the right-hand side of (2.2.8), while the integral over the rest of $\partial S_{M}$ vanishes in the $M \rightarrow \infty$ limit by (2.2.9). The factor of 3 adjustment in (2.2.11) compared to (2.2.8) is justified by the rank-four tensor $\epsilon_{i j k l}$, which causes each term on the right-hand side of (2.2.10) to appear three times in (2.2.11).

By assumption, all singularities of $\sigma_{z}^{-1}$ lie in $S_{M} \cap S$ when $M$ is sufficiently large. Hence $\sigma_{z}^{-1}$ is well-defined in $S_{M} \Delta S:=\left(S_{M} \cup S\right) \backslash\left(S \cap S_{M}\right)$. It follows that for $(\omega, y, \xi, \zeta) \in S_{M} \Delta S$, we have

$$
\begin{aligned}
\nabla \cdot u= & \epsilon_{i j k l} \operatorname{tr} \partial_{l}\left(\partial_{i} \sigma_{z} \sigma_{z}^{-1} \partial_{j} \sigma_{z} \sigma_{z}^{-1} \partial_{k} \sigma_{z} \sigma_{z}^{-1}\right) \\
= & -\epsilon_{i j k l} \operatorname{tr}\left(\partial_{i} \sigma_{z} \sigma_{z}^{-1} \partial_{l} \sigma_{z} \sigma_{z}^{-1} \partial_{j} \sigma_{z} \sigma_{z}^{-1} \partial_{k} \sigma_{z} \sigma_{z}^{-1}+\partial_{i} \sigma_{z} \sigma_{z}^{-1} \partial_{j} \sigma_{z} \sigma_{z}^{-1} \partial_{l} \sigma_{z} \sigma_{z}^{-1} \partial_{k} \sigma_{z} \sigma_{z}^{-1}\right. \\
& \left.+\frac{1}{2} \partial_{i} \sigma_{z} \sigma_{z}^{-1} \partial_{j} \sigma_{z} \sigma_{z}^{-1} \partial_{k} \sigma_{z} \sigma_{z}^{-1} \partial_{l} \sigma_{z} \sigma_{z}^{-1}+\frac{1}{2} \partial_{l} \sigma_{z} \sigma_{z}^{-1} \partial_{i} \sigma_{z} \sigma_{z}^{-1} \partial_{j} \sigma_{z} \sigma_{z}^{-1} \partial_{k} \sigma_{z} \sigma_{z}^{-1}\right) \\
= & 0
\end{aligned}
$$

where we have used cyclicity of the trace to justify the second equality, and the antisymmetry of $\epsilon_{i j k l}$ for the last equality (the third term cancels half the second term, the fourth term cancels half the first term, half the second term cancels half the first term). Using the Divergence Theorem, we thus replace $\partial S_{M}$ in (2.2.11) by $\partial S$ and the proof is complete.

### 2.2.2 Computing the conductivity

This section contains two simplifications of Theorem 2.2.1 that will become useful when calculating the interface conductivity for systems of interest. The first result (Proposition
2.2.4) is more general, as it requires $\sigma$ to have differentiable eigenvalues and eigenvectors (up to a set of measure zero), and makes no additional assumptions besides (H1). The second (Proposition 2.2.5) provides an accessible formula for the conductivity as the degree of an appropriate map, but is restricted to certain two-dimensional models.

Throughout, we will use the shorthand $\sigma_{I}:=\sigma_{I}(H, P, \varphi)$, where it is implied that $P(x)=$ $P \in \mathfrak{S}(0,1)$ and $\varphi \in \mathfrak{S}\left(0,1 ; E_{1}, E_{2}\right)$. Recall Propositions 2.1.3 and 2.1.4, which state that $\sigma_{I}$ is independent of $\varphi$ and $P$.

Proposition 2.2.4. Suppose $H=\operatorname{Op}(\sigma)$ satisfies (H1). Fix $x_{0} \in \mathbb{R}$ and let $\lambda_{\ell}(y, \xi, \zeta)$ denote the eigenvalues of $\sigma\left(x_{0}, y, \xi, \zeta\right)$, and $\psi_{\ell}(y, \xi, \zeta)$ the corresponding eigenvectors. Given $E_{1}<\alpha<E_{2}$ and $R$ as in Theorem 2.2.1, define $n_{+}:=\left\{\ell: \lambda_{\ell}>\alpha\right.$ on $\left.\partial R\right\}$ and $n_{-}:=$ $\left\{\ell: \lambda_{\ell}<\alpha\right.$ on $\left.\partial R\right\}$. Assume that the $\lambda_{\ell}$ and $\psi_{\ell}$ are differentiable on $\partial R$, up to a set of $\Sigma$-measure zero. Then

$$
\begin{equation*}
2 \pi \sigma_{I}=\frac{i}{2 \pi} \int_{\partial R} \varepsilon_{i j k} \sum_{\ell_{+} \in n_{+}, \ell_{-} \in n_{-}} \partial_{i} \psi_{\ell_{+}}^{*} \psi_{\ell_{-}} \psi_{\ell_{-}}^{*} \partial_{j} \psi_{\ell_{+}} \nu_{k} d \Sigma \tag{2.2.12}
\end{equation*}
$$

Proof. Recall that by the definition of $\partial R, n_{+} \cup n_{-}=\{1,2, \ldots, n\}$. Define the projectors $\Pi_{\ell}:=\psi_{\ell} \psi_{\ell}^{*}$. We know that $\sigma=\sum_{\ell=1}^{n} \lambda_{\ell} \Pi_{\ell}$, and hence $\sigma_{z}^{-1}=\sum_{\ell=1}^{n}\left(z-\lambda_{\ell}\right)^{-1} \Pi_{\ell}$ whenever $\Im z \neq 0$. It follows from (2.2.1) and cyclicity of the trace that

$$
\begin{equation*}
2 \pi \sigma_{I}=\frac{i}{8 \pi^{2}} \int_{\partial R} \int_{\mathbb{R}} \sum_{\ell_{1}, \ell_{2}=1}^{n}\left(z-\lambda_{\ell_{1}}\right)^{-2}\left(z-\lambda_{\ell_{2}}\right)^{-1} d \omega \varepsilon_{i j k} \operatorname{tr} \Pi_{\ell_{1}} \partial_{i} \sigma \Pi_{\ell_{2}} \partial_{j} \sigma \nu_{k} d \Sigma \tag{2.2.13}
\end{equation*}
$$

If $\lambda_{\ell_{1}} \neq \lambda_{\ell_{2}}$, then

$$
\left(z-\lambda_{\ell_{1}}\right)^{-2}\left(z-\lambda_{\ell_{2}}\right)^{-1}=\left(\lambda_{\ell_{2}}-\lambda_{\ell_{1}}\right)^{-2}\left(\left(z-\lambda_{\ell_{2}}\right)^{-1}-\left(z-\lambda_{\ell_{1}}\right)^{-1}\right)-\left(\lambda_{\ell_{2}}-\lambda_{\ell_{1}}\right)^{-1}\left(z-\lambda_{\ell_{1}}\right)^{-2} .
$$

Thus the integral over $\omega$ vanishes if $\lambda_{\ell_{1}}-\alpha$ and $\lambda_{\ell_{2}}-\alpha$ have the same sign, and otherwise
equals $2 \pi\left(\lambda_{\ell_{2}}-\lambda_{\ell_{1}}\right)^{-2}$. It follows that

$$
2 \pi \sigma_{I}=\frac{i}{2 \pi} \int_{\partial R_{\ell_{+} \in n_{+}, \ell_{-} \in n_{-}}}\left(\lambda_{\ell_{+}}-\lambda_{\ell_{-}}\right)^{-2} \varepsilon_{i j k} \operatorname{tr} \Pi_{\ell_{+}} \partial_{i} \sigma \Pi_{\ell_{-}} \partial_{j} \sigma \nu_{k} d \Sigma,
$$

where we have added the contributions of $\left(\ell_{1}, \ell_{2}\right) \in n_{+} \times n_{-}$and $\left(\ell_{1}, \ell_{2}\right) \in n_{-} \times n_{+}$in (2.2.13). Finally, we observe that

$$
\begin{aligned}
\operatorname{tr} \Pi_{\ell_{+}} \partial_{i} \sigma \Pi_{\ell_{-}} \partial_{j} \sigma= & \operatorname{tr}\left(\lambda_{\ell_{+}}^{2} \Pi_{\ell_{+}} \partial_{i} \Pi_{\ell_{+}} \Pi_{\ell_{-}} \partial_{j} \Pi_{\ell_{+}}+\lambda_{\ell_{+}} \lambda_{\ell_{-}} \Pi_{\ell_{+}} \partial_{i} \Pi_{\ell_{+}} \Pi_{\ell_{-}} \partial_{j} \Pi_{\ell_{-}}\right. \\
& \left.+\lambda_{\ell_{+}} \lambda_{\ell_{-}} \Pi_{\ell_{+}} \partial_{i} \Pi_{\ell_{-}} \Pi_{\ell_{-}} \partial_{j} \Pi_{\ell_{+}}+\lambda_{\ell_{-}}^{2} \Pi_{\ell_{+}} \partial_{i} \Pi_{\ell_{-}} \Pi_{\ell_{-}} \partial_{j} \Pi_{\ell_{-}}\right) \\
= & -\left(\lambda_{\ell_{+}}-\lambda_{\ell_{-}}\right)^{2} \operatorname{tr} \psi_{\ell_{+}} \partial_{i} \psi_{\ell_{+}}^{*} \psi_{\ell_{-}} \partial_{j} \psi_{\ell_{-}}^{*} \\
= & \left(\lambda_{\ell_{+}}-\lambda_{\ell_{-}}\right)^{2} \partial_{i} \psi_{\ell_{+}}^{*} \psi_{\ell_{-}} \psi_{\ell_{-}}^{*} \partial_{j} \psi_{\ell_{+}}
\end{aligned}
$$

and the result is complete.

We now simplify Theorem 2.2.1 for a large class of 2-dimensional models. An analogous result for tight-binding models can be found in the literature [46]. For the current (non tightbinding) setting, one can find a similar result in [42] that expresses topological invariants in terms of the resolvent $(i \omega-H)^{-1}$. Our result is unique in that it provides a simple explicit expression for $\sigma_{I}$ that can easily be computed without the resolvent or a spectral decomposition.

We say that a point $a \in \mathbb{R}^{d}$ is a regular value of a map $f: \mathbb{R}^{d} \rightarrow \mathbb{R}^{d}$ if the preimage $f^{-1}(a)$ is a finite collection of points $\left\{x_{1}, x_{2}, \ldots, x_{p}\right\} \subset \mathbb{R}^{d}$ such that the Jacobian $M_{j}^{m n}=\partial_{m} f_{n} \mid x_{j}$ is nonsingular for all $j$. We let $\mathcal{R}(f)$ denote the set of all regular values of $f$. By Sard's Theorem [75], we know that $\mathbb{R}^{d} \backslash \mathcal{R}(f)$ has measure zero.

Proposition 2.2.5. Let $H=\operatorname{Op}(\sigma)$ satisfy (H1), and fix $x_{0} \in \mathbb{R}$. Suppose there exist
(smooth and real-valued) functions $f_{1}, f_{2}, f_{3}$ such that

$$
\sigma\left(x_{0}, y, \xi, \zeta\right)=f_{1}(y, \xi, \zeta) \sigma_{1}+f_{2}(y, \xi, \zeta) \sigma_{2}+f_{3}(y, \xi, \zeta) \sigma_{3}
$$

for all $(y, \xi, \zeta) \in \mathbb{R}^{3}$, where

$$
\sigma_{1}=\left(\begin{array}{cc}
0 & 1  \tag{2.2.14}\\
1 & 0
\end{array}\right), \quad \sigma_{2}=\left(\begin{array}{cc}
0 & -i \\
i & 0
\end{array}\right), \quad \sigma_{3}=\left(\begin{array}{cc}
1 & 0 \\
0 & -1
\end{array}\right)
$$

are the Pauli matrices. In particular, this means there exists a three-dimensional ball $S$ such that the vector field $f:=\left(f_{1}, f_{2}, f_{3}\right)$ satisfies $|f|:=\sqrt{f_{1}^{2}+f_{2}^{2}+f_{3}^{2}} \geq \varepsilon_{0}$ in $\mathbb{R}^{3} \backslash S$, for some $\varepsilon_{0}>0$. Then the set $\mathcal{R}_{\varepsilon_{0}}:=\mathcal{R}(f) \cap\left\{|a|<\varepsilon_{0}\right\}$ is nonempty. Let $a \in \mathcal{R}_{\varepsilon_{0}}$, and define $f^{-1}(a)=:\left\{\left(y_{j}, \xi_{j}, \zeta_{j}\right)\right\}_{j=1}^{p}$. Then

$$
\begin{equation*}
2 \pi \sigma_{I}=-\sum_{j=1}^{p} \operatorname{sgn} \operatorname{det} M_{j} \tag{2.2.15}
\end{equation*}
$$

where $M_{j} \in \mathbb{R}^{3 \times 3}$ is the Jacobian matrix defined by $M_{j}^{m n}=\left.\partial_{m} f_{n}\right|_{\left(y_{j}, \xi_{j}, \zeta_{j}\right)}$.

Proof. The fact that $\mathcal{R}_{\varepsilon_{0}} \neq \emptyset$ is an immediate consequence of Sard's Theorem [75]. Given the structure of $\sigma$, we take $\alpha=0$ in Theorem 2.2.1, meaning that $z=i \omega$ in (2.2.1). It follows that

$$
\sigma_{z}^{-1}=-\left(\omega^{2}+|f|^{2}\right)^{-1}(i \omega+\sigma)
$$

from which we conclude that

$$
\begin{aligned}
& \sigma_{z}^{-1} \partial_{i} \sigma_{z} \sigma_{z}^{-1} \partial_{j} \sigma_{z} \sigma_{z}^{-1}=-\left(\omega^{2}+|f|^{2}\right)^{-3}(i \omega+\sigma) \partial_{i} \sigma(i \omega+\sigma) \partial_{j} \sigma(i \omega+\sigma) \\
= & -\left(\omega^{2}+|f|^{2}\right)^{-3} \sigma \partial_{i} \sigma \sigma \partial_{j} \sigma \sigma+\omega^{2}\left(\omega^{2}+|f|^{2}\right)^{-3}\left(\partial_{i} \sigma \partial_{j} \sigma \sigma+\partial_{i} \sigma \sigma \partial_{j} \sigma+\sigma \partial_{i} \sigma \partial_{j} \sigma\right)+\ldots,
\end{aligned}
$$

where we have left out terms that are odd in $\omega$ in the last line. Carrying out the integral over $\omega$ in (2.2.1), we find that

$$
2 \pi \sigma_{I}=\frac{i}{8 \pi^{2}} \int_{\partial R}\left(-\frac{3 \pi \epsilon_{i j k}}{8|f|^{5}} \operatorname{tr} \sigma \partial_{i} \sigma \sigma \partial_{j} \sigma \sigma+\frac{\pi \epsilon_{i j k}}{8|f|^{3}} \operatorname{tr}\left(\partial_{i} \sigma \partial_{j} \sigma \sigma+\partial_{i} \sigma \sigma \partial_{j} \sigma+\sigma \partial_{i} \sigma \partial_{j} \sigma\right)\right) \nu_{k} d \Sigma .
$$

Simplifying the above using cyclicity of the trace and the fact that $\sigma^{2}=|f|^{2}$, we have

$$
2 \pi \sigma_{I}=\frac{i}{16 \pi} \int_{\partial R} \frac{\epsilon_{i j k}}{|f|^{3}} \operatorname{tr} \sigma \partial_{i} \sigma \partial_{j} \sigma \nu_{k} d \Sigma
$$

Using commutation relations of the Pauli matrices, we see that

$$
\operatorname{tr} \sigma \partial_{i} \sigma \partial_{j} \sigma=4 i\left(f_{1}\left(\partial_{i} f_{2} \partial_{j} f_{3}-\partial_{i} f_{3} \partial_{j} f_{2}\right)+[(1,2,3) \rightarrow(2,3,1)]+[(1,2,3) \rightarrow(3,1,2)]\right),
$$

which we may rewrite as $4 i \epsilon_{m n p} f_{m} \partial_{i} f_{n} \partial_{j} f_{p}$. Letting $w \in \mathbb{R}^{3}$ such that $w_{k}=\frac{\epsilon_{i j k}}{|f|^{3}} \operatorname{tr} \sigma \partial_{i} \sigma \partial_{j} \sigma$, we have

$$
\nabla \cdot w=\frac{4 i}{|f|^{3}} \tilde{\epsilon}_{i j k} \epsilon_{m n p} \partial_{k} f_{m} \partial_{i} f_{n} \partial_{j} f_{p}\left(1-\frac{3 f_{m}^{2}}{f_{1}^{2}+f_{2}^{2}+f_{3}^{2}}\right)
$$

away from the zeros of $|f|^{2}$, where the only nonzero entries of $\tilde{\epsilon}_{i j k}$ are $\tilde{\epsilon}_{123}=\tilde{\epsilon}_{231}=\tilde{\epsilon}_{312}=$ 1. By symmetry, we know that $\tilde{\epsilon}_{i j k} \epsilon_{m n p} f_{m}^{2} \partial_{k} f_{m} \partial_{i} f_{n} \partial_{j} f_{p}=\tilde{\epsilon}_{i j k} \epsilon_{m n p} f_{n}^{2} \partial_{k} f_{m} \partial_{i} f_{n} \partial_{j} f_{p}=$ $\tilde{\epsilon}_{i j k} \epsilon_{m n p} f_{p}^{2} \partial_{k} f_{m} \partial_{i} f_{n} \partial_{j} f_{p}$, meaning that

$$
\nabla \cdot w=\frac{4 i}{|f|^{3}} \tilde{\epsilon}_{i j k} \epsilon_{m n p} \partial_{k} f_{m} \partial_{i} f_{n} \partial_{j} f_{p}\left(1-\frac{f_{m}^{2}+f_{n}^{2}+f_{p}^{2}}{f_{1}^{2}+f_{2}^{2}+f_{3}^{2}}\right)=0
$$

Thus $2 \pi \sigma_{I}=\frac{i}{16 \pi} \int_{\partial R} w \cdot \hat{\nu} d \Sigma$, where $\hat{\nu}$ is the unit vector normal to the surface $\partial R$ and $w$ has zero divergence everywhere except for a finite number of points in $\mathbb{R}^{3}$. Hence we can
deform $\partial R$ in any way so long as it encloses the same zeros of $|f|^{3}$. It follows that

$$
2 \pi \sigma_{I}=\frac{i}{16 \pi} \int_{\partial S} w \cdot \hat{\nu} d \Sigma ; \quad w \cdot \hat{\nu}=\frac{-4 i}{|f|^{3} \sqrt{\xi^{2}+\zeta^{2}+y^{2}}} \operatorname{det}\left[\begin{array}{cccc}
0 & f_{1} & f_{2} & f_{3} \\
\xi & \partial_{\xi} f_{1} & \partial_{\xi} f_{2} & \partial_{\xi} f_{3} \\
\zeta & \partial_{\zeta} f_{1} & \partial_{\zeta} f_{2} & \partial_{\zeta} f_{3} \\
y & \partial_{y} f_{1} & \partial_{y} f_{2} & \partial_{y} f_{3}
\end{array}\right]
$$

Using the geometric interpretation of the determinant, we have

$$
2 \pi \sigma_{I}=-\frac{1}{4 \pi} \int_{\partial S} \frac{1}{|f|^{3}} \operatorname{det}\left[\begin{array}{ccc}
f_{1} & f_{2} & f_{3} \\
\partial_{u} f_{1} & \partial_{u} f_{2} & \partial_{u} f_{3} \\
\partial_{v} f_{1} & \partial_{v} f_{2} & \partial_{v} f_{3}
\end{array}\right] d u \wedge d v
$$

where $u=-\hat{\xi} \sin \phi+\hat{\zeta} \cos \phi$ and $v=-\hat{\xi} \cos \theta \cos \phi-\hat{\zeta} \cos \theta \sin \phi+\hat{y} \sin \theta$, with $\theta$ and $\phi$ the polar and azimuthal angles that parametrize $\partial S$. We conclude from [38, Corollary 14.2.1] and continuity of the degree of a map that $2 \pi \sigma_{I}=-\operatorname{deg} g$, where $g: \partial S \rightarrow \mathbb{S}^{2}$ is the Gauss map defined by

$$
(y, \xi, \zeta) \mapsto \frac{f(y, \xi, \zeta)-a}{|f(y, \xi, \zeta)-a|}
$$

The fact that $\operatorname{deg} g=\sum_{j=1}^{p} \operatorname{sgn} \operatorname{det} M_{j}$ is a direct consequence of [38, Theorems 14.4.3 and 14.4.4].

### 2.3 Applications

We now apply the preceding results to standard Hamiltonians that appear in the analysis of topological insulators and superconductors, and in particular the $2 \times 2$ Dirac system
$[7,8,10,11,68,87]$

$$
\begin{equation*}
H=D_{x} \sigma_{1}+D_{y} \sigma_{2}+m(y) \sigma_{3}, \tag{2.3.1}
\end{equation*}
$$

and models for $p$-wave and $d$-wave superconductors following [86],

$$
\begin{align*}
H & =\left(\frac{1}{2 m}\left(D_{x}^{2}+D_{y}^{2}\right)-\mu\right) \sigma_{1}+\frac{1}{2}\left(c(y) D_{y}+D_{y} c(y)\right) \sigma_{2}+c_{0} D_{x} \sigma_{3}  \tag{2.3.2}\\
H & =\left(\frac{1}{2 m}\left(D_{x}^{2}+D_{y}^{2}\right)-\mu\right) \sigma_{1}+c_{0}\left(D_{y}^{2}-D_{x}^{2}\right) \sigma_{2}+\frac{1}{2} D_{x}\left(c(y) D_{y}+D_{y} c(y)\right) \sigma_{3} \tag{2.3.3}
\end{align*}
$$

Here, $c_{0}, m$, and $\mu$ are fixed positive constants, the $\sigma_{i}$ are the Pauli matrices (2.2.14), and we have defined $D_{\alpha}:=-i \partial_{\alpha}$. We assume that $m(y)$ and $c(y)$ are smooth domain walls with

$$
m(y)=\left\{\begin{array}{ll}
m_{-}, & y \leq-y_{0} \\
m_{+}, & y \geq y_{0}
\end{array} \quad \text { and } \quad c(y)= \begin{cases}c_{-}, & y \leq-y_{0} \\
c_{+}, & y \geq y_{0}\end{cases}\right.
$$

for some $y_{0}>0$, where the constants $m_{ \pm}$and $c_{ \pm}$are all nonzero. Without loss of generality, assume $m^{\prime}(y) \neq 0$ for all $y \in m^{-1}(0)$, and $c^{\prime}(y) \neq 0$ for all $y \in c^{-1}(0)$.

Throughout this section, let $\sigma_{\min }$ denote the smallest-magnitude eigenvalue of $\sigma$. We continue to use the shorthand $\sigma_{I}:=\sigma_{I}(H, P, \varphi)$, where $P(x)=P \in \mathfrak{S}(0,1)$ and $\varphi \in$ $\mathfrak{S}\left(0,1 ; E_{1}, E_{2}\right)$ is implied.
$2 \times 2$ Dirac system (2.3.1). We have $H=\operatorname{Op}(\sigma)$, with $\sigma=\xi \sigma_{1}+\zeta \sigma_{2}+m(y) \sigma_{3}$. Thus it is clear that $\sigma \in S_{1,0}^{1}$ with $\sigma^{2}=\xi^{2}+\zeta^{2}+m^{2}(y) \geq \xi^{2}+\zeta^{2}$. Using that $\langle\xi, \zeta\rangle-|(\xi, \zeta)| \leq 1$, it follows that $\left|\sigma_{\min }\right| \geq\langle\xi, \zeta\rangle-1$. Now, take $\sigma_{ \pm}:=\xi \sigma_{1}+\zeta \sigma_{2} \pm m_{ \pm} \sigma_{3}$, with $E_{2}:=$ $\min \left\{\left|m_{+}\right|,\left|m_{-}\right|\right\}$and $E_{1}:=-E_{2}$. Since $\sigma_{ \pm}^{2}=\xi^{2}+\zeta^{2}+m_{ \pm}^{2} \geq m_{ \pm}^{2},(H 1)$ is satisfied.

We now apply Proposition 2.2 .5 with $f=(\xi, \zeta, m(y))$, so that $f^{-1}(0)=\{(y, 0,0)$ : $m(y)=0\}$. Given a point $y \in m^{-1}(0)$, the determinant of the Jacobian of $f$ evaluated at $(y, 0,0)$ is $m^{\prime}(y)$. Thus it follows that $2 \pi \sigma_{I}=\frac{1}{2}\left(\operatorname{sgn}\left(m_{-}\right)-\operatorname{sgn}\left(m_{+}\right)\right)$, as in e.g., $[7,10]$.
$p$-wave superconductor model (2.3.2). We have $H=\operatorname{Op}(\sigma)$, with

$$
\sigma=\left(\frac{1}{2 m}\left(\xi^{2}+\zeta^{2}\right)-\mu\right) \sigma_{1}+c(y) \zeta \sigma_{2}+c_{0} \xi \sigma_{3} .
$$

It follows that $\sigma \in S_{1,0}^{2}$ with, moreover, $\sigma^{2}=\left(\frac{1}{2 m}\left(\xi^{2}+\zeta^{2}\right)-\mu\right)^{2}+c^{2}(y) \zeta^{2}+c_{0}^{2} \xi^{2} \geq$ $\left(\frac{1}{2 m}\left(\xi^{2}+\zeta^{2}\right)-\mu\right)^{2}$, which implies

$$
\left|\sigma_{\min }\right| \geq\left|\frac{1}{2 m}\left(\xi^{2}+\zeta^{2}\right)-\mu\right|=\left|\frac{1}{2 m}\left(\langle\zeta, \zeta\rangle^{2}-1\right)-\mu\right|
$$

Thus there exists $C_{1}>0$ and a compact set $K \subset \mathbb{R}^{2}$ such that $\left|\sigma_{\min }\right| \geq C_{1}\langle\xi, \zeta\rangle^{2}$ for all $(\xi, \zeta) \in \mathbb{R}^{2} \backslash K$. Moreover, there exists $C_{2}>0$ such that $\langle\xi, \zeta\rangle^{2} \leq C_{2}$ and hence $1+\left|\sigma_{\min }\right| \geq 1 \geq \frac{1}{C_{2}}\langle\xi, \zeta\rangle^{2}$ for all $(\xi, \zeta) \in K$. It follows that for $c:=\min \left\{C_{1}, \frac{1}{C_{2}}\right\}$, we have $1+\left|\sigma_{\min }\right| \geq c\langle\xi, \zeta\rangle^{2}$, for $(\xi, \zeta) \in \mathbb{R}^{2}$. Now, take $\sigma_{ \pm}:=\left(\frac{1}{2 m}\left(\xi^{2}+\zeta^{2}\right)-\mu\right) \sigma_{1}+c_{ \pm} \zeta \sigma_{2}+c_{0} \xi \sigma_{3}$, so that $\sigma_{ \pm}^{2}=\left(\frac{1}{2 m}\left(\xi^{2}+\zeta^{2}\right)-\mu\right)^{2}+c_{ \pm}^{2} \zeta^{2}+c_{0}^{2} \xi^{2}$. Minimizing the above with respect to $(\xi, \zeta)$ is equivalent to minimizing

$$
f_{ \pm}(a, b):=\left(\frac{1}{2 m}(a+b)-\mu\right)^{2}+c_{ \pm}^{2} b+c_{0}^{2} a
$$

with respect to $(a, b) \geq 0$. We see that

$$
\begin{gathered}
\partial_{a} f_{ \pm}=\frac{1}{m}\left(\frac{1}{2 m}(a+b)-\mu\right)+c_{0}^{2}, \quad \partial_{b} f_{ \pm}=\frac{1}{m}\left(\frac{1}{2 m}(a+b)-\mu\right)+c_{ \pm}^{2} \\
\partial_{a a}^{2} f_{ \pm}=\partial_{a b}^{2} f_{ \pm}=\partial_{b b}^{2} f_{ \pm}=\frac{1}{2 m^{2}}
\end{gathered}
$$

hence $f_{ \pm}$is convex. Since $\partial_{a} f_{ \pm}-\partial_{b} f_{ \pm}=c_{0}^{2}-c_{ \pm}^{2}$, we know that the minimum value of $f_{ \pm}$is attained when either $a=0$ or $b=0$. We have $f_{ \pm}(0, b)=\left(\frac{1}{2 m} b-\mu\right)^{2}+c_{ \pm}^{2} b$, which is a convex function that is minimized when $\frac{1}{m}\left(\frac{1}{2 m} b-\mu\right)+c_{ \pm}^{2}=0$, or equivalently $b=2 m \mu-2 m^{2} c_{ \pm}^{2}=: b^{*}$. Similarly, $f_{ \pm}(a, 0)$ is a convex function that is minimized when
$a=2 m \mu-2 m^{2} c_{0}^{2}=: a^{*}$.
Case 1: Suppose $a^{*}, b^{*}<0$. Then the solution for our minimization problem is $a=b=0$, which yields a value of $f_{ \pm}(0,0)=\frac{1}{4 m^{2}} \mu^{2}$.

Case 2: Suppose $a^{*}<0<b^{*}$. Then the minimum value of $f_{ \pm}$over the allowed domain is

$$
f_{ \pm}\left(0, b^{*}\right)=m^{2} c_{ \pm}^{4}+\left(2 m \mu-2 m^{2} c_{ \pm}^{2}\right) c_{ \pm}^{2}=2 m \mu c_{ \pm}^{2}-m^{2} c_{ \pm}^{4}=m c_{ \pm}^{2}\left(2 \mu-m c_{ \pm}^{2}\right)
$$

which is bounded below by $m^{2} c_{ \pm}^{4}$.
Case 3: Suppose $b^{*}<0<a^{*}$. Then the minimum value of $f_{ \pm}$over the allowed domain is

$$
f_{ \pm}\left(a^{*}, 0\right)=m^{2} c_{0}^{4}+c_{0}^{2}\left(2 m \mu-2 m^{2} c_{0}^{2}\right)=m c_{0}^{2}\left(2 \mu-m c_{0}^{2}\right),
$$

which is bounded below by $m^{2} c_{0}^{4}$.
Case 4: Suppose $a^{*}, b^{*}>0$. Then the minimum value of $f_{ \pm}$over the allowed domain is

$$
\min \left\{f_{ \pm}\left(a^{*}, 0\right), f_{ \pm}\left(0, b^{*}\right)\right\}=m c_{1}^{2}\left(2 \mu-m c_{1}^{2}\right), \quad c_{1}:=\min \left\{c_{0}, c_{ \pm}\right\}
$$

In each case, we have proven the existence of $E_{ \pm}>0$ such that $\sigma_{ \pm}^{2} \geq E_{ \pm}^{2}$. Define $E_{0}:=$ $\min \left\{E_{+}, E_{-}\right\}$, and take $\left(E_{1}, E_{2}\right):=\left(-E_{0}, E_{0}\right)$. Thus we have verified (H1).

We now apply Proposition 2.2 .5 with $f_{1}=\frac{1}{2 m}\left(\xi^{2}+\zeta^{2}\right)-\mu, f_{2}=\zeta c(y)$, and $f_{3}=c_{0} \xi$. The zeros of $f_{1}^{2}+f_{2}^{2}+f_{3}^{2}$ are $(y, \xi, \zeta)=\left(y_{1}, 0, \pm \sqrt{2 m \mu}\right)$, for all $y_{1} \in c^{-1}(0)$. A straightforward calculation reveals that $\left.\operatorname{sgn} \operatorname{det} \partial_{m} f_{n}\right|_{\left(y_{1}, 0, \pm \sqrt{2 m \mu}\right)}=\operatorname{sgn} c^{\prime}\left(y_{1}\right)$, hence $2 \pi \sigma_{I}=\operatorname{sgn}\left(c_{-}\right)-$ $\operatorname{sgn}\left(c_{+}\right)$.
$d$-wave superconductor model (2.3.3). Here, the symbol of $H$ is

$$
\sigma=\left(\frac{1}{2 m}\left(\xi^{2}+\zeta^{2}\right)-\mu\right) \sigma_{1}+c_{0}\left(\zeta^{2}-\xi^{2}\right) \sigma_{2}+\xi \zeta c(y) \sigma_{3}
$$

We see that $\sigma \in S_{1,0}^{2}$, and $\sigma^{2}=\left(\frac{1}{2 m}\left(\xi^{2}+\zeta^{2}\right)-\mu\right)^{2}+c_{0}^{2}\left(\zeta^{2}-\xi^{2}\right)^{2}+\xi^{2} \zeta^{2} c^{2}(y) \geq\left(\frac{1}{2 m}\left(\xi^{2}+\right.\right.$ $\left.\left.\zeta^{2}\right)-\mu\right)^{2}$, hence $\sigma$ is elliptic. As with the $p$-wave superconductor, we set $\sigma_{ \pm}$equal to $\sigma$ with $c(y)$ replaced by $c_{ \pm}$.

We now apply Proposition 2.2.5, with $f_{1}=\frac{1}{2 m}\left(\xi^{2}+\zeta^{2}\right)-\mu, f_{2}=c_{0}\left(\zeta^{2}-\xi^{2}\right)$, and $f_{3}=\xi \zeta c(y)$. The zeros of $f_{1}^{2}+f_{2}^{2}+f_{3}^{2}$ are $(y, \xi, \zeta)=\left(y_{1}, \varepsilon_{1} \sqrt{m \mu}, \varepsilon_{2} \sqrt{m \mu}\right)$, for all $y_{1} \in c^{-1}(0)$ and $\varepsilon_{1}, \varepsilon_{2} \in\{-1,1\}$. (That is, there are four zeros for every $y_{1} \in c^{-1}(0)$.) One can then easily verify that $\left.\operatorname{sgn} \operatorname{det} \partial_{m} f_{n}\right|_{\left(y_{1}, \varepsilon_{1} \sqrt{m \mu}, \varepsilon_{2} \sqrt{m \mu}\right)}=\operatorname{sgn} c^{\prime}\left(y_{1}\right)$. Thus we conclude that $2 \pi \sigma_{I}=2\left(\operatorname{sgn}\left(c_{-}\right)-\operatorname{sgn}\left(c_{+}\right)\right)$.

Regularized model of equatorial waves. The two-dimensional water wave model is given by a Hamiltonian $H_{0}=\operatorname{Op}\left(\sigma_{0}\right)$ with

$$
\sigma_{0}(x, y, \xi, \zeta)=\left(\begin{array}{ccc}
0 & \xi & \zeta  \tag{2.3.4}\\
\xi & 0 & -i f(y) \\
\zeta & i f(y) & 0
\end{array}\right)
$$

where $f(y)$ is a Coriolis force that is positive when $y>0$ (northern hemisphere of (necessarily) flat Earth) and negative when $y<0$; see $[10,33,79]$ for background on this water wave problem and in particular the observation that the bulk-interface correspondence fails for certain profiles $f(y)[10]$.

We thus consider here a regularized version given by Hamiltonian $H_{\mu}=\operatorname{Op}(\sigma)$ given by
$\sigma=\lambda_{+} \Pi_{+}+\lambda_{-} \Pi_{-}+\lambda_{0} \Pi_{0}$, where, following calculations in [10, 33]

$$
\Pi_{j}=\psi_{j} \psi_{j}^{*}, \quad \psi_{0}=\frac{1}{\kappa}\left[\begin{array}{c}
i f \\
\zeta \\
-\xi
\end{array}\right], \quad \psi_{ \pm}=\frac{1}{\rho}\left[\begin{array}{c}
i f \xi \pm \kappa \zeta \\
\xi \zeta \pm i f \kappa \\
\zeta^{2}+f^{2}
\end{array}\right], \quad \lambda_{0}=\mu \kappa^{2}\left(1+\kappa^{2}\right)^{-\frac{1}{2}}, \quad \lambda_{ \pm}= \pm \kappa
$$

with $f \in \mathfrak{S}\left(f_{-}, f_{+}\right), \kappa=\sqrt{f^{2}+\xi^{2}+\zeta^{2}}$, and $\rho=\kappa \sqrt{2\left(f^{2}+\zeta^{2}\right)}$ for some nonzero constants $\mu$ and $f_{ \pm}$. We verify that $H_{\mu}=H_{0}$ when the regularization parameter $\mu=0$.

The role of the regularization is to replace the infinitely degenerate (flat band) eigenvalue 0 by a topologically trivial band with eigenvalues tending to $\pm \infty$ as $|(\xi, \zeta)| \rightarrow \infty$. The choice of the regularization for $\lambda_{0}$ ensure that $\sigma \in S_{1,0}^{1}$ and is elliptic when $\mu \neq 0$.

As above, we set $\sigma_{ \pm}$equal to $\sigma$ with $f(y)$ replaced by $f_{ \pm}$. Defining $\mu_{1}:=\min \{|\mu|, 1\}$ and $f_{0}:=\min \left\{\left|f_{+}\right|,\left|f_{-}\right|\right\}$, we can set $E_{2}:=f_{0} \min \left\{\mu_{1},|\mu| f_{0}\right\}$ and $E_{1}:=-E_{2}$, so that (H1) is satisfied.

We apply Proposition 2.2.4, with $\mu>0$ for concreteness. Take $\partial R_{3}:=\left\{y^{2}+\xi^{2}+\zeta^{2}=\right.$ $\left.r^{2}\right\} \subset \mathbb{R}^{3}$, with $r>0$ sufficiently large so that $|f(y)| \geq f_{0}$ whenever $|y| \geq r$. The $\lambda_{j}$ and $\psi_{j}$ are differentiable everywhere on $\partial R_{3}$, except perhaps where $\rho=0$. But $\{\rho=0\} \cap \partial R_{3} \subset$ $\{\zeta=0\} \cap\left\{y^{2}+\xi^{2}=r^{2}\right\}$ has $\Sigma$-measure zero, so indeed the regularity requirement is satisfied. Observe that

$$
\partial_{\xi} \psi_{0}=\frac{1}{\kappa}\left[\begin{array}{c}
0 \\
0 \\
-1
\end{array}\right]+c_{1} \psi_{0}, \quad \partial_{\zeta} \psi_{0}=\frac{1}{\kappa}\left[\begin{array}{l}
0 \\
1 \\
0
\end{array}\right]+c_{2} \psi_{0}, \quad \partial_{y} \psi_{0}=\frac{1}{\kappa}\left[\begin{array}{c}
i f^{\prime} \\
0 \\
0
\end{array}\right]+c_{3} \psi_{0}
$$

where the $c_{i}$ are scalar-valued functions. The terms in the integrand of (2.2.12) corresponding to the pair $\left\{\lambda_{0}, \lambda_{-}\right\}$are $\Im\left(\left(\psi_{-}^{*} \partial_{\xi} \psi_{0}\right)^{*} \psi_{-}^{*} \partial_{\zeta} \psi_{0}\right)=-\frac{1}{\kappa^{2} \rho^{2}} f \kappa\left(\zeta^{2}+f^{2}\right), \Im\left(\left(\psi_{-}^{*} \partial_{\xi} \psi_{0}\right)^{*} \psi_{-}^{*} \partial_{y} \psi_{0}\right)=$ $\frac{1}{\kappa^{2} \rho^{2}} f^{\prime} \kappa \zeta\left(\zeta^{2}+f^{2}\right)$, and $\Im\left(\left(\psi_{-}^{*} \partial_{\zeta} \psi_{0}\right)^{*} \psi_{-}^{*} \partial_{y} \psi_{0}\right)=-\frac{1}{\kappa^{2} \rho^{2}} f^{\prime} \kappa \xi\left(\zeta^{2}+f^{2}\right)$. Thus the contribution
to $2 \pi \sigma_{I}$ of $\left\{\lambda_{0}, \lambda_{-}\right\}$is

$$
\begin{aligned}
& \frac{i}{2 \pi} \int_{\partial R_{3}} \varepsilon_{i j k} \partial_{i} \psi_{0}^{*} \psi_{-} \psi_{-}^{*} \partial_{j} \psi_{0} \nu_{k} d \Sigma \\
& \quad=\frac{i}{2 \pi} \int_{\partial R_{3}} 2 i\left(\Im\left(\left(\psi_{-}^{*} \partial_{\xi} \psi_{0}\right)^{*} \psi_{-}^{*} \partial_{\zeta} \psi_{0}\right) \nu_{y}-\Im\left(\left(\psi_{-}^{*} \partial_{\xi} \psi_{0}\right)^{*} \psi_{-}^{*} \partial_{y} \psi_{0}\right) \nu_{\zeta}\right. \\
& \left.\quad+\Im\left(\left(\psi_{-}^{*} \partial_{\zeta} \psi_{0}\right)^{*} \psi_{-}^{*} \partial_{y} \psi_{0}\right) \nu_{\xi}\right) d \Sigma \\
& \quad=\frac{1}{\pi} \int_{\partial R_{3}} \frac{\zeta^{2}+f^{2}}{\kappa^{2} \rho^{2}}\left(f y+f^{\prime} \zeta^{2}+f^{\prime} \xi^{2}\right) d \Sigma=\frac{1}{2 \pi} \int_{\partial R_{3}} \frac{1}{\kappa^{4}}\left(f y+f^{\prime} \zeta^{2}+f^{\prime} \xi^{2}\right) d \Sigma
\end{aligned}
$$

Considering the case $f_{-}<0<f_{+}$, we can without loss of generality take $f(y)=y$ on $\partial R_{3}$, so that the integral becomes $\frac{1}{2 \pi} \int_{\partial R_{3}} \frac{1}{\kappa^{2}} d \Sigma=2$. One verifies that $\psi_{-}^{*} \partial_{\xi} \psi_{+}=\psi_{-}^{*} \partial_{\zeta} \psi_{+}=$ $\psi_{-}^{*} \partial_{y} \psi_{+}=0$, meaning there is no contribution to the conductivity from the pair $\left\{\lambda_{+}, \lambda_{-}\right\}$.

We conclude that $2 \pi \sigma_{I}=2$, which in practice corresponds to two (observed) eastwardmoving asymmetric modes along the equator. A similar calculation shows that $2 \pi \sigma_{I}=2$ also when $\mu<0$, so that the conductivity is independent of the regularization parameter $\mu \neq 0$. If instead we assume that $f_{+}<0<f_{-}$(with south pole pointing upwards), then $2 \pi \sigma_{I}=-2$.

### 2.4 Extension to junction models

Above, we considered models of two 2-dimensional insulators that were joined together along a flat (e.g. $y=0$ ) interface. In this subsection, we will generalize the theory to more complicated geometries; see Figure 2.1 for an illustration. Our goal is to describe systems exhibiting surface waves that propagate along the interfaces between the + and - regions with a preferred direction (e.g. moving from left to right). Analogous to (1.0.1), we will define a junction conductivity that quantifies the net current associated with these waves at a particular energy. One can define the conductivity so that it measures the contribution from any combination of interfaces (more details below Corollary 2.4.3).


Figure 2.1: Two illustrations of the models considered in Section 2.4. Here, two insulators (labeled by + and - ) are joined together at a junction. We analyze the asymmetric transport along each interface and through the junction. Two example level curves $\{g=0\}$ are presented in each case (solid and dashed curves), for some arbitrary choices of $g$ satisfying the appropriate assumptions below. The scale of this picture is much smaller than 1 , so that we are not in the region where $g(r, \theta)=r g_{\Theta}(\theta)$.

The main result of this subsection (Theorem 2.4.8 below) is a bulk-interface correspondence, which equates the conductivity to the difference of two integrals depending only on the associated insulators. In particular, the conductivity does not depend on local properties of the material, such as the way the insulators are joined together at each interface. Theorem 2.4.8 allows for straight-forward evaluations of the conductivity using the previously developed theory for interface models in Sections 2.1-2.3.

The proof of Theorem 2.4.8 requires familiar stability properties of the conductivity, resembling those of the previous subsections. As an alternative approach, we tie the conductivity to a Fredholm index (Theorem 2.4.2), with the latter known to be quantized and invariant with respect to a large class of perturbations. This implies a conservation law (Corollary 2.4.3) showing (among other things) that the currents entering and leaving the junction are the same. We also prove invariance of the junction conductivity with respect to the corresponding density of states (Corollary 2.4.4), semiclassical rescaling (Theorem 2.4.7) and other "continuous" perturbations of the Hamiltonian (Corollary 2.4.5). We conclude the section with applications to a continuum model of twisted bilayer graphene and a $2 \times 2$
junction Dirac model.
Let us introduce some notation. As above, the Hamiltonians $H$ act on $\mathcal{H}:=L^{2}\left(\mathbb{R}^{2}\right) \otimes \mathbb{C}^{n}$. We label the spatial coordinates $(x, y) \in \mathbb{R}^{2}$ and the corresponding dual variables $(\xi, \zeta) \in \mathbb{R}^{2}$. Given functions $u \in \mathcal{C}^{\infty}\left(\mathbb{R}^{2}\right)$ and $A \in \mathcal{C}^{\infty}\left(\mathbb{R}^{2} ; \mathbb{C}^{n \times n}\right)$, fixed matrices $A_{1}$ and $A_{2}$ in $\mathbb{C}^{n \times n}$, and fixed constants $c_{1} \leq c_{2}$, we write $A \in \mathfrak{S}\left(A_{1}, A_{2} ; c_{1}, c_{2} ; u\right)$ to mean that

$$
A= \begin{cases}A_{1}, & u<c_{1}  \tag{2.4.1}\\ A_{2}, & u>c_{2}\end{cases}
$$

We let $\mathfrak{S}\left(A_{1}, A_{2} ; u\right)$ denote the union of $\mathfrak{S}\left(A_{1}, A_{2} ; c_{1}, c_{2} ; u\right)$ over all $-\infty<c_{1} \leq c_{2}<$ $\infty$. If $A=A(\alpha)$ is a function of just one variable, then we define $\mathfrak{S}\left(A_{1}, A_{2} ; c_{1}, c_{2}\right):=$ $\mathfrak{S}\left(A_{1}, A_{2} ; c_{1}, c_{2} ; \alpha\right)$ and $\mathfrak{S}\left(A_{1}, A_{2}\right):=\mathfrak{S}\left(A_{1}, A_{2} ; \alpha\right)$. Note that if $A(x, y)=\chi_{A}(u(x, y))$ for some $\chi_{A} \in \mathfrak{S}\left(A_{1}, A_{2}\right)$, then $A \in \mathfrak{S}\left(A_{1}, A_{2} ; u\right)$.

Let $(r, \theta)$ denote the polar coordinates for $(x, y)$. Fix $k \in \mathbb{N}$ and let $\Theta_{k} \subset \mathbb{T}$ be any set containing $2 k$ elements, where $\mathbb{T}:=[0,2 \pi)$ is the one-dimensional torus. Let $f_{\Theta} \in \mathcal{C}^{\infty}(\mathbb{T})$ such that $f_{\Theta}(\theta)=0$ if and only if $\theta \in \Theta_{k}$, and $f_{\Theta}^{\prime}(\theta) \neq 0$ whenever $\theta \in \Theta_{k}$. This means $f_{\Theta}$ changes sign across every point in $\Theta_{k}$ and is bounded away from 0 for all $\theta$ away from $\Theta_{k}$. Let $f \in \mathcal{C}^{\infty}\left(\mathbb{R}^{2}\right)$ such that $f(x, y)=f(r, \theta):=r f_{\Theta}(\theta)$ whenever $r \geq 1$.

As before, we will fix $-\infty<\tilde{E}_{1}<\tilde{E}_{2}<\infty$ and let $\Phi \in \mathcal{C}_{c}^{\infty}(\mathbb{R})$ such that $\Phi \equiv 1$ in $\left[\tilde{E}_{1}, \tilde{E}_{2}\right]$. We make the following assumption.
( $\mathbf{H} \mathbf{0}^{\prime} \mathbf{)}$ Let $H \in \operatorname{Op}\left(E S_{1,0}^{m}\right)$ such that $\Phi(H) \in \operatorname{Op}\left(S\left(\langle f(x, y), \xi, \zeta\rangle^{-\infty}\right)\right)$.

Let $\left\{\theta_{1}, \theta_{2}\right\} \subset \mathbb{T}$ such that $\theta_{1}, \theta_{2} \notin \Theta_{k}$. Let $g_{\Theta} \in \mathcal{C}^{\infty}(\mathbb{T})$ such that $g_{\Theta}(\theta)=0$ if and only if $\theta \in\left\{\theta_{1}, \theta_{2}\right\}$, and $g_{\Theta}^{\prime}\left(\theta_{j}\right) \neq 0$ for $j \in\{1,2\}$. Let $g \in \mathcal{C}^{\infty}\left(\mathbb{R}^{2}\right)$ such that $g(x, y)=g(r, \theta)=r g_{\Theta}(\theta)$ for all $r \geq 1$. Our assumptions on $f$ and $g$ imply the existence of
positive constants $C_{1}<C_{2}$ such that

$$
\begin{equation*}
C_{1}\langle x, y\rangle \leq\langle f(x, y), g(x, y)\rangle \leq C_{2}\langle x, y\rangle . \tag{2.4.2}
\end{equation*}
$$

Moreover, there exists a smooth function $\left(x_{0}, y_{0}\right): \mathbb{R} \rightarrow \mathbb{R}^{2}$ with $\lim _{|t| \rightarrow \infty}\left|\left(x_{0}(t), y_{0}(t)\right)\right|=$ $\infty$ and whose range separates the $x y$-plane into two regions, $R_{+}$and $R$, such that $g \geq 0$ in $R_{+}$. One can for example take a curve contained in $\{g=0\}$, in which case $\left(x_{0}(t), y_{0}(t)\right)=$ $\left(\beta_{1} t, \beta_{2} t\right)$ for all $|t|$ sufficiently large, with the constants $\beta_{1}$ and $\beta_{2}$ depending only on sgn $t$. Our convention is that $R_{+}$be on the right side of the curve defined by $\left(x_{0}, y_{0}\right)$ as $t$ increases. That is, if you were to move along the curve with $t$ increasing, then $R_{+}$would be on your right.

The similarity between (H0) and (H0') naturally implies the following

Lemma 2.4.1. Suppose $H$ satisfies (H0'), and let $P(x, y)=P \in \mathfrak{S}(0,1 ; g)$ and $\varphi \in$ $\mathfrak{S}\left(0,1 ; \tilde{E}_{1}, \tilde{E}_{2}\right)$. Then $[H, P] \varphi^{\prime}(H)$ is trace-class. If $\psi \in \mathcal{C}_{c}^{\infty}\left(\tilde{E}_{1}, \tilde{E}_{2}\right)$, then $q(H)[\psi(H), P]$ is trace-class for any polynomial $q$.

Proof. Replace $(x, y)$ by $(f(x, y), g(x, y))$ in the proof of Lemma 2.1.1, and use (2.4.2).
Thus we can define a junction conductivity $\sigma_{I}(H, P, \varphi)$ by (1.0.1), where the switch functions satisfy $\varphi \in \mathfrak{S}\left(0,1 ; \tilde{E}_{1}, \tilde{E}_{2}\right)$ as before and $P(x, y)=P \in \mathfrak{S}(0,1 ; g)$. The stability results from Section 2.1 extend easily to this new setting. For pedagogical purposes, we instead present an alternative method that ties the conductivity to a Fredholm index.

Theorem 2.4.2. Suppose $H$ satisfies (H0'), and let $P(x, y)=P \in \mathfrak{S}(0,1 ; g)$ and $\varphi \in$ $\mathfrak{S}\left(0,1 ; \tilde{E}_{1}, \tilde{E}_{2}\right)$. Define $U(H):=e^{i 2 \pi \varphi(H)}$ and let $\bar{P}=\bar{P}(x, y)=\chi(g(x, y)-\alpha)$, where $\chi: \mathbb{R} \rightarrow \mathbb{R}$ is the Heavyside step function and $\alpha \in \mathbb{R}$. Then $\bar{P} U(H) \bar{P}$ is a Fredholm operator on the range of $\bar{P}$, with

$$
2 \pi \sigma_{I}(H, P, \varphi)=\operatorname{Index}(\bar{P} U(H) \bar{P})
$$

Proof. We will prove that

$$
\begin{equation*}
2 \pi \sigma_{I}(H, P, \varphi)=\operatorname{Tr}[U, P] U^{*}=\operatorname{Tr}[U, \bar{P}] U^{*}=\operatorname{Index}(\bar{P} U \bar{P}) \tag{2.4.3}
\end{equation*}
$$

where we use the shorthand $U:=U(H)$.
We begin by proving the first equality in (2.4.3). Let $V:=U-I$, and observe that $V \in \mathcal{C}_{c}^{\infty}\left(\{\Phi=1\}^{\circ}\right)$. We have

$$
\operatorname{Tr}[U, P] U^{*}=\operatorname{Tr}[V, P] U^{*}=\operatorname{Tr}[V, P] V^{*}+\operatorname{Tr}[V, P] .
$$

Using that $\left[(z-H)^{-1}, P\right]=(z-H)^{-1}[H, P](z-H)^{-1}$, the Helffer-Sjöstrand formula (A.1.6) and cyclicity of the trace [57] imply that

$$
\operatorname{Tr}[V, P] V^{*}=\operatorname{Tr}\left(-\frac{1}{\pi} \int_{\mathbb{C}} \tilde{\partial} \tilde{V}(z)[H, P](z-H)^{-2} d^{2} z V^{*}(H)\right)
$$

After integrating by parts in $\partial$ and using that $\partial \tilde{V}=\tilde{V}^{\prime}$ for some almost analytic extension $\tilde{V}^{\prime}$ of $V^{\prime}$, we see that

$$
\begin{align*}
\operatorname{Tr}[V, P] V^{*}=\operatorname{Tr}[H, P] V^{\prime} V^{*} & =2 \pi \sigma_{I}(H, P)-\operatorname{Tr} 2 \pi i[H, P] \varphi^{\prime} U  \tag{2.4.4}\\
& =2 \pi \sigma_{I}(H, P, \varphi)-\operatorname{Tr}[H, P] V^{\prime}
\end{align*}
$$

We have thus shown that

$$
\operatorname{Tr}[U, P] U^{*}=2 \pi \sigma_{I}(H, P)-\operatorname{Tr}[H, P] V^{\prime}+\operatorname{Tr}[V, P] .
$$

By the same logic used in (2.4.4), we obtain that $\operatorname{Tr}[H, P] V^{\prime}=\operatorname{Tr}[H, P] V^{\prime} \Phi=\operatorname{Tr}[V, P] \Phi$,
hence

$$
\operatorname{Tr}[U, P] U^{*}=2 \pi \sigma_{I}(H, P, \varphi)+\operatorname{Tr}[V, P](1-\Phi)
$$

Again using cyclicity of the trace, this means

$$
\operatorname{Tr}[U, P] U^{*}=2 \pi \sigma_{I}(H, P, \varphi)+\operatorname{Tr} \Psi_{1}[V, P] \Psi_{2},
$$

for some $\Psi_{j}=\psi_{j}(H)$, where $\psi_{j} \in \mathcal{C}^{\infty}(\mathbb{R})$ vanishes on $\operatorname{supp}(V)$. The first equality in (2.4.3) follows.

We now prove the second equality of (2.4.3). Let $\chi(x, y)=\chi_{1}(g(x, y))$ where $\chi_{1} \in \mathcal{C}_{c}^{\infty}$ such that $\chi(\bar{P}-P)=\bar{P}-P$. Then

$$
[U, \bar{P}-P] U^{*}=[V, \bar{P}-P] U^{*}=V \chi(\bar{P}-P) U^{*}-(\bar{P}-P) \chi V U^{*}
$$

is trace class, as $V \chi$ and $\chi V$ are both trace-class by the $\Psi \mathrm{DO}$ calculus with $P-P_{1}$ and $U^{*}$ bounded. Since we know that $[U, P] U^{*}$ is trace-class, this proves that

$$
[U, \bar{P}] U^{*}=[U, \bar{P}-P] U^{*}+[U, P] U^{*}
$$

is trace-class. Now,

$$
\operatorname{Tr}[U, \bar{P}-P] U^{*}=\operatorname{Tr} V \chi(\bar{P}-P) U^{*}-\operatorname{Tr}(\bar{P}-P) \chi V U^{*},
$$

with

$$
\operatorname{Tr} V \chi(\bar{P}-P) U^{*}=\operatorname{Tr}(\bar{P}-P) U^{*} V \chi
$$

and

$$
\operatorname{Tr}(\bar{P}-P) \chi V U^{*}=\operatorname{Tr} \chi(\bar{P}-P) \chi V U^{*}=\operatorname{Tr}(\bar{P}-P) \chi V U^{*} \chi=\operatorname{Tr}(\bar{P}-P) V U^{*} \chi .
$$

Since $\left[V, U^{*}\right]=0$, we have proven that $\operatorname{Tr}[U, \bar{P}-P] U^{*}=0$, which verifies the second equality of (2.4.3).

Finally, the last equality of (2.4.3) follows immediately from [6, Proposition 2.4] and the fact that $[U, \bar{P}] U^{*}$ is trace-class.

Theorem 2.4.2 is a powerful result with many implications. For example, we immediately have the following three corollaries.

Corollary 2.4.3. Suppose $H$ satisfies $\left(H 0^{\prime}\right), P, P_{1} \in \mathfrak{S}(0,1 ; g(x, y))$ and $\varphi \in \mathfrak{S}\left(0,1 ; \tilde{E}_{1}, \tilde{E}_{2}\right)$.
Then

$$
\sigma_{I}(H, P, \varphi)=\sigma_{I}\left(H, P_{1}, \varphi\right)
$$

Proof. Applying Theorem 2.4.2, we have $2 \pi \sigma_{I}(H, P, \varphi)=\operatorname{Index}(\bar{P} U(H) \bar{P})=2 \pi \sigma_{I}\left(H, P_{1}, \varphi\right)$.

For an intuitive explanation of the above corollary, recall that our model is of two types of materials $(+$ and -$)$ that are smoothly glued together at a junction. See Figure 2.1 for two examples, where we assume for concreteness that $P$ transitions from 0 to 1 in the vicinity of $\{g=0\}$. The solid curve (in both panels) can see only one transition ( + to - and - to + in the left and right panels, respectively), while the dashed level curves contain multiple transitions $(+\rightarrow-\rightarrow+\rightarrow-$ for the left panel and $-\rightarrow+\rightarrow-\rightarrow+\rightarrow-\rightarrow+$ for the right panel). As expected, the conductivity only cares about the "starting" and "ending" topology along the level curve and is unaffected by oscillations in between. Observe that Corollary 2.4.3 describes a conservation law, as the solid and dashed curves respectively measure the conductivity entering and leaving the junction.

Corollary 2.4.4. Suppose $H$ satisfies (H0'), and let $P(x, y)=P \in \mathfrak{S}(0,1 ; g)$ and $\varphi, \varphi_{1} \in$ $\mathfrak{S}\left(0,1 ; \tilde{E}_{1}, \tilde{E}_{2}\right)$. Then

$$
\sigma_{I}(H, P, \varphi)=\sigma_{I}\left(H, P, \varphi_{1}\right)
$$

Proof. For $\mu \in[0,1]$, define $\varphi_{\mu}:=\varphi+\mu\left(\varphi_{1}-\varphi\right)$. Theorem 2.4.2 implies that $\sigma_{I}\left(H, P, \varphi_{\mu}^{\prime}\right)=$ $\operatorname{Index}\left(\bar{P} U_{\mu} \bar{P}\right)$ for all $\mu \in[0,1]$, where $U_{\mu}:=e^{i 2 \pi \varphi_{\mu}(H)}$. With $V_{\mu}:=U_{\mu}-I$, the HelfferSjöstrand formula (A.1.6) implies

$$
U_{\mu_{2}}-U_{\mu_{1}}=V_{\mu_{2}}-V_{\mu_{1}}=-\frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial}\left(\tilde{V}_{\mu_{2}}(z)-\tilde{V}_{\mu_{1}}(z)\right)(z-H)^{-1} d^{2} z
$$

Since $\left|\bar{\partial}\left(\tilde{V}_{\mu_{2}}(z)-\tilde{V}_{\mu_{1}}(z)\right)\right| \leq C\left|\mu_{2}-\mu_{1}\right|$ uniformly in $\mu_{1}, \mu_{2} \in[0,1]$ and $z \in \mathbb{C}$, it follows that $\left\|U_{\mu_{2}}-U_{\mu_{1}}\right\| \rightarrow 0$ as $\mu_{2}-\mu_{1} \rightarrow 0$. Therefore, by [55, Theorem 19.1.5], Index $\left(\bar{P} U_{\mu} \bar{P}\right)$ is independent of $\mu \in[0,1]$, and the result is complete.

Corollary 2.4.5. Let $H \in \operatorname{Op}\left(E S_{1,0}^{m}\right)$ and $W \in \operatorname{Op}\left(S_{1,0}^{m}\right)$ such that $H^{(\mu)}:=H+\mu W$ satisfies (H0') for all $\mu \in[0,1]$. Let $P(x, y)=P \in \mathfrak{S}(0,1 ; g)$ and $\varphi \in \mathfrak{S}\left(0,1 ; \tilde{E}_{1}, \tilde{E}_{2}\right)$. Then $\sigma_{I}\left(H^{(1)}, P, \varphi\right)=\sigma_{I}(H, P, \varphi)$.

As before, in the following proofs we write $A \in \operatorname{Op}(\mu S(\mathfrak{m}))$ to mean that $A=\operatorname{Op}(a)$ with $a \in S(\mathfrak{m})$ for all $\mu$ and $\left|\partial^{\alpha} a\right| \leq C_{\alpha} \mu \mathfrak{m}$ uniformly in $\mu$.

Proof. We use the shorthand $\sigma_{I}\left(H^{(\mu)}\right):=\sigma_{I}\left(H^{(\mu)}, P, \varphi\right)$. By definition, $\Phi \in \mathcal{C}_{c}^{\infty}\left(E_{1}, E_{2}\right)$ satisfies $\varphi^{\prime} \Phi=\varphi^{\prime}$. We see that

$$
\begin{aligned}
& \sigma_{I}\left(H^{\left(\mu_{2}\right)}\right)-\sigma_{I}\left(H^{\left(\mu_{1}\right)}\right)= \\
& \quad \operatorname{Tr} i\left[H^{\left(\mu_{2}\right)}, P\right]\left(\varphi^{\prime}\left(H^{\left(\mu_{2}\right)}\right)-\varphi^{\prime}\left(H^{\left(\mu_{1}\right)}\right)\right)+\left(\mu_{2}-\mu_{1}\right) \operatorname{Tr} i[W, P] \varphi^{\prime}\left(H^{\left(\mu_{1}\right)}\right),
\end{aligned}
$$

with

$$
\varphi^{\prime}\left(H^{\left(\mu_{2}\right)}\right)-\varphi^{\prime}\left(H^{\left(\mu_{1}\right)}\right)=\frac{1}{\pi}\left(\mu_{2}-\mu_{1}\right) \int_{\mathbb{C}} \bar{\partial} \tilde{\varphi}^{\prime}(z)\left(z-H^{\left(\mu_{2}\right)}\right)^{-1} W\left(z-H^{\left(\mu_{1}\right)}\right)^{-1} d^{2} z
$$

by the Helffer-Sjöstrand formula (A.1.6). Observe that

$$
\varphi^{\prime}\left(H^{\left(\mu_{2}\right)}\right)-\varphi^{\prime}\left(H^{\left(\mu_{1}\right)}\right) \in \mathrm{Op}\left(\left|\mu_{2}-\mu_{1}\right| S\left(\langle x, y, \xi, \zeta\rangle^{m}\right)\right) \cap \operatorname{Op}\left(S\left(\langle f(x, y), \xi, \zeta\rangle^{-\infty}\right)\right)
$$

and thus

$$
\varphi^{\prime}\left(H^{\left(\mu_{2}\right)}\right)-\varphi^{\prime}\left(H^{\left(\mu_{1}\right)}\right) \in \operatorname{Op}\left(\left|\mu_{2}-\mu_{1}\right|^{1 / 2} S\left(\langle f(x, y), \xi, \zeta\rangle^{-\infty}\right)\right)
$$

by interpolation. Since $\left[H^{\left(\mu_{2}\right)}, P\right] \in \operatorname{Op}\left(S\left(\langle x, y, \xi, \zeta\rangle^{m}\langle g(x, y)\rangle^{-\infty}\right)\right)$, the composition calculus implies that

$$
\left|\operatorname{Tr} i\left[H^{\left(\mu_{2}\right)}, P\right]\left(\varphi^{\prime}\left(H^{\left(\mu_{2}\right)}\right)-\varphi^{\prime}\left(H^{\left(\mu_{1}\right)}\right)\right)\right| \leq C\left|\mu_{2}-\mu_{1}\right|^{1 / 2} .
$$

Since $\left|\operatorname{Tr} i[W, P] \varphi^{\prime}\left(H^{\left(\mu_{1}\right)}\right)\right| \leq C$ independent of $\mu_{2}$, we have shown that $\mu \mapsto \sigma_{I}\left(H^{(\mu)}\right)$ is continuous on $\mu \in[0,1]$. But Theorem 2.4.2 implies that $2 \pi \sigma_{I}\left(H^{(\mu)}\right)$ is equal to an (integervalued) Fredholm index for all $\mu$, and thus $\sigma_{I}\left(H^{(\mu)}\right)$ is independent of $\mu \in[0,1]$.

Analogous to (H1), we now introduce the following class of symbols.
(H1') Let $H=\operatorname{Op}(\sigma)$ with $\sigma \in E S_{1,0}^{m}$. Suppose there exist symbols $\sigma_{ \pm} \in S_{1,0}^{m}$ independent of $(x, y)$ with no spectrum in the open interval $\left(E_{1}, E_{2}\right)$, such that $\sigma=\sigma_{ \pm}$whenever $\pm f(x, y)>0$ is sufficiently large.

We then have

Proposition 2.4.6. Suppose $H=\operatorname{Op}(\sigma)$ satisfies ( $\left.H 1^{\prime}\right)$ and $\Phi \in \mathcal{C}_{c}^{\infty}\left(E_{1}, E_{2}\right)$. Define $H_{h}:=\operatorname{Op}_{h}(\sigma)$ for $h \in(0,1]$. Then $\Phi\left(H_{h}\right) \in \operatorname{Op}_{h}\left(S\left(\langle f(x, y), \xi, \zeta\rangle^{-\infty}\right)\right)$.

Proof. See the proof of Proposition 2.1.10.

Proposition 2.4.6 proves that the class (H1') is a subset of (H0'). We now show that under ( $\mathrm{H} 1^{\prime}$ ), the interface conductivity is stable with respect to semi-classical rescaling.

Theorem 2.4.7. Suppose $H=\mathrm{Op}(\sigma)$ satisfies ( $\left.H 1^{\prime}\right)$, and define $H_{h}:=\mathrm{Op}_{h}(\sigma)$. Let $P(x, y)=P \in \mathfrak{S}(0,1 ; g)$ and $\varphi \in \mathfrak{S}\left(0,1 ; E_{1}, E_{2}\right)$. Then $\sigma_{I}\left(H_{h}, P, \varphi\right)=\sigma_{I}(H, P, \varphi)$ for all $h \in(0,1]$.

Proof. For $h \in(0,1]$, define $U_{h}:=e^{i 2 \pi \varphi\left(H_{h}\right)}$. Proposition 2.4.6 and Theorem 2.4.2 imply that $2 \pi \sigma_{I}\left(H_{h}, P, \varphi\right)=\operatorname{Index}\left(\bar{P} U_{h} \bar{P}\right)$ for all $h \in(0,1]$. It thus suffices to show that for any fixed $h \in(0,1]$, $\operatorname{Index}\left(\bar{P} U_{h^{\prime}} \bar{P}\right)$ is constant over $h^{\prime}$ in an open neighborhood of $h$. Using the Helffer-Sjöstrand formula (A.1.6), we write

$$
U_{h^{\prime}}-U_{h}=\frac{1}{\pi} \int_{\mathbb{R}} \bar{\partial} \tilde{V}(z)\left(z-H_{h^{\prime}}\right)^{-1}\left(H_{h^{\prime}}-H_{h}\right)\left(z-H_{h}\right)^{-1} d^{2} z
$$

Since $\left\|\left(z-H_{h^{\prime}}\right)^{-1}\right\| \leq|\Im z|^{-1}$ and $H_{h^{\prime}}-H_{h} \in \operatorname{Op}\left(\left|h^{\prime}-h\right| S\left(\langle\xi, \zeta\rangle^{m}\right)\right)$ with $\left(z-H_{h}\right)^{-1} \in$ $\operatorname{Op}\left(S\left(\langle\xi, \zeta\rangle^{-m}\right)\right)$ for all $\Im z \neq 0$ (with bounds growing at most algebraically in $|\Im z|^{-1}$ ), the rapid decay of $\bar{\partial} \tilde{V}(z)$ near the real axis implies that $\left\|U_{h^{\prime}}-U_{h}\right\| \leq C\left|h^{\prime}-h\right|$. The result then follows from [55, Theorem 19.1.5].

### 2.4.1 Bulk-interface correspondence

As in Section 2.2, the above results allow for an accessible formula for $\sigma_{I}(H, P, \varphi)$. In particular, we will prove a bulk-interface correspondence, extending Corollary 2.2.2 to the junction setting.

For $(t, \xi, \zeta) \in \mathbb{R}^{3}$, define

$$
\begin{equation*}
\tau(t, \xi, \zeta):=\sigma\left(x_{0}(t), y_{0}(t), \xi, \zeta\right), \quad \tau_{z}:=z-\tau \tag{2.4.5}
\end{equation*}
$$

where we recall the definition of $\left(x_{0}, y_{0}\right)$ below (2.4.2). By (H1'), there exist $t_{0}>0$ and symbols $\tau_{ \pm} \in\left\{\sigma_{+}, \sigma_{-}\right\}$such that $\tau(t, \xi, \zeta)=\tau_{ \pm}(\xi, \zeta)$ whenever $\pm t \geq t_{0}$.

Theorem 2.4.8. Suppose $H$ satisfies (H1'), $P(x, y)=P \in \mathfrak{S}(0,1 ; g)$ and $\varphi \in \mathfrak{S}\left(0,1 ; E_{1}, E_{2}\right)$. Let $\alpha \in\left(E_{1}, E_{2}\right)$ and define $\tau_{ \pm, z}:=z-\tau_{ \pm}$with $z=\alpha+i \omega$. Then

$$
2 \pi \sigma_{I}(H, P, \varphi)=\frac{i}{8 \pi^{2}}\left(I_{+}-I_{-}\right), \quad I_{ \pm}=\int_{\mathbb{R}^{3}} \operatorname{tr}\left[\tau_{ \pm, z}^{-1} \partial_{\xi} \tau_{ \pm, z}, \tau_{ \pm, z}^{-1} \partial_{\zeta} \tau_{ \pm, z}\right] \tau_{ \pm, z}^{-1} d \omega d \xi d \zeta
$$

As with Theorem 2.2.1, we will expand $\sigma_{I}\left(H_{h}, P, \varphi\right)$ in the semiclassical parameter $h$ and apply Theorem 2.4.7 to eliminate terms that are not $O(1)$.

Proof. By Theorem 2.4.4, $\sigma_{I}(H, P, \varphi)$ is independent of $\varphi \in \mathfrak{S}\left(0,1 ; E_{1}, E_{2}\right)$, thus we can take $\varphi^{\prime} \in \mathcal{C}_{c}^{\infty}\left(\alpha_{0}, \alpha\right)$ for some $\alpha_{0}>E_{1}$. We will use the shorthand $\sigma_{I}:=\sigma_{I}(H, P, \varphi)$.

Let $\mathrm{Op}_{h} \nu_{h}:=\varphi^{\prime}\left(H_{h}\right)$. By Proposition A.3.3, we have

$$
\nu_{h}+\frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial}_{\varphi^{\prime}}(z) \tilde{q}_{z, h} d^{2} z \in S^{-3 / 2}\left(\langle f(x, y), \xi, \zeta\rangle^{-\infty}\right)
$$

where

$$
\tilde{q}_{z, h}=\sigma_{z}^{-1}+\frac{i h}{2}\left\{\sigma_{z}^{-1}, \sigma_{z}\right\} \sigma_{z}^{-1}, \quad\{a, b\}:=\partial_{\xi} a \partial_{x} b+\partial_{\zeta} a \partial_{y} b-\partial_{x} a \partial_{\xi} b-\partial_{y} a \partial_{\zeta} b
$$

and $\sigma_{z}:=z-\sigma$. With $\mathrm{Op}_{h}\left(\kappa_{h}\right):=\left[H_{h}, P\right]$, we have that

$$
\kappa_{h}+i h k_{1}-\frac{h^{2}}{4} k_{2} \in S^{-3}\left(\langle g(x, y)\rangle^{-\infty}\langle\xi, \zeta\rangle^{m}\right)
$$

where

$$
\begin{equation*}
k_{1}:=\partial_{\xi} \sigma \partial_{x} P+\partial_{\zeta} \sigma \partial_{y} P, \quad k_{2}:=\partial_{\xi \xi} \sigma \partial_{x x} P+2 \partial_{\xi \zeta} \sigma \partial_{x y} P+\partial_{\zeta \zeta} \sigma \partial_{y y} P . \tag{2.4.6}
\end{equation*}
$$

Since $\nu_{h} \in S\left(\langle f(x, y), \xi, \zeta\rangle^{-\infty}\right)$ and $\kappa_{h} \in S^{-1}\left(\langle g(x, y)\rangle^{-\infty}\langle\xi, \zeta\rangle^{m}\right)$, the composition calculus implies that

$$
\kappa_{h} \sharp_{h} \nu_{h}-\kappa_{h} \nu_{h}+\frac{i h}{2}\left\{\kappa_{h}, \nu_{h}\right\} \in S^{-3}\left(\langle x, y, \xi, \zeta\rangle^{-\infty}\right),
$$

with $S^{-3}$ (rather than $S^{-2}$ ) above because $\kappa_{h}$ is $O(h)$ in $S\left(\langle g(x, y)\rangle^{-\infty}\langle\xi, \zeta\rangle^{m}\right)$. Therefore,

$$
\sigma_{I}=\frac{i}{(2 \pi h)^{2}} \operatorname{tr} \int_{\mathbb{R}^{4}} \kappa_{h} \sharp H_{h} \nu_{h} d R_{4}=\frac{i}{(2 \pi h)^{2}} \operatorname{tr} \int_{\mathbb{R}^{4}}\left(\kappa_{h} \nu_{h}-\frac{i h}{2}\left\{\kappa_{h}, \nu_{h}\right\}\right) d R_{4}+o(1)
$$

as $h \rightarrow 0$, with $d R_{4}:=d x d y d \xi d \zeta$. Since

$$
\kappa_{h} \nu_{h}=\left(i h k_{1}-\frac{h^{2}}{4} k_{2}\right) \frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\varphi}^{\prime}(z)\left(\sigma_{z}^{-1}+\frac{i h}{2}\left\{\sigma_{z}^{-1}, \sigma_{z}\right\} \sigma_{z}^{-1}\right) d^{2} z+h^{5 / 2} a_{h}
$$

for some $a_{h} \in S\left(\langle x, y, \xi, \zeta\rangle^{-\infty}\right)$ and

$$
\left\{\kappa_{h}, \nu_{h}\right\}=\left\{i h k_{1}, \frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\varphi}^{\prime}(z) \sigma_{z}^{-1} d^{2} z\right\}+h^{2} b_{h}, \quad b_{h} \in S\left(\langle x, y, \xi, \zeta\rangle^{-\infty}\right)
$$

it follows that

$$
\begin{array}{r}
\sigma_{I}=\frac{i}{(2 \pi h)^{2}} \frac{1}{\pi} \operatorname{tr} \int_{\mathbb{R}^{4}} \int_{\mathbb{C}} \bar{\partial} \tilde{\varphi}^{\prime}(z)\left(i h k_{1} \sigma_{z}^{-1}-\frac{h^{2}}{2} k_{1}\left\{\sigma_{z}^{-1}, \sigma_{z}\right\} \sigma_{z}^{-1}\right. \\
\left.-\frac{h^{2}}{4} k_{2} \sigma_{z}^{-1}+\frac{h^{2}}{2}\left\{k_{1}, \sigma_{z}^{-1}\right\}\right) d^{2} z d R_{4}+o(1)
\end{array}
$$

as $h \rightarrow 0$. Since $\sigma_{I}$ is independent of $h$, it follows that the $O\left(h^{-1}\right)$ term above vanishes, and
thus

$$
\sigma_{I}=\frac{i}{(2 \pi)^{3}} \operatorname{tr} \int_{\mathbb{R}^{4}} \int_{\mathbb{C}} \bar{\partial} \tilde{\varphi}^{\prime}(z)\left(-k_{1} \sigma_{z}^{-1}\left\{\sigma_{z}, \sigma_{z}^{-1}\right\}-\frac{1}{2} k_{2} \sigma_{z}^{-1}+\left\{k_{1}, \sigma_{z}^{-1}\right\}\right) d^{2} z d R_{4}
$$

Observe that whenever $\langle f(x, y), \xi, \zeta\rangle$ is sufficiently large, $z \mapsto \sigma_{z}^{-1}$ is holomorphic and thus the above integral over $z$ vanishes (this is verified via an integration by parts in $\bar{\partial}$ ). Using that $k_{1}$ and $k_{2}$ vanish whenever $\langle g(x, y)\rangle$ is sufficiently large, we can replace the above integration limit $\mathbb{R}^{4}$ by a sufficiently large rectangle $B:=B_{x y} \times B_{\xi \zeta} \subset \mathbb{R}^{2} \times \mathbb{R}^{2}$. We require that $B$ contain all points $\left(x_{0}(t), y_{0}(t), \xi, \zeta\right)$ for which $\sigma\left(x_{0}(t), y_{0}(t), \xi, \zeta\right)$ has an eigenvalue of $\alpha$. Moreover, assume that $B_{x y}$ contains $\left(x_{0}\left(t_{0}\right), y_{0}\left(t_{0}\right)\right)$ and $\left(x_{0}\left(-t_{0}\right), y_{0}\left(-t_{0}\right)\right)$, where we recall the respective definitions of $t_{0}$ and $\left(x_{0}, y_{0}\right)$ above Theorem 2.4.8 and below (2.4.2).

At this point, $\left\{k_{1}, \sigma_{z}^{-1}\right\}$ can be written in divergence form and the corresponding term converted (via integration by parts in $(x, y, \xi, \zeta)$ ) to an integral over the surface $\partial B$. An integration by parts in $\bar{\partial}$ then reveals that the contribution of this term vanishes. Thus we are left with

$$
\begin{equation*}
\sigma_{I}=-\frac{i}{(2 \pi)^{3}} \operatorname{tr} \int_{B} \int_{\mathbb{C}} \bar{\partial} \tilde{\varphi}^{\prime}(z)\left(k_{1} \sigma_{z}^{-1}\left\{\sigma_{z}, \sigma_{z}^{-1}\right\}+\frac{1}{2} k_{2} \sigma_{z}^{-1}\right) d^{2} z d R_{4} \tag{2.4.7}
\end{equation*}
$$

We now simplify the first term above. Integrating by parts in $\bar{\partial}$ with $z=: \lambda+i \omega$, we see that

$$
\begin{aligned}
\sigma_{I, 1} & :=\frac{-i}{(2 \pi)^{3}} \operatorname{tr} \int_{B} \int_{\mathbb{C}} \bar{\partial} \tilde{\varphi}^{\prime}(z) k_{1} \sigma_{z}^{-1}\left\{\sigma_{z}, \sigma_{z}^{-1}\right\} d^{2} z d R_{4} \\
& =\left.\frac{-1}{2(2 \pi)^{3}} \operatorname{tr} \int_{B} \int_{\alpha_{0}}^{\alpha} \varphi^{\prime}(\lambda) k_{1} \sigma_{z}^{-1}\left\{\sigma_{z}, \sigma_{z}^{-1}\right\}\right|_{\omega=0^{-}} ^{\omega=0^{+}} d \lambda d R_{4}
\end{aligned}
$$

Let $\mathcal{L}:=\left\{x_{0}(t), y_{0}(t): t \in \mathbb{R}\right\}$ be the range of $\left(x_{0}, y_{0}\right)$. For simplicity, assume that $\mathcal{L} \subset$
$g^{-1}(0)$. By cyclicity of the trace and recalling the definition (2.4.6) of $k_{1}$, it follows that

$$
\operatorname{tr} k_{1}\left\{\sigma_{z}^{-1}, \sigma_{z}\right\} \sigma_{z}^{-1}=-\operatorname{tr}\left(\partial_{x} P \partial_{\xi} \sigma_{z}\left\{\sigma_{z}^{-1}, \sigma_{z}\right\}_{\zeta, y}+\partial_{y} P \partial_{\zeta} \sigma_{z}\left\{\sigma_{z}^{-1}, \sigma_{z}\right\}_{\xi, x}\right) \sigma_{z}^{-1}
$$

where we have defined $\{a, b\}_{\alpha, \beta}:=\partial_{\alpha} a \partial_{\beta} b-\partial_{\beta} a \partial_{\alpha} b$ and used the fact that $\left(\partial_{\xi} \sigma, \partial_{\zeta} \sigma\right)=$ $-\left(\partial_{\xi} \sigma_{z}, \partial_{\zeta} \sigma_{z}\right)$. Integrating by parts in $x$ (first term) and $y$ (second term), we obtain

$$
\begin{aligned}
\sigma_{I, 1}= & -\frac{1}{2(2 \pi)^{3}} \operatorname{tr} \int_{B} \int_{\alpha_{0}}^{\alpha} \varphi^{\prime}(\lambda) P
\end{aligned} \begin{aligned}
& \left(\partial_{x}\left(\partial_{\xi} \sigma_{z}\left\{\sigma_{z}^{-1}, \sigma_{z}\right\}_{\zeta, y} \sigma_{z}^{-1}\right)\right. \\
& \left.+\partial_{y}\left(\partial_{\zeta} \sigma_{z}\left\{\sigma_{z}^{-1}, \sigma_{z}\right\}_{\xi, x} \sigma_{z}^{-1}\right)\right)\left.\right|_{\omega=0^{-}} ^{\omega=0^{+}} d \lambda d R_{4} \\
& +\frac{1}{2(2 \pi)^{3}} \operatorname{tr} \int_{B_{\xi \zeta}} \int_{B_{x y} \cap \mathcal{L}} \int_{\alpha_{0}}^{\alpha} \varphi^{\prime}(\lambda)\left(\partial_{\xi} \sigma_{z}\left\{\sigma_{z}^{-1}, \sigma_{z}\right\}_{\zeta, y} \nu_{x}\right. \\
& \left.+\partial_{\zeta} \sigma_{z}\left\{\sigma_{z}^{-1}, \sigma_{z}\right\}_{\xi, x} \nu_{y}\right)\left.\sigma_{z}^{-1}\right|_{\omega=0^{-}} ^{\omega=0^{+}} d \lambda d \ell d R_{2} \\
= & \sigma_{I, 10}+\sigma_{I, 11}
\end{aligned}
$$

where $d R_{2}:=d \xi d \zeta$ and $d \ell$ is the integration measure on $\mathcal{L}$. Here, $\nu$ is the unit vector (outwardly) normal to the surface $\partial\left(\{g(x, y) \leq 0\} \cap B_{x y}\right)$. Note that the other surface terms do not contribute; over $\left(\partial\left(\{g(x, y) \leq 0\} \cap B_{x y}\right)\right) \backslash\left(B_{x y} \cap \mathcal{L}\right)$, either $P=0$ or the map $z \mapsto \sigma_{z}^{-1}$ is holomorphic and thus the difference between $\omega=0^{+}$and $\omega=0^{-}$vanishes.

We now verify that the volume term $\sigma_{I, 10}$ vanishes. First, observe that

$$
\left.\int_{B} \int_{\alpha_{0}}^{\alpha} \varphi^{\prime}(\lambda) P\left(\partial_{\xi}\left(\partial_{x} \sigma_{z}\left\{\sigma_{z}^{-1}, \sigma_{z}\right\}_{\zeta, y} \sigma_{z}^{-1}\right)+\partial_{\zeta}\left(\partial_{y} \sigma_{z}\left\{\sigma_{z}^{-1}, \sigma_{z}\right\}_{\xi, x} \sigma_{z}^{-1}\right)\right)\right|_{\omega=0^{-}} ^{\omega=0^{+}} d \lambda d R_{4}=0
$$

as we can use the fact that $P$ is independent of $(\xi, \zeta)$ to integrate the above left-hand side by parts in $\xi$ (first term) and $\zeta$ (second term) to obtain integrals over $\partial B_{\xi \zeta}$, over which
$z \mapsto \sigma_{z}^{-1}$ is holomorphic. Therefore,

$$
\begin{aligned}
\sigma_{I, 10}=-\frac{1}{2(2 \pi)^{3}} & \int_{B} \int_{\alpha_{0}}^{\alpha} \varphi^{\prime}(\lambda) P \operatorname{tr}\left(\partial_{x}\left(\partial_{\xi} \sigma_{z}\left\{\sigma_{z}^{-1}, \sigma_{z}\right\}_{\zeta, y} \sigma_{z}^{-1}\right)+\partial_{y}\left(\partial_{\zeta} \sigma_{z}\left\{\sigma_{z}^{-1}, \sigma_{z}\right\}_{\xi, x} \sigma_{z}^{-1}\right)\right. \\
& \left.-\partial_{\xi}\left(\partial_{x} \sigma_{z}\left\{\sigma_{z}^{-1}, \sigma_{z}\right\}_{\zeta, y} \sigma_{z}^{-1}\right)-\partial_{\zeta}\left(\partial_{y} \sigma_{z}\left\{\sigma_{z}^{-1}, \sigma_{z}\right\}_{\xi, x} \sigma_{z}^{-1}\right)\right)\left.\right|_{\omega=0^{-}} ^{\omega=0^{+}} d \lambda d R_{4}
\end{aligned}
$$

A brute force calculation reveals that the above trace vanishes, and thus indeed $\sigma_{I, 10}=0$.
For $\sigma_{I, 11}$, we again use cyclicity of the trace to verify that

$$
\begin{aligned}
\operatorname{tr}\left(\partial_{\xi} \sigma_{z}\left\{\sigma_{z}^{-1}, \sigma_{z}\right\}_{\zeta, y} \sigma_{z}^{-1} \nu_{x}\right. & \left.+\partial_{\zeta} \sigma_{z}\left\{\sigma_{z}^{-1}, \sigma_{z}\right\}_{\xi, x} \sigma_{z}^{-1} \nu_{y}\right) \\
& =\operatorname{tr}\left(\left(\nu_{x} \partial_{y} \sigma_{z}-\nu_{y} \partial_{x} \sigma_{z}\right)\left\{\sigma_{z}^{-1}, \sigma_{z}\right\}_{\xi, \zeta} \sigma_{z}^{-1}\right)
\end{aligned}
$$

We recognize the first factor on the above right-hand side as the derivative of $\sigma_{z}$ in the direction of $\mathcal{L}$ (with $t$ increasing). Recalling the definition of $\tau$ in (2.4.5), we see that

$$
\begin{aligned}
& \partial_{t} \tau(t, \xi, \zeta)=x_{0}^{\prime}(t) \partial_{x} \sigma\left(x_{0}(t), y_{0}(t), \xi, \zeta\right)+y_{0}^{\prime}(t) \partial_{y} \sigma\left(x_{0}(t), y_{0}(t), \xi, \zeta\right), \\
& \partial_{\xi} \tau(t, \xi, \zeta)=\partial_{\xi} \sigma\left(x_{0}(t), y_{0}(t), \xi, \zeta\right), \quad \partial_{\zeta} \tau(t, \xi, \zeta)=\partial_{\zeta} \sigma\left(x_{0}(t), y_{0}(t), \xi, \zeta\right),
\end{aligned}
$$

with $\left(x_{0}^{\prime}, y_{0}^{\prime}\right)=N\left(-\nu_{y}, \nu_{x}\right)$ and $d \ell=N d t$, where $N:=\sqrt{\left(x_{0}^{\prime}(t)\right)^{2}+\left(y_{0}^{\prime}(t)\right)^{2}}$. We conclude that

$$
\begin{equation*}
\sigma_{I, 1}=\left.\frac{1}{16 \pi^{3}} \operatorname{tr} \int_{R} \int_{\left[\alpha_{0}, \alpha\right]} \varphi^{\prime}(\lambda) \partial_{t} \tau_{z}\left\{\tau_{z}^{-1}, \tau_{z}\right\}_{\xi, \zeta} \tau_{z}^{-1}\right|_{\omega=0^{-}} ^{\omega=0^{+}} d \lambda d R_{3}, \tag{2.4.8}
\end{equation*}
$$

where $R=\left(t_{1}, t_{2}\right) \times B_{\xi \zeta}$ with $t_{1}=\inf \left\{t:\left(x_{0}(t), y_{0}(t)\right) \in B_{x y}\right\}$ and $t_{2}=\sup \{t:$ $\left.\left(x_{0}(t), y_{0}(t)\right) \in B_{x y}\right\}$, and $d R_{3}:=d t d \xi d \zeta$.

We next eliminate $\varphi^{\prime}$ from (2.4.8). Since $\partial_{t} \tau_{z}\left\{\tau_{z}^{-1}, \tau_{z}\right\}_{\xi, \zeta} \tau_{z}^{-1} \rightarrow 0$ as $|\omega| \rightarrow \infty$, we have

$$
\left.\partial_{t} \tau_{z}\left\{\tau_{z}^{-1}, \tau_{z}\right\}_{\xi, \zeta} \tau_{z}^{-1}\right|_{\omega=0^{-}} ^{\omega=0^{+}}=-\int_{-\infty}^{+\infty} \partial_{\omega}\left(\partial_{t} \tau_{z}\left\{\tau_{z}^{-1}, \tau_{z}\right\}_{\xi, \zeta} \tau_{z}^{-1}\right) d \omega
$$

Cyclicity of the trace and the fact that $\partial_{\omega} \tau_{z}=i$ imply that

$$
\begin{equation*}
\operatorname{tr} \partial_{\omega}\left(\partial_{t} \tau_{z}\left\{\tau_{z}^{-1}, \tau_{z}\right\}_{\xi, \zeta} \tau_{z}^{-1}\right)=-i \operatorname{tr} \varepsilon_{i j k} \partial_{k}\left(\tau_{z}^{-1} \partial_{i} \tau_{z} \tau_{z}^{-1} \partial_{j} \tau_{z} \tau_{z}^{-1}\right) \tag{2.4.9}
\end{equation*}
$$

where $\varepsilon_{i j k}$ is the anti-symmetric tensor with $\varepsilon_{123}=1$, and the variables are identified by $(1,2,3)=(\xi, \zeta, t)$. Pulling $\partial_{k}$ out of the integral over $\omega$ and integrating by parts, we get

$$
\left.\operatorname{tr} \int_{R} \partial_{t} \tau_{z}\left\{\tau_{z}^{-1}, \tau_{z}\right\}_{\xi, \zeta} \tau_{z}^{-1}\right|_{\omega=0^{-}} ^{\omega=0^{+}} d R_{3}=i \int_{\partial R} \int_{-\infty}^{+\infty} \Theta d \omega d \Sigma,
$$

where

$$
\Theta:=\operatorname{tr} \varepsilon_{i j k} \tau_{z}^{-1} \partial_{i} \tau_{z} \tau_{z}^{-1} \partial_{j} \tau_{z} \tau_{z}^{-1} \nu_{k},
$$

and $\nu$ is the outward unit normal vector to the surface $\partial R$ with $\Sigma$ the Euclidean surface measure in $\mathbb{R}^{3}$. Thus we have shown that

$$
\sigma_{I, 1}=\frac{i}{16 \pi^{3}} \int_{\left[\alpha_{0}, \alpha\right]} \varphi^{\prime}(\lambda) \int_{\partial R} \int_{-\infty}^{+\infty} \Theta d \omega d \Sigma d \lambda
$$

Integrating by parts in $\lambda$, we obtain

$$
\sigma_{I, 1}=\frac{i}{16 \pi^{3}} \int_{\partial R} \int_{-\infty}^{+\infty} \Theta d \omega d \Sigma
$$

with now $z=\alpha+i \omega$ in the above integrand (and from now on). The fact that only the boundary term survives follows from analyticity of $\Theta$ in $z$ over the region of integration (so that $\left.\partial_{\lambda} \Theta=-i \partial_{\omega} \Theta\right)$.

Recall that $R=\left(-t_{0}, t_{0}\right) \times(-M, M)^{2}$, where $M>0$ can be chosen as large as necessary. Observe that $|\Theta| \leq C\langle\omega\rangle^{-3}$ and $|\Theta| \leq C\langle\xi, \zeta\rangle^{-m-2}$ uniformly in $(t, \xi, \zeta) \in \partial R$ and $M>0$ sufficiently large, hence $|\Theta| \leq C\langle\omega\rangle^{-3 / 2}\langle\xi, \zeta\rangle^{-\frac{m+2}{2}}$ by interpolation. It follows that
$\int_{-\infty}^{+\infty}|\Theta| d \omega \leq C\langle\xi, \zeta\rangle^{-\frac{m+2}{2}}$. Therefore, sending $M \rightarrow \infty$, we see that the contributions to $\sigma_{I, 1}$ from the sides of $\partial R$ with normal vector in the $\xi$ and $\zeta$ directions vanish. Indeed, the area of these surfaces is proportional to $M$, with the maximum of the integrand bounded by $C M^{-1-m / 2}$. Thus we are left with integrals over the sides corresponding to $t= \pm t_{0}$, over which $\tau(t, \xi, \zeta)=\tau_{ \pm}(\xi, \zeta)$. As a consequence, $\sigma_{I, 1}=\frac{i}{16 \pi^{3}}\left(I_{+}-I_{-}\right)$.

Recalling (2.4.7), it remains to show that

$$
\begin{equation*}
\sigma_{I, 2}:=-\frac{i}{2(2 \pi)^{3}} \operatorname{tr} \int_{\mathbb{R}^{4}} k_{2} \int_{\mathbb{C}} \bar{\partial} \tilde{\varphi}^{\prime}(z) \sigma_{z}^{-1} d^{2} z d R_{4}=0 . \tag{2.4.10}
\end{equation*}
$$

Define $P_{\varepsilon}(x, y):=P(\varepsilon(x-\tilde{x}, y-\tilde{y}))$, where $(\tilde{x}, \tilde{y}) \in \mathbb{R}^{2}$ is chosen such that $P_{\varepsilon}\left(x_{0}(t), y_{0}(t)\right)=1$ for all $\varepsilon>0$ and $t \in \mathbb{R}$. Corollary 2.4.3 implies that $\sigma_{I}\left(H, P_{\varepsilon}\right)$ is independent of $\varepsilon$, while $\sigma_{I, 1}$ is independent of $\varepsilon$ since $I_{ \pm}$are. It follows that $\sigma_{I, 2}$ must also be independent of $\varepsilon>0$.

We will thus replace $P$ by $P_{\varepsilon}$ and show that $\sigma_{I, 2} \rightarrow 0$ as $\varepsilon \downarrow 0$. As stated in the paragraph below (2.2.4), the integral over $\mathbb{C}$ in (2.4.10) vanishes whenever $\langle f(x, y), \xi, \zeta\rangle$ is sufficiently large (uniformly in $\varepsilon$ ). From the definition (2.4.6), it follows that $k_{2}=\varepsilon^{2} \tilde{k}_{2, \varepsilon}$, where $\tilde{k}_{2, \varepsilon}$ vanishes whenever $\langle\varepsilon g(x-\tilde{x}, y-\tilde{y})\rangle$ is sufficiently large. We conclude that there exist positive constants $M_{1}$ and $M_{2}$ such that

$$
\left|\sigma_{I, 2}\right| \leq C \varepsilon^{2} \operatorname{Vol}\left(\left\{\langle f(x, y)\rangle \leq M_{1}\right\} \cap\left\{\langle\varepsilon g(x-\tilde{x}, y-\tilde{y})\rangle \leq M_{2}\right\}\right) \leq C \varepsilon
$$

with the above volume taken in the $x y$-plane. We have thus shown that $\sigma_{I, 2}=0$, and the proof is complete.

An immediate consequence of Theorem 2.4.8 is

Corollary 2.4.9. Let $\tilde{H}=\operatorname{Op}(\tau)$ with $\tau=\tau(y, \xi, \zeta)$ and $\tilde{P}(x)=\tilde{P} \in \mathfrak{S}(0,1)$. Then $\sigma_{I}(H, P)=\sigma_{I}(\tilde{H}, \tilde{P})$.

Proof. Apply Theorem 2.4.8 to $H=\operatorname{Op}(\sigma)$ and $\tilde{H}=\operatorname{Op}(\tau)$.


Figure 2.2: Triangular domains form in mechanically relaxed $t$ BLG. We highlight two regions of interest for our analysis. Region 1 corresponds to an edge, resembling the setting in Sections 2.1-2.3. Region 2 is at a junction, where the theory from this subsection applies.

Observe that Corollary 2.4.9 reduces the junction conductivity to the simpler setting of a flat interface analyzed in Sections 2.1-2.3. The insulating materials are described by $\tau_{ \pm}=\tau( \pm \infty, \xi, \zeta)$.

### 2.4.2 Applications

Twisted bilayer graphene (tBLG) is widely studied for its unique mechanical and electronic properties including the magic angle superconductivity [23, 18]. Upon gating, it acts as host of a network of topological interface channels, which can be seen experimentally and theoretically $[74,72,24,1]$. tBLG is constructed by taking two periodic 2 D sheets of graphene and stacking them with a relative twist, typically small. The atoms relax to minimize energy, forming large triangular regions of the energetically favorable AB and BA Bernal stacking [80, 24, 31]. It is relevant to note this asymmetric transport under gating is a separate phenomena from magic angle superconductivity. Indeed, the asymmetric transport phenomena only requires sufficiently small twist angles and vertical gating, i.e. by inducing a potential difference between the two layers, while superconductivity can only occur precisely at the magic twist angles. In this section, we make use of Figure 2.2 for discussion of the geometry. The interior of one of these triangles can be considered as approximately an
infinite periodic material as the side of one of these triangles scales inversely proportional to the small twist angle [24].

Following the Bistritzer-MacDonald model [18], the junction Hamiltonian corresponding to Region 2 from Figure 2.2 is

$$
\operatorname{Op}(\sigma)=H=\left(\begin{array}{cc}
\Omega I+D \cdot \sigma^{(\eta)} & \lambda U^{*}(x, y)  \tag{2.4.11}\\
\lambda U(x, y) & -\Omega I+D \cdot \sigma^{(\eta)}
\end{array}\right)
$$

Here, $\lambda$ and $\Omega$ are fixed real constants, $D \cdot \sigma^{(\eta)}:=-i \partial_{x} \sigma_{1}-i \eta \partial_{y} \sigma_{2}$ with $\eta \in\{-1,1\}$, and $U(x, y):=\frac{1}{2}\left((1+\tilde{m}(x, y)) A+(1-\tilde{m}(x, y)) A^{*}\right)$. The function $\tilde{m} \in \mathcal{C}^{\infty}$ is positive in regions of AB stacking, negative in regions of BA stacking, and zero on the boundaries between regions. We assume that $\tilde{m} \in\{-1,1\}$ away from these boundaries.

To construct such a function $\tilde{m}$ explicitly, define $f_{\Theta}: \mathbb{T} \rightarrow \mathbb{T}$ by $f_{\Theta}(\theta)=\sin (3 \theta)$ and (as above) let $f \in \mathcal{C}^{\infty}\left(\mathbb{R}^{2}\right)$ such that $f(x, y)=r f_{\Theta}(\theta)$ for all $r \geq 1$, where $(r, \theta)$ are the polar coordinates corresponding to $(x, y)$. For concreteness, take $f(x, y):=\chi(r) r f_{\Theta}(\theta)$, where $\chi \in \mathfrak{S}(0,1 ; \varepsilon, 1)$ and $0<\varepsilon<1$. Since $\chi$ is smooth and vanishes near the origin, it is immediate that $f$ is smooth.

Let $m \in \mathfrak{S}(-1,1)$ be monotonically increasing with $m(0)=0$, and define $\tilde{m}(x, y):=$ $m(f(x, y))$. This definition is consistent with the above constraints, as $f_{\Theta}>0$ if and only if $2 k \pi / 3<\theta<(2 k+1) \pi / 3$ for some $k \in\{0,1,2\}$. Since $m$ and all of its derivatives are bounded, it follows that $\sigma \in S_{1,0}^{1}$ with the right-hand side defined in Appendix B. Moreover, the structure of $H$ (with $D \cdot \sigma^{(\eta)}$ on the diagonal and no other derivatives) implies the existence of a constant $c>0$ such that $\left|\sigma_{\min }(x, y, \xi, \zeta)\right| \geq c\langle\xi, \zeta\rangle-1$, where $\sigma_{\min }$ is the smallest magnitude eigenvalue of $\sigma$. We see that when $f(x, y)$ is sufficiently large (resp.
small), $\sigma=\sigma_{+}\left(\right.$resp. $\left.\sigma=\sigma_{-}\right)$, where

$$
\sigma_{+}=\left(\begin{array}{cc}
\Omega I+\xi \sigma_{1}+\eta \zeta \sigma_{2} & \lambda A^{*} \\
\lambda A & -\Omega I+\xi \sigma_{1}+\eta \zeta \sigma_{2}
\end{array}\right)
$$

and

$$
\sigma_{-}=\left(\begin{array}{cc}
\Omega I+\xi \sigma_{1}+\eta \zeta \sigma_{2} & \lambda A \\
\lambda A^{*} & -\Omega I+\xi \sigma_{1}+\eta \zeta \sigma_{2}
\end{array}\right) .
$$

By [13, Appendix A], there exists $E>0$ such that $\sigma_{ \pm}\left(\right.$equivalently $\left.H_{ \pm}:=\operatorname{Op}\left(\sigma_{ \pm}\right)\right)$has a spectral gap in the interval $\left(E_{1}, E_{2}\right):=(-E, E)$.

Thus we have shown that $H$ given by (2.4.11) satisfies (H1'), meaning that Theorem 2.4.8 applies. For concreteness, define $g_{\Theta}: \mathbb{T} \rightarrow \mathbb{T}$ by $g_{\Theta}(\theta)=\cos (\theta)+\cos (\pi / 6)=\cos (\theta)+\sqrt{3} / 2$ (note that $g_{\Theta}<0$ if and only if $5 \pi / 6<\theta<7 \pi / 6$ ) and take $g(x, y)=\chi(r) r g_{\Theta}(\theta)$. Due to the factor of $\chi$ in the definition of $g$, the level set $g^{-1}(0)$ is not a curve in the $x y$-plane. Still, there exists a smooth function $\left(x_{0}, y_{0}\right): \mathbb{R} \rightarrow \mathbb{R}^{2}$ whose range separates the $x y$-plane into two regions, $R_{+}$and $R_{-}$, such that $g \geq 0$ in $R_{+}$and $g \leq 0$ in $R_{-}$. More specifically, we can take

$$
\left(x_{0}(t), y_{0}(t)\right)= \begin{cases}\left(\frac{\sqrt{3}}{2} t, \frac{1}{2} t\right), & t \leq-1 \\ \left(-\frac{\sqrt{3}}{2} t, \frac{1}{2} t\right), & t \geq 1\end{cases}
$$

and smoothly connect $\left(x_{0}, y_{0}\right)$ for $|t|<1$ so that $g\left(x_{0}(t), y_{0}(t)\right)=0$ for all $t \in \mathbb{R}$. It follows that $\tau(t, \xi, \zeta)=\sigma_{ \pm}(\xi, \zeta)$ whenever $\pm t>0$ is sufficiently large, where $\tau$ is defined in terms of $\left(x_{0}, y_{0}\right)$ by (2.4.5). By [13, Theorem 2.1], this means

$$
2 \pi \sigma_{I}(H, P, \varphi)=-2 \eta \operatorname{sgn}(\Omega) .
$$

If $P_{j}=\chi_{p}\left(g_{j}(x, y)\right)$ for $j \in\{0,1,2\}$ with the $g_{j}$ appropriately chosen to satisfy the growth condition for $\langle f, g\rangle$, then $\sigma_{I}\left(H, \sum_{j} P_{j}, \varphi\right)=\sum_{j} \sigma_{I}\left(H, P_{j}, \varphi\right)$. Thus if $g_{j}(x, y)=\chi(r) r g_{j_{\Theta}}(\theta)$ with $g_{j_{\Theta}}=\cos (\theta-2 \pi j / 3)+\cos (\pi / 6)$, then the $2 \pi / 3$-rotational symmetry of the tBLG Hamiltonian implies that $\sigma_{I}\left(H, \sum_{j} P_{j}, \varphi\right)=3 \sigma_{I}\left(H, P_{0}, \varphi\right)$. That is, each term $\sigma_{I}\left(H, P_{j}, \varphi\right)$ in the sum over $j$ sees the same transition from $\sigma_{-}$to $\sigma_{+}$. In words, the superposition of three appropriately chosen regularized indicator functions increases the conductivity by a factor of 3. This makes perfect sense, as the conductivity is now measured through three distinct regions $\left(\operatorname{supp} \nabla P_{j}\right.$ for $\left.j=0,1,2\right)$, each of which contributes the same value of $-\eta \operatorname{sgn}(\Omega) / \pi$.
$2 \times 2$ Dirac Hamiltonian. We conclude this section with an application of Theorem 2.4.8 to the operator

$$
\begin{equation*}
H=D_{x} \sigma_{1}+D_{y} \sigma_{2}+\tilde{m}(x, y) \sigma_{3}, \tag{2.4.12}
\end{equation*}
$$

where $\tilde{m}(x, y)=m(f(x, y))$ for some $m \in \mathfrak{S}(-1,1)$, and

$$
f(r, \theta)=\chi(r) r f_{\Theta}(\theta), \quad \chi \in \mathfrak{S}(0,1 ; \varepsilon, 1), \quad f_{\Theta}(\theta)=\sin (k \theta)
$$

for some $0<\varepsilon<1$ and $k \in \mathbb{N}_{+}$. Note that when $k=1$, we recover the setting of a flat interface analyzed in detail in Sections 2.1-2.3, while $k=3$ yields the same hexagonal structure on the right panel of Figure 2.1. As with the Bistritzer-MacDonald Hamiltonian (2.4.11), it is straightforward to verify that $H$ defined by (2.4.12) satisfies (H1'), and thus the above theory applies. The bulk spectral gap in this case is $\left(E_{1}, E_{2}\right)=(-1,1)$.

Take $P(x, y)=\chi_{p}(g(x, y))$ for $\chi_{p} \in \mathfrak{S}(0,1)$, where $g(x, y)=g(r, \theta)=\chi(r) r g_{\Theta}(\theta)$ and $g_{\Theta}(\theta)=\cos \left(\theta-\theta_{1}\right)-\cos \left(\theta_{0}\right)$, for some $0<\theta_{0}<\pi$ and $-\frac{\pi}{k}<\theta_{1}<\frac{\pi}{k}$ satisfying

$$
\theta_{+} \notin \Theta_{k}, \quad 2 \pi-\theta_{+} \notin \Theta_{k}, \quad \Theta_{k}=\left\{\frac{j}{k \pi}: j \in \mathbb{Z}\right\}, \quad \theta_{+}:=\theta_{0}+\theta_{1}
$$

This way, the zeros of $g_{\Theta}$ and $f_{\Theta}$ are disjoint. Indeed, $g_{\Theta}(\theta)=0$ if and only if $\cos \left(\theta-\theta_{1}\right)=$
$\cos \left(\theta_{0}\right)$, which occurs exactly when $\{\theta+2 \pi j: j \in \mathbb{Z}\} \cap\left\{\theta_{+}, 2 \pi-\theta_{+}\right\} \neq \emptyset$. It follows that $\tau(t, \xi, \zeta)=\xi \sigma_{1}+\zeta \sigma_{2}+\mu(t) \sigma_{3}$ for some $\mu \in \cup_{\varepsilon_{1}, \varepsilon_{2} \in\{-1,1\}} \mathfrak{S}\left(\varepsilon_{1}, \varepsilon_{2}\right)$, where the $\varepsilon_{j}$ are determined by $\theta_{0}$ and $\theta_{1}$. We have reduced the problem to computing the interface conductivity for the translation-invariant $2 \times 2$ Dirac system, and hence $2 \pi \sigma_{I}=\frac{1}{2}\left(\varepsilon_{1}-\varepsilon_{2}\right)$ by Section 2.3. For the case $\left(k, \theta_{1}\right)=(3,0)$ which is analyzed numerically in the following section, we have

$$
2 \pi \sigma_{I}= \begin{cases}-1, & 0<\theta_{0}<\pi / 3 \\ 1, & \pi / 3<\theta_{0}<2 \pi / 3 \\ -1, & 2 \pi / 3<\theta_{0}<\pi\end{cases}
$$

If $g$ from Figure 2.1 were to have the above form, then $\theta_{1}=0$ and $\theta_{0}$ would be the angle between [the line making up the top of the level curve $\{g=0\}$ ] and [the positive $x$-axis]. So $\pi / 2<\theta_{0}<\pi$ and $2 \pi / 3<\theta_{0}<\pi$ for the left and right panels, respectively.

## CHAPTER 3

## MAGNETIC DIRAC EQUATIONS

In this section, we derive a quantized conductivity at the interface of two (distinct) sheets of graphene (or other similar materials). As we will see below, the Hamiltonians considered here have symbols that are unbounded in the spatial coordinates $(x, y)$ and not elliptic. As a result, none of the theory from Section 2 applies directly.

In the absence of external electric and magnetic fields, the edge state dynamics is governed by a two-dimensional Schrödinger Hamiltonian whose coefficients obey the appropriate honeycomb symmetry away from the interface. It was shown in $[8,36,44]$ that at low energies, this Hamiltonian is well approximated by the following Dirac operator,

$$
H_{D}=D_{x} \sigma_{1}+D_{y} \sigma_{2}+m(x) \sigma_{3} ; \quad D_{\alpha}:=-i \partial_{\alpha}, \quad m \in \mathfrak{S}\left(m_{-}, m_{+}\right), \quad 0 \neq m_{ \pm} \in \mathbb{R}
$$

Here, $\mathfrak{S}(a, b):=\cup_{\alpha<\beta} \mathfrak{S}(a, b ; \alpha, \beta)$ with $\mathfrak{S}(a, b ; \alpha, \beta)$ the set of smooth functions $f: \mathbb{R} \rightarrow \mathbb{R}$ such that $f(\lambda)=a($ resp. $f(\lambda)=b)$ whenever $\lambda \leq \alpha$ (resp. $\lambda \geq \beta$ ), and the Pauli matrices are given by

$$
\sigma_{1}=\left(\begin{array}{cc}
0 & 1 \\
1 & 0
\end{array}\right), \quad \sigma_{2}=\left(\begin{array}{cc}
0 & -i \\
i & 0
\end{array}\right), \quad \sigma_{3}=\left(\begin{array}{cc}
1 & 0 \\
0 & -1
\end{array}\right)
$$

Observe that the interface material described by $H_{D}$ is made up of two insulators (given by $\left.H_{ \pm}:=D_{x} \sigma_{1}+D_{y} \sigma_{2}+m_{ \pm} \sigma_{3}\right)$ that are glued together along a one-dimensional interface, say $\{m(x)=0\}$. The mass term $m$ models the transition from one insulator to the other. Although $H_{ \pm}$has a spectral gap in the interval $\left(-\left|m_{ \pm}\right|,\left|m_{ \pm}\right|\right)$, the interface material is a conductor whenever $m_{+}$and $m_{-}$have opposite sign. In this case $H_{D}$ no longer has a spectral gap, and the energies in the interval $\left(-m_{0}, m_{0}\right)$ with $m_{0}:=\min \left\{\left|m_{-}\right|,\left|m_{+}\right|\right\}$correspond to propagating edge modes as described above. As before, the asymmetric transport associated
with this model is quantified by the following interface conductivity,

$$
\begin{equation*}
\sigma_{I}(H):=\operatorname{Tr} i[H, P] \varphi^{\prime}(H) \tag{3.0.1}
\end{equation*}
$$

Here, $P(y)=P \in \mathfrak{S}(0,1)$ and $\varphi \in \mathfrak{S}\left(0,1 ;-m_{0}, m_{0}\right)$. Explicit formulas for $\sigma_{I}$ (that apply to $H_{D}$ among other models) are derived in [10, 11]. In [7] it was shown that $2 \pi \sigma_{I}\left(H_{D}\right)=$ $\operatorname{SF}\left(H_{D} ; \alpha\right)=\frac{1}{2} \operatorname{sgn}\left(m_{+}-m_{-}\right)$, where $\operatorname{SF}\left(H_{D} ; \alpha\right)$ is the spectral flow of $H_{D}$ through $\alpha$ (more on this below) and $\alpha \in\left(-m_{0}, m_{0}\right)$. Observe that the quantity $\operatorname{sgn}\left(m_{-} m_{+}\right)$is independent of $\alpha, P$ and $\varphi$, and is robust with respect to changes in $m$.

The goal of this paper is to extend the above results for $H_{D}$ to Dirac operators with electric and magnetic fields; we will explicitly calculate the interface conductivity by relating it to a spectral flow, and prove its stability in the presence of perturbations. The electromagnetic Dirac operators are given by

$$
\begin{equation*}
H=D_{x} \sigma_{1}+\left(D_{y}-A_{2}(x)\right) \sigma_{2}+m(x) \sigma_{3}+V(x) \sigma_{0} \tag{3.0.2}
\end{equation*}
$$

where $A_{2}(x)=x B(x)$ and

$$
\begin{equation*}
B \in \mathfrak{S}\left(B_{-}, B_{+}\right), \quad m \in \mathfrak{S}\left(m_{-}, m_{+}\right), \quad V \in \mathfrak{S}\left(V_{-}, V_{+}\right) \tag{3.0.3}
\end{equation*}
$$

for some constants $B_{ \pm}, m_{ \pm}, V_{ \pm} \in \mathbb{R}$ with $B_{ \pm} \neq 0$. Here, $\sigma_{0}$ is the $2 \times 2$ identity matrix. The vector-valued function $A=\left(0, A_{2}\right)$ is the magnetic potential (in the Landau gauge), with $\left(B+x B^{\prime}\right) \hat{z}=\nabla \times A$ the magnetic field. The function $V$ represents the electric potential. See [82, Section 4.2] for a derivation of these models.

We see that the Hamiltonian (3.0.2) implements three domain walls given by (3.0.3). The magnetic domain wall $B$ gives rise to the unbounded term $A_{2}(x)$, while the functions $m$ and $V$ are necessarily bounded. We have already discussed the domain wall in $m$ as a mechanism
for transport. To motivate the domain wall in $B$, we provide an illustrative example.
Consider a two-dimensional electron gas confined to the right-half plane and subject to a constant (and strong) orthogonal magnetic field. Away from the edge $\{x=0\}$ each electron will move in a (small) circular path, meaning the gas is insulating in its bulk. However the presence of the edge causes nearby electrons to propagate, giving rise to a current along the boundary. The same can be said of a (no longer confined) two-dimensional electron gas subject to an orthogonal magnetic field that changes signs across the $y$-axis.

The domain wall in $V$ gives rise to an electric field $E=-\nabla V$ perpendicular to the interface that vanishes whenever $|x|$ is sufficiently large. Note that if the mass term were to grow linearly at infinity, the interface conductivity would no longer depend on the magnetic field [11].

The spectral properties of $H$ are acquired from the following bulk Hamiltonians,

$$
H_{ \pm}=D_{x} \sigma_{1}+\left(D_{y}-x B_{ \pm}\right) \sigma_{2}+m_{ \pm} \sigma_{3}+V_{ \pm} \sigma_{0}
$$

which model the two insulating materials that are glued together to form the conductor with Hamiltonian $H$. Since $H$ and $H_{ \pm}$are translation-invariant in $y$, we can define their Fourier transforms by

$$
\begin{aligned}
\hat{H}(\zeta) & =D_{x} \sigma_{1}+\left(\zeta-A_{2}(x)\right) \sigma_{2}+m(x) \sigma_{3}+V(x) \sigma_{0} \\
\hat{H}_{ \pm}(\zeta) & =D_{x} \sigma_{1}+\left(\zeta-x B_{ \pm}\right) \sigma_{2}+m_{ \pm} \sigma_{3}+V_{ \pm} \sigma_{0}
\end{aligned}
$$

for $\zeta \in \mathbb{R}$. It is known that each operator $\mathcal{O} \in\left\{H, H_{+}, H_{-}, \hat{H}(\zeta), \hat{H}_{+}(\zeta), \hat{H}_{-}(\zeta)\right\}$ is selfadjoint with domain of definition $\mathcal{D}(\mathcal{O})=(i-\mathcal{O})^{-1} \mathcal{H}$; see e.g. [11, 20, 82]. Here, $\mathcal{H}:=$ $L^{2}\left(\mathbb{R}^{d}\right) \otimes \mathbb{C}^{2}$, where $d=2$ for $H$ and $H_{ \pm}$, and $d=1$ for $\hat{H}$ and $\hat{H}_{ \pm}$. Note that for a similar edge model, it was shown that the domain of definition of the Hamiltonian depends on the strength of the magnetic field [29]. This suggests that $\mathcal{D}(\mathcal{O})$ likely depends on $B_{ \pm}$, though
we do not investigate this issue here.
Throughout this thesis, we denote the spectrum of an operator $\mathcal{O}$ by $\sigma(\mathcal{O})$ and the resolvent set of $\mathcal{O}$ by $\rho(\mathcal{O})$. Suppose temporarily that $V_{ \pm}=0$, as these constants contribute merely a uniform shift to the spectrum of $H_{ \pm}$. Then $\hat{H}_{ \pm}^{2}(\zeta)=D_{x}^{2}+\left(\zeta-x B_{ \pm}\right)^{2}+m_{ \pm}^{2}-B_{ \pm} \sigma_{3}$ is block diagonal, meaning that up to shifts by $m_{ \pm}^{2}-B_{ \pm}$and $m_{ \pm}^{2}+B_{ \pm}$, the spectra of $\hat{H}_{ \pm}^{2}(\zeta)$ and $L_{ \pm}(\zeta):=D_{x}^{2}+\left(\zeta-x B_{ \pm}\right)^{2}$ are the same. But $L_{ \pm}(\zeta)$ is the Hamiltonian for the quantum harmonic oscillator (up to rescaling) and has spectrum consisting entirely of simple eigenvalues and given by $\sigma\left(L_{ \pm}(\zeta)\right)=\{(2 k+1) B: k \in \mathbb{N}\}$. The elements of $\sigma\left(L_{ \pm}(\zeta)\right)$ are known as "Landau levels." Since $\sigma\left(L_{ \pm}(\zeta)\right)$ (and hence $\sigma\left(\hat{H}_{ \pm}(\zeta)\right)$ ) is independent of $\zeta$, it follows that $\sigma\left(H_{ \pm}\right)$is made up of a countable collection of points (all corresponding to essential spectrum now) going to infinity in absolute value; see Lemma 3.1.2. The domain wall in $B$ can cause spectral gaps between Landau levels to close, with eigenvalues of $\hat{H}^{2}(\zeta)$ potentially going to infinity as $\zeta \rightarrow \pm \infty$; see Lemma 3.1.7.

When spectral branches of $\hat{H}(\zeta)$ do not go to infinity, they converge to elements of $\sigma\left(H_{ \pm}\right)$ as shown by Lemma 3.1.6. Each branch converges both as $\zeta \rightarrow \infty$ and $\zeta \rightarrow-\infty$ in the case that $B_{+}$and $B_{-}$have the same sign. When $B_{-}<0<B_{+}$(resp. $B_{+}<0<B_{-}$), the branches converge only as $\zeta \rightarrow \infty$ (resp. $\zeta \rightarrow-\infty$ ). Hence no matter the signs of $B_{+}$and $B_{-}$, the branches of spectrum do not go to infinity as $|\zeta| \rightarrow \infty$. This means $H$ lacks the ellipticity that is assumed by (H1) and in $[10,11,37]$. Indeed, the term $\left(D_{y}-A_{2}(x)\right) \sigma_{2}$ has (Weyl) symbol $\left(\zeta-A_{2}(x)\right) \sigma_{2}$, which can remain small even when $|\zeta|$ and $|x|$ are large.

Below is a brief summary of this section. We first define the spectral flow, and in doing so give an outline of Section 3.1. We refer to [66] for a more generally applicable definition. The set $\{\hat{H}(\zeta): \zeta \in \mathbb{R}\}$ is a one-parameter family of self-adjoint operators, with $\hat{H}(\zeta)$ holomorphic in $\zeta \in \mathbb{C}$ (see [58, Chapter VII.1.1] for a precise definition of holomorphic operators). We will show with Lemmas 3.1.3 and 3.1.4 that the spectrum of $\hat{H}(\zeta)$ consists
entirely of simple eigenvalues $\left\{\mu_{j}(\zeta)\right\}_{j \in \mathbb{Z}}$, where each branch $\mu_{j}: \mathbb{R} \rightarrow \mathbb{R}$ is analytic. Let

$$
\begin{equation*}
\mathcal{R}:=\rho\left(H_{+}\right) \cap \rho\left(H_{-}\right) \cap \mathbb{R} \tag{3.0.4}
\end{equation*}
$$

denote the bulk band gaps. It turns out that for any energy level $\alpha \in \mathcal{R}$, there exists $\zeta_{\alpha}>0$ such that no branches attain the value $\alpha$ when $|\zeta|>\zeta_{\alpha}$ (Lemmas 3.1.6 and 3.1.7). Moreover, the number of branches to ever attain the value $\alpha$ is finite (Lemma 3.1.8).

This means we can define the spectral flow of $H$ through $\alpha$ to be the signed number of crossings of branches through $\alpha$ :

$$
\begin{equation*}
\operatorname{SF}(H ; \alpha):=N_{\uparrow}-N_{\downarrow}, \tag{3.0.5}
\end{equation*}
$$

where $N_{\uparrow}$ (resp. $N_{\downarrow}$ ) is the number of branches that are less than $\alpha$ when $\zeta<-\zeta_{\alpha}$ and greater than $\alpha$ when $\zeta>\zeta_{\alpha}$ (resp. greater than $\alpha$ when $\zeta<-\zeta_{\alpha}$ and less than $\alpha$ when $\zeta>\zeta_{\alpha}$ ). With Theorem 3.1.1 we compute the spectral flow using the max-min principle [71, 81] and perturbation theory [58].

In Section 3.2, we relate the interface conductivity in (3.0.1) to the spectral flow in (3.0.5) (Theorem 3.2.5) and prove its stability under a large class of perturbations (Theorems 3.2.93.2.12). The stability results are proved using pseudo-differential calculus as in Section 2 and $[10,11,37]$. But since $H$ is not elliptic, much of the pseudo-differential operator theory in this existing literature does not apply directly. Still, we are able to use Beals's criterion [34, Proposition 8.3] and specific properties of $H$ (as a first-order differential operator with the spectrum of $H_{ \pm}$known) to obtain the necessary decay properties; see Lemmas 3.2.2 and 3.2.3. Similar stability results for (also non-elliptic) magnetic Schrödinger operators can be found in $[28,35]$.

### 3.1 Spectral analysis

In this section we calculate the spectral flow $\operatorname{SF}(H ; \alpha)$ for $\alpha \in \mathcal{R}$, i.e. an energy level in a spectral gap for both bulk Hamiltonians. This involves analyzing the limiting behavior of the branches of spectrum $\mu_{j}(\zeta)$ of $\hat{H}(\zeta)$ as $|\zeta| \rightarrow \infty$. The multisets $S_{ \pm}:=\left\{\lim _{\zeta \rightarrow \pm \infty} \mu_{j}(\zeta)\right\}_{j \in \mathbb{Z}}$ are determined using a standard max-min argument. Perturbation theory is then used to match elements in $S_{+}$with those in $S_{-}$to determine the quantities $\lim _{\zeta \rightarrow \pm \infty} \mu_{j}(\zeta)$ for every $j$. We now state the main result, which can be interpreted as a bulk-interface correspondence. We recall that $B_{ \pm} \neq 0$ throughout this paper.

Theorem 3.1.1. Fix $\alpha \in \mathcal{R}$. Then $S F(H ; \alpha)=I\left(H_{-} ; \alpha\right)-I\left(H_{+} ; \alpha\right)$, where

$$
\begin{equation*}
I\left(H_{ \pm} ; \alpha\right)=\operatorname{sgn}\left(B_{ \pm}\right) \operatorname{sgn}\left(\alpha-V_{ \pm}-m_{ \pm} \operatorname{sgn}\left(B_{ \pm}\right)\right)\left(N\left(H_{ \pm} ; \alpha\right)+\frac{1}{2}\right) \tag{3.1.1}
\end{equation*}
$$

and

$$
N\left(H_{ \pm} ; \alpha\right)= \begin{cases}0 ; & \left|\alpha-V_{ \pm}\right|<\sqrt{2\left|B_{ \pm}\right|+m_{ \pm}^{2}}  \tag{3.1.2}\\ k ; & \sqrt{2 k\left|B_{ \pm}\right|+m_{ \pm}^{2}}<\left|\alpha-V_{ \pm}\right|<\sqrt{2(k+1)\left|B_{ \pm}\right|+m_{ \pm}^{2}}, \quad k \in \mathbb{N}_{+}\end{cases}
$$

Note that $N\left(H_{ \pm} ; \alpha\right)$ counts the ( $\zeta$-independent) number of eigenvalues of $\hat{H}_{ \pm}(\zeta)-V_{ \pm}$ in the interval $\left(\left|m_{ \pm}\right|,\left|\alpha-V_{ \pm}\right|\right)$. The above values of $\alpha$ for which $\operatorname{SF}(\alpha)$ is not defined are precisely the elements of $\sigma\left(H_{+}\right) \cup \sigma\left(H_{-}\right)$. Although the above expresses the (integer-valued) spectral flow as a difference of two bulk quantities, it is unclear whether each bulk quantity may be interpreted as an invariant as in $[6,16]$.

We refer to Figures 3.1-3.3 for an illustration of the branches of spectrum of $H$, under various choices of the parameters $B_{ \pm}, m_{ \pm}$and $V_{ \pm} .{ }^{1}$ For example, in the top right panel of

[^0]

Figure 3.1: Plots of the lowest-magnitude eigenvalues of $\hat{H}(\zeta)$ as a function of $\zeta$, for different combinations of domain walls in $B$ and $m$. For all plots, $V \equiv 0$. The choices of parameters are $B \equiv 2$ and $m \equiv 2$ (top left); $B \equiv 2$ and $m_{+}=2=-m_{-}$(top right); $B_{+}=2=-B_{-}$ and $m \equiv 0$ (bottom left); $B_{+}=2=-B_{-}$and $m \equiv 2$ (bottom right). As $\zeta \rightarrow+\infty$, these eigenvalues converge to the elements of $\sigma\left(\hat{H}_{+}(\zeta)\right) \cup \sigma\left(\hat{H}_{+}(\zeta)\right)$ as predicted by the theory; see Lemma 3.1.7. The top left panel illustrates the spectrum of a bulk Hamiltonian; see Lemma 3.1.2. As demonstrated by the top right plot, a transition in the sign of $m$ generates a nonzero spectral flow for $\alpha$ near 0 (in this example, $\operatorname{SF}(H ; \alpha)=1$ ). Comparing the two bottom plots, we see how a constant nonzero $m$ opens a gap at 0 . Although it may look like the eigenvalues in the bottom plots are degenerate for large $\zeta$, this is not the case (Lemma 3.1.3). They only have the same limit as $\zeta \rightarrow \infty$.

Figure 3.2 we see that $\operatorname{SF}(H ; \alpha)=1$ for all $\alpha$ near 0 , while $\operatorname{SF}(H ; \alpha)=-1$ for any $\alpha$ near 2.5.
periodic boundary conditions. The periodic Hamiltonian also has spurious eigenvalues (corresponding to eigenfunctions that are localized near the boundary; see Section 4 for more details), which are not included in our plots.


Figure 3.2: Effect of the electric potential $V$ on the lowest-magnitude eigenvalues of $\hat{H}(\zeta)$. All plots have $B_{+}=2=-B_{-}$and $m_{+}=2=-m_{-}$. The choices of electric potential are $V \equiv 0$ (top left); $V_{+}=0.1=-V_{-}($top right $) ; V_{+}=0.5=-V_{-}$(bottom left); $V_{+}=2=-V_{-}$ (bottom right). Notice that, as predicted by Lemma 3.1.2, a small domain wall in $V$ opens a gap between pairs of branches for large $\zeta$ (top panels).

The rest of this section is devoted to proving Theorem 3.1.1 (and the statements we made before it). We begin by determining the spectrum of $\hat{H}_{ \pm}(\zeta)$.

Lemma 3.1.2. For any $\zeta \in \mathbb{R}$, the spectrum of $\hat{H}_{ \pm}(\zeta)$ consists entirely of eigenvalues and


Figure 3.3: Effect of the signs of $B_{ \pm}$and $m_{ \pm}$on the lowest-magnitude eigenvalues of $\hat{H}(\zeta)$. Both plots have $-B_{+}=2=B_{-}$and $V \equiv 0$. The left panel corresponds to $m_{+}=1=-m_{-}$ and the right to $-m_{+}=1=m_{-}$. Since $B_{+}<0<B_{-}$(in contrast with Figures 3.1 and 3.2), the branches now converge to the elements of $\sigma\left(\hat{H}_{+}(\zeta)\right) \cup \sigma\left(\hat{H}_{+}(\zeta)\right)$ as $\zeta \rightarrow-\infty$ while going to infinity as $\zeta \rightarrow+\infty$. By comparing the left and right panels, observe that the sign-flips in $m_{ \pm}$change the spectral flow only for $-1<\alpha<1$.
is given by

$$
\begin{equation*}
\sigma\left(\hat{H}_{ \pm}(\zeta)\right)=\left\{\varepsilon \sqrt{2 k\left|B_{ \pm}\right|+m_{ \pm}^{2}}+V_{ \pm}: \varepsilon \in\{-1,1\}, k \in \mathbb{N}_{+}\right\} \bigcup\left\{m_{ \pm} \operatorname{sgn}\left(B_{ \pm}\right)+V_{ \pm}\right\} \tag{3.1.3}
\end{equation*}
$$

Above, the subscripts $\pm$ (and one operation $\mp$ ) are understood to correspond to the operator $\hat{H}_{ \pm}$in question. Note that the spectrum of $\hat{H}_{ \pm}(\zeta)$ is independent of $\zeta$; thus the spectrum of $H_{ \pm}$is also given by (3.1.3), only now these values all belong to the essential spectrum.

Proof. Suppose for concreteness that $B_{+}>0$. Set $V_{+}=0$ without loss of generality, as this term only contributes a uniform shift of the spectrum. Any eigenpair $\mu$ and $\psi$ of $\hat{H}_{+}(\zeta)$ must satisfy $\hat{H}_{+}^{2}(\zeta) \psi=\mu^{2} \psi$, with $\hat{H}_{+}^{2}(\zeta)=D_{x}^{2}+\left(\zeta-B_{+} x\right)^{2}+m_{+}^{2}-B_{+} \sigma_{3}$. The eigenelements of $\hat{H}_{+}^{2}(\zeta)$ are well known. The eigenvalues are $\nu_{k}=2 k\left|B_{+}\right|+m_{+}^{2}$ for $k \in \mathbb{N}$. When $k \in \mathbb{N}_{+}, \nu_{k}$ has multiplicity 2 and the eigenfunctions are $\psi_{k, \uparrow}=\left(\phi_{k}, 0\right)$ and $\psi_{k, \downarrow}=\left(0, \phi_{k-1}\right)$, where $\phi_{k}(x)=\left|B_{+}\right|^{-1 / 4} \tilde{\phi}_{k}\left(\sqrt{\left|B_{+}\right|} x-\zeta\right)$ with $\tilde{\phi}_{k}$ the Hermite functions.

We see that $\nu_{0}$ has multiplicity one with eigenfunction $\psi_{0}=\left(\phi_{0}, 0\right)$. We then verify that the eigenvalues of $\hat{H}_{+}(\zeta)$ are $\mu_{k, \varepsilon}=\varepsilon \sqrt{\nu_{k}}$ for $\varepsilon \in\{-1,1\}, k \in \mathbb{N}_{+}$and $\mu_{0}=m_{+}$, with corresponding eigenfunctions $\psi_{k, \varepsilon}=c_{1, \varepsilon} \psi_{k, \uparrow}+c_{2, \varepsilon} \psi_{k, \downarrow}$ and $\psi_{0}$ for some $\left|c_{1, \varepsilon}\right|^{2}+\left|c_{2, \varepsilon}\right|^{2}=1$. If instead $B_{+}<0$, we would instead have $\psi_{k, \uparrow}=\left(\phi_{k-1}, 0\right)$ and $\psi_{k, \downarrow}=\left(0, \phi_{k}\right)$. All elements of the spectrum would be the same as before with the exception of $\mu_{0}=-m_{+}$now. The eigenelements of $\hat{H}_{-}(\zeta)$ are calculated similarly. This completes the result.

The following two lemmas state important properties of the spectrum of $\hat{H}(\zeta)$, with $\zeta \in \mathbb{C}$ now. The extension to complex $\zeta$ will be necessary when we later apply the holomorphic perturbation theory from [58].

Lemma 3.1.3. For any $\zeta \in \mathbb{C}$, the spectrum of $\hat{H}(\zeta)$ consists entirely of simple eigenvalues.
Proof. Fix $\zeta_{0}+i \zeta_{1}:=\zeta \in \mathbb{C}$. We know that $\hat{H}\left(\zeta_{0}\right)$ is self-adjoint [11, 20, 82], hence $|\Im \mu| \leq\left|\zeta_{1}\right|$ for all $\mu \in \sigma(\hat{H}(\zeta))$. The Weyl symbol (see Appendix A. 1 for the definition) of $\hat{H}(\zeta)$ grows linearly in $\langle x, \xi\rangle$, as demonstrated by (3.2.2) below. This implies that $\left(i\left(1+\left|\zeta_{1}\right|\right)+\hat{H}(\zeta)\right)^{-1}$ is compact (see e.g. [11, 20]), and hence $\sigma(\hat{H}(\zeta))$ consists entirely of eigenvalues. We now prove that each eigenvalue has multiplicity one. Suppose $\psi$ and $\phi$ are eigenfunctions of $\hat{H}$ corresponding to eigenvalue $\mu$. This means

$$
\begin{aligned}
& -i \psi_{2}^{\prime}-i\left(\zeta-A_{2}\right) \psi_{2}+(V+m) \psi_{1}=\mu \psi_{1} \\
& -i \psi_{1}^{\prime}+i\left(\zeta-A_{2}\right) \psi_{1}+(V-m) \psi_{2}=\mu \psi_{2}
\end{aligned}
$$

with the same equations holding also with $\psi$ replaced by $\phi$. It follows that

$$
\psi_{1} \phi_{2}^{\prime}+\psi_{1}^{\prime} \phi_{2}=i(\mu-m-V) \psi_{1} \phi_{1}+i(\mu+m-V) \psi_{2} \phi_{2}=\psi_{2} \phi_{1}^{\prime}+\psi_{2}^{\prime} \phi_{1}
$$

so that

$$
\partial_{x}\left(\psi_{1} \phi_{2}-\psi_{2} \phi_{1}\right)=\psi_{1} \phi_{2}^{\prime}+\psi_{1}^{\prime} \phi_{2}-\left(\psi_{2} \phi_{1}^{\prime}+\psi_{2}^{\prime} \phi_{1}\right)=0 .
$$

Since $\psi, \phi, \psi^{\prime}, \phi^{\prime}$ all go to zero as $|x| \rightarrow \infty$, we have shown that $\psi_{1} \phi_{2}-\psi_{2} \phi_{1} \equiv 0$. Thus for every $x \in \mathbb{R}$, the vectors $\left(\psi_{1}(x), \psi_{2}(x)\right)$ and $\left(\phi_{1}(x), \phi_{2}(x)\right)$ are linearly dependent. Normalize the eigenfunctions so that $\psi\left(x_{0}\right)=\phi\left(x_{0}\right)$ for some $x_{0} \in \mathbb{R}$ (this can be done because $\psi$ and $\phi$ are continuous, hence there must exist a point at which both functions are zero or both functions are nonzero). Letting $\nu:=\psi-\phi$, we see that

$$
\nu^{\prime}(x)=F(x) \nu(x), \quad x \in \mathbb{R} ; \quad \nu\left(x_{0}\right)=0
$$

for some $F \in \mathcal{C}^{\infty}\left(\mathbb{R} ; \mathbb{C}^{2 \times 2}\right)$. By regularity of $F$, the standard uniqueness result for first-order ODE implies that $\nu \equiv 0$. This completes the result.

Lemma 3.1.4. There exists a countable collection of holomorphic functions, $\left\{\mu_{j}\right\}_{j \in \mathbb{Z}} \subset$ $\mathcal{C}^{\omega}(\mathbb{C})$, such that for each $\zeta \in \mathbb{C}, \sigma(\hat{H}(\zeta))=\left\{\mu_{j}(\zeta): j \in \mathbb{Z}\right\}$. The corresponding eigenprojections are also holomorphic in $\zeta \in \mathbb{C}$. Finally, the restriction of each $\mu_{j}$ to the real axis defines a real-analytic function.

Here, $\mathcal{C}^{\omega}(\mathbb{C})$ denotes the set of holomorphic functions on $\mathbb{C}$. We refer to [58, Chapter VII.1.1] for the precise definition of an operator-valued holomorphic function.

Proof. Note that $\hat{H}(\zeta)$ is holomorphic in $\zeta \in \mathbb{C}$, thus Lemma 3.1.3 and [58, Theorems VII.1.7 and VII.1.8] imply that the eigenvalues and eigenprojections of $\hat{H}(\zeta)$ are also holomorphic in $\zeta \in \mathbb{C}$. That $\mu_{j}$ maps $\mathbb{R}$ to itself follows directly from the self-adjointness of $\hat{H}(\zeta)$ when $\zeta \in \mathbb{R}$.

Henceforth (unless otherwise specified) it is assumed that $\zeta \in \mathbb{R}$. We now show that bounded perturbations of $\hat{H}$ cannot change the $\mu_{j}$ by too much.

Lemma 3.1.5. Let $m_{j} \in \mathfrak{S}\left(m_{j,-}, m_{j,+}\right)$ and $V_{j} \in \mathfrak{S}\left(V_{j,-}, V_{j,+}\right)$ for $j \in\{1,2\}$, where the $m_{j, \pm}$ and $V_{j, \pm}$ are real numbers. For $\zeta \in \mathbb{R}$ and $\lambda \in[0,1]$, set

$$
\begin{aligned}
\hat{H}(\zeta, \lambda)=D_{x} \sigma_{1}+\left(\zeta-A_{2}(x)\right) \sigma_{2} & +m_{1}(x) \sigma_{3}+V_{1}(x) \sigma_{0} \\
& +\lambda\left(\left(m_{2}(x)-m_{1}(x)\right) \sigma_{3}+\left(V_{2}(x)-V_{1}(x)\right) \sigma_{0}\right)
\end{aligned}
$$

Let $\left\{\mu_{j}(\zeta, \lambda)\right\}_{j \in \mathbb{Z}}$ denote the eigenvalues of $\hat{H}(\zeta, \lambda)$. Then the $\mu_{j}$ can be chosen real-analytic in $(\zeta, \lambda) \in \mathbb{R} \times[0,1]$ and

$$
\left|\partial_{\lambda} \mu_{j}(\zeta, \lambda)\right| \leq\left\|m_{2}-m_{1}\right\|_{\infty}+\left\|V_{2}-V_{1}\right\|_{\infty}
$$

Proof. Since $m$ and $V$ were arbitrary switch functions, it follows from Lemma 3.1.3 and [58, Theorems VII.1.7 and VII.1.8] as before that the eigenvalues and eigenprojections of $\hat{H}(\zeta, \lambda)$ are holomorphic in $(\zeta, \lambda) \in \mathbb{C}^{2}$. Since the eigenvalues are real whenever $\zeta$ and $\lambda$ are real, we have proved the first part of the lemma. Letting $\mu(\zeta, \lambda)$ denote such an eigenvalue with $\Pi_{\zeta, \lambda}=\psi_{\zeta, \lambda} \psi_{\zeta, \lambda}^{*}$ the projection onto the corresponding (one-dimensional) eigenspace, we have

$$
\hat{H}(\zeta, \lambda) \psi_{\zeta, \lambda}=\mu(\zeta, \lambda) \psi_{\zeta, \lambda}
$$

Now fix $\zeta \in \mathbb{R}$ and $\lambda \in[0,1]$, and let $h>0$. Evaluating the difference of the above between $\lambda$ and $\lambda+h$ yields

$$
\begin{aligned}
(\hat{H}(\zeta, \lambda & +h)-\hat{H}(\zeta, \lambda)) \psi_{\zeta, \lambda+h}+\hat{H}(\zeta, \lambda)\left(\psi_{\zeta, \lambda+h}-\psi_{\zeta, \lambda}\right) \\
& =\hat{H}(\zeta, \lambda+h) \psi_{\zeta, \lambda+h}-\hat{H}(\zeta, \lambda) \psi_{\zeta, \lambda}=\mu(\zeta, \lambda+h) \psi_{\zeta, \lambda+h}-\mu(\zeta, \lambda) \psi_{\zeta, \lambda} \\
& =(\mu(\zeta, \lambda+h)-\mu(\zeta, \lambda)) \psi_{\zeta, \lambda+h}+\mu(\zeta, \lambda)\left(\psi_{\zeta, \lambda+h}-\psi_{\zeta, \lambda}\right)
\end{aligned}
$$

Multiplying both sides by $\bar{\psi}_{\zeta, \lambda}$ and taking inner products, we obtain

$$
\left(\psi_{\zeta, \lambda},(\hat{H}(\zeta, \lambda+h)-\hat{H}(\zeta, \lambda)) \psi_{\zeta, \lambda+h}\right)=(\mu(\zeta, \lambda+h)-\mu(\zeta, \lambda))\left(\psi_{\zeta, \lambda}, \psi_{\zeta, \lambda+h}\right)
$$

Using that $\hat{H}(\zeta, \lambda+h)-\hat{H}(\zeta, \lambda)=h\left(\left(m_{2}(x)-m_{1}(x)\right) \sigma_{3}+\left(V_{2}(x)-V_{1}(x)\right) \sigma_{0}\right)$, we divide both sides by $h$ to get

$$
\begin{align*}
\left(\psi_{\zeta, \lambda},\left(\left(m_{2}(x)-m_{1}(x)\right) \sigma_{3}+\left(V_{2}(x)\right.\right.\right. & \left.\left.\left.-V_{1}(x)\right) \sigma_{0}\right) \psi_{\zeta, \lambda+h}\right)  \tag{3.1.4}\\
& =h^{-1}(\mu(\zeta, \lambda+h)-\mu(\zeta, \lambda))\left(\psi_{\zeta, \lambda}, \psi_{\zeta, \lambda+h}\right)
\end{align*}
$$

Since $\Pi_{\zeta, \lambda}$ is holomorphic in $\lambda$, the operator norm $\left\|\Pi_{\zeta, \lambda+h}-\Pi_{\zeta, \lambda}\right\|$ goes to zero as $h \rightarrow 0$. Hence

$$
\left\|\left(\psi_{\zeta, \lambda+h}, \psi_{\zeta, \lambda}\right) \psi_{\zeta, \lambda+h}-\psi_{\zeta, \lambda}\right\| \rightarrow 0
$$

meaning that $\left|\left(\psi_{\zeta, \lambda+h}, \psi_{\zeta, \lambda}\right)\right| \rightarrow 1$. Taking the absolute value of (3.1.4) and sending $h \rightarrow 0$, we thus obtain

$$
\left|\partial_{\lambda} \mu(\zeta, \lambda)\right| \leq\left\|m_{2}-m_{1}\right\|_{\infty}+\left\|V_{2}-V_{1}\right\|_{\infty}
$$

This completes the proof.
Next we determine the multisets $\left\{\lim _{\zeta \rightarrow \pm \infty} \mu_{j}(\zeta)\right\}_{j \in \mathbb{Z}}$, which depend on the signs of $B_{+}$ and $B_{-}$. Given two sets $A$ and $B$, we use the notation $A+B$ to denote the multiset formed by combining $A$ and $B$. That is, $x \in A+B$ if and only if $x \in A \cup B$, where the multiplicity of $x$ is 2 if $x \in A \cap B$ and 1 otherwise. Although each $\mu_{j}$ has multiplicity 1 (Lemma 3.1.3), it is possible that two distinct $\mu_{j}$ converge to the same value (as $\zeta \rightarrow \infty$ if $B_{-}<0<B_{+}$; as $\zeta \rightarrow-\infty$ if $B_{+}<0<B_{-}$).

Lemma 3.1.6. 1. Suppose $B_{-}<0<B_{+}$. Then for each $j \in \mathbb{Z}, \mu_{j, \infty}:=\lim _{\zeta \rightarrow \infty} \mu_{j}(\zeta)$ exists and belongs to $\sigma\left(\hat{H}_{+}(\zeta)\right) \cup \sigma\left(\hat{H}_{-}(\zeta)\right)$. For each $\nu \in \sigma\left(\hat{H}_{+}(\zeta)\right)+\sigma\left(\hat{H}_{-}(\zeta)\right)$, there
exists exactly one index $j \in \mathbb{Z}$ such that $\mu_{j, \infty}=\nu$.
2. Suppose $B_{+}<0<B_{-}$. Then for each $j \in \mathbb{Z}, \mu_{j,-\infty}:=\lim _{\zeta \rightarrow-\infty} \mu_{j}(\zeta)$ exists and belongs to $\sigma\left(\hat{H}_{+}(\zeta)\right) \cup \sigma\left(\hat{H}_{-}(\zeta)\right)$. For each $\nu \in \sigma\left(\hat{H}_{+}(\zeta)\right)+\sigma\left(\hat{H}_{-}(\zeta)\right)$, there exists exactly one index $j \in \mathbb{Z}$ such that $\mu_{j,-\infty}=\nu$.
3. Suppose $0<B_{+}, B_{-}$. Then for each $j \in \mathbb{Z}$, $\mu_{j, \pm \infty}:=\lim _{\zeta \rightarrow \pm \infty} \mu_{j}(\zeta)$ exists and belongs to $\sigma\left(\hat{H}_{ \pm}(\zeta)\right)$. For each $\nu \in \sigma\left(\hat{H}_{ \pm}(\zeta)\right)$, there exists exactly one index $j \in \mathbb{Z}$ such that $\mu_{j, \pm \infty}=\nu$.
4. Suppose $B_{+}, B_{-}<0$. Then for each $j \in \mathbb{Z}, \mu_{j, \pm \infty}:=\lim _{\zeta \rightarrow \pm \infty} \mu_{j}(\zeta)$ exists and belongs to $\sigma\left(\hat{H}_{\mp}(\zeta)\right)$. For each $\nu \in \sigma\left(\hat{H}_{\mp}(\zeta)\right)$, there exists exactly one index $j \in \mathbb{Z}$ such that $\mu_{j, \pm \infty}=\nu$.

Proof. Suppose $B_{-}<0<B_{+}$. Let $\left\{\nu_{k, \pm}: k \in \mathbb{N}\right\}$ denote the full set of eigenvalues of $\hat{H}_{ \pm}$. Applying the max-min principle (Theorem 4.1.5) to $\hat{H}^{2}(\zeta)$ (see [30, 88] and Section 4.1 for similar arguments), it follows that there is a bijection $\iota: \mathbb{Z} \rightarrow \mathbb{N} \times\{+,-\}$ such that for every $j \in \mathbb{Z}, \mu_{j}^{2}(\zeta) \rightarrow \nu_{\iota(j)}^{2}$ as $\zeta \rightarrow \infty$.

Now fix $j \in \mathbb{Z}$ and let $K_{j}:=\left\{i \in \mathbb{Z}: \nu_{\iota(i)}^{2}=\nu_{\iota(j)}^{2}\right\}$. (Lemma 3.1.2 implies that $K_{j}$ can have at most four elements.) By Lemma 3.1.5, the quantities $\lim _{\zeta \rightarrow \infty} \mu_{j}^{2}(\zeta)$ are continuous in shifts of $V$. Thus there is a bijection $\iota_{j}: K_{j} \rightarrow \iota\left(K_{j}\right)$ such that for all $i \in K_{j}$ and $\alpha$ sufficiently small, $\left(\mu_{i}(\zeta)+\alpha\right)^{2} \rightarrow\left(\nu_{\iota j}(i)+\alpha\right)^{2}$ as $\zeta \rightarrow \infty$. Indeed, the previous paragraph applies also to $H+\alpha$, as $V \in \mathfrak{S}\left(V_{-}, V_{+}\right)$was arbitrary. Hence $\lim _{\zeta \rightarrow \infty} \mu_{i}(\zeta)=\nu_{\iota_{j}(i)}$ for all $i \in K_{j}$, as desired.

The proofs for cases 2-4 are similar.
Lemma 3.1.7. 1. Suppose $B_{-}<0<B_{+}$. For every $M>0$, there exists $\zeta_{0} \in \mathbb{R}$ such that $\left|\mu_{j}(\zeta)\right|>M$ for all $\zeta<\zeta_{0}$ and $j \in \mathbb{Z}$.
2. Suppose $B_{+}<0<B_{-}$. For every $M>0$, there exists $\zeta_{0} \in \mathbb{R}$ such that $\left|\mu_{j}(\zeta)\right|>M$ for all $\zeta>\zeta_{0}$ and $j \in \mathbb{Z}$.

By continuity of the $\mu_{j}$, Lemma 3.1.7 implies that when $\pm B_{-}<0< \pm B_{+}$, each $\mu_{j}$ goes either to $+\infty$ or $-\infty$ as $\zeta \rightarrow \mp \infty$.

Proof. Suppose $B_{-}<0<B_{+}$. We see that

$$
\begin{align*}
\hat{H}^{2}(\zeta)=D_{x}^{2}+\left(\zeta-A_{2}(x)\right)^{2} & +m^{2}(x)-A_{2}^{\prime}(x) \sigma_{3}-m^{\prime}(x) \sigma_{2}+V^{2}(x) \\
& +2 V(x)\left(D_{x} \sigma_{1}+\left(\zeta-A_{2}(x)\right) \sigma_{2}+m(x) \sigma_{3}\right)-i V^{\prime}(x) \sigma_{1} \tag{3.1.5}
\end{align*}
$$

where $\sigma_{0}$ has been dropped from the notation. Since $V$ is bounded, we know there exists a constant $C>0$ such that

$$
\left|\left(\psi, V D_{x} \sigma_{1} \psi\right)\right| \leq C\|\psi\|\left\|\psi^{\prime}\right\|
$$

for all $\psi \in \mathcal{H}^{1}$. Since $\left(\psi, D_{x}^{2} \psi\right)=\left\|\psi^{\prime}\right\|^{2}$, it follows that the operator $D_{x}^{2}+V(x) D_{x} \sigma_{1}$ is bounded from below. The function $A_{2}$ is also bounded from below, hence

$$
\lim _{\zeta \rightarrow-\infty} \inf _{x \in \mathbb{R}}\left(\zeta-A_{2}(x)\right)^{2}=\infty
$$

Since $m, V, m^{\prime}, V^{\prime}, A_{2}^{\prime}$ are all bounded, we conclude that all eigenvalues of $\hat{H}^{2}(\zeta)$ go to infinity uniformly as $\zeta \rightarrow-\infty$. The proof for the case $B_{+}<0<B_{-}$is similar. This completes the result.

We also need the following result, which ensures that the spectral flow is well defined.

Lemma 3.1.8. For every $\alpha \in \mathcal{R}$, the set $\mathcal{T}_{\alpha}:=\left\{j: \alpha \in \operatorname{Ran}\left(\mu_{j}\right)\right\}$ is finite.

Above, $\operatorname{Ran}\left(\mu_{j}\right):=\left\{\mu_{j}(\zeta): \zeta \in \mathbb{R}\right\}$ is the range (or image) of $\mu_{j}$.
Proof. Fix $\alpha \in \mathcal{R}$. The previous lemmas imply the existence of $\bar{\zeta}>0$ such that $\mu_{j}(\zeta) \neq \alpha$ for all $j \in \mathbb{Z}$ and $|\zeta|>\bar{\zeta}$. For $\zeta \in \mathbb{R}$ and $\beta>0$, let $N(\zeta, \beta):=\left|\left\{j: \mu_{j}^{2}(\zeta)<\beta\right\}\right|$. Suppose
by contradiction that $\mathcal{T}_{\alpha}$ is infinite. Then there exists a sequence $\left(\zeta_{k}\right) \subset[-\bar{\zeta}, \bar{\zeta}]$ and a number $\zeta_{*} \in[-\bar{\zeta}, \bar{\zeta}]$ such that $\zeta_{k} \rightarrow \zeta_{*}$ and $N\left(\zeta_{k}, \alpha^{2}\right) \rightarrow \infty$ as $k \rightarrow \infty$. But Lemma 3.1.3 implies that $N\left(\zeta_{*}, \alpha^{2}+1\right)<\infty$, hence there exist $i, j \in \mathbb{Z}$ such that $\mu_{i}\left(\zeta_{*}\right) \leq-\sqrt{\alpha^{2}+1}$ and $\mu_{j}\left(\zeta_{*}\right) \geq \sqrt{\alpha^{2}+1}$. Lemma 3.1.3 and the fact that $N\left(\zeta_{k}, \alpha^{2}\right) \rightarrow \infty$ imply that for all $k$ sufficiently large, there exists $\ell \in\{i, j\}$ such that $\mu_{\ell}^{2}\left(\zeta_{k}\right)<\alpha^{2}$. Thus either $\mu_{i}$ or $\mu_{j}$ is not continuous at $\zeta_{*}$, which contradicts Lemma 3.1.4.

To compute the spectral flow in (3.0.5), it remains to determine the sign of $\mu_{j}(\zeta)$ as $\zeta \rightarrow-\infty$. We will do this first by imposing additional constraints on $\hat{H}$.

Lemma 3.1.9. Fix $\zeta \in \mathbb{R}$, let $m_{0} \in \mathbb{R}$ such that $m_{0}^{2}>\left\|A_{2}^{\prime}\right\|_{\infty}$, and define

$$
\hat{H}_{0}(\zeta):=D_{x} \sigma_{1}+\left(\zeta-A_{2}(x)\right) \sigma_{2}+m_{0} \sigma_{3}
$$

Then $\hat{H}_{0}(\zeta)$ has a spectral gap in the interval $(-\Delta, \Delta)$, with $\Delta:=\sqrt{m_{0}^{2}-\left\|A_{2}^{\prime}\right\|_{\infty}}$.
The above implies that for $\hat{H}_{0}$, if $B_{-}<0<B_{+}$, every branch converging to a positive (resp. negative) value as $\zeta \rightarrow \infty$ goes to $+\infty$ (resp. $-\infty$ ) as $\zeta \rightarrow-\infty$.

Proof. Observe that $\hat{H}_{0}^{2}(\zeta):=D_{x}^{2}+\left(\zeta-A_{2}(x)\right)^{2}+m_{0}^{2}-A_{2}^{\prime}(x) \sigma_{3}$, and the result easily follows.

We are now ready to complete the proof of Theorem 3.1.1. The idea is to treat $\hat{H}$ as a perturbation of $\hat{H}_{0}$. For $\zeta \in \mathbb{R}$ and $\left(\lambda_{1}, \lambda_{2}\right) \in[0,1]^{2}$, define

$$
\hat{H}\left(\zeta ; \lambda_{1}, \lambda_{2}\right):=D_{x} \sigma_{1}+\left(\zeta-A_{2}(x)\right) \sigma_{2}+m_{0} \sigma_{3}+\lambda_{1}\left(m(x)-m_{0}\right) \sigma_{3}+\lambda_{2} V(x) \sigma_{0}
$$

and let $\left\{\mu_{j}\left(\zeta ; \lambda_{1}, \lambda_{2}\right)\right\}_{j \in \mathbb{Z}}$ denote the eigenvalues of $\hat{H}\left(\zeta ; \lambda_{1}, \lambda_{2}\right)$. We use the shorthand $\lambda:=\left(\lambda_{1}, \lambda_{2}\right)$ and $\mu_{j, \pm \infty}(\lambda):=\lim _{\zeta \rightarrow \pm \infty} \mu_{j}(\zeta ; \lambda)$. By Lemma 3.1.5, the $\mu_{j}(\zeta ; \lambda)$ are realanalytic in $(\zeta ; \lambda) \in \mathbb{R} \times[0,1]^{2}$ with $\partial_{\lambda_{i}} \mu_{j}(\zeta ; \lambda)$ bounded uniformly in $(\zeta ; \lambda)$ for $i \in\{1,2\}$.

This means the limits as $\zeta \rightarrow \pm \infty$ of the $\mu_{j}(\zeta ; \lambda)$ depend continuously on $\lambda$. Recall that Lemmas 3.1.6, 3.1.7 and 3.1.9 give us a full description of the $\mu_{j}(\zeta ; 0,0)$.

When $B_{-}<0<B_{+}$, the $\mu_{j, \infty}(0,0)$ are known (and finite) and $\mu_{j,-\infty}(0,0)= \pm \infty$ if and only if $\pm \mu_{j, \infty}(0,0)>0$. The uniform bounds on $\partial_{\lambda_{i}} \mu_{j}$ imply that $\mu_{j,-\infty}(\lambda)=\mu_{j,-\infty}(0,0)$ for all $\lambda \in[0,1]^{2}$. Combined with Lemma 3.1.3 and the fact that the multiset $\left\{\mu_{j, \infty}(1,1)\right\}_{j \in \mathbb{Z}}$ is known (Lemmas 3.1.2 and 3.1.6), this will allow us to obtain the limits as $\zeta \rightarrow \pm \infty$ of each $\mu_{j}(\zeta ; 1,1)$ via a smooth transition of $\lambda$ from $(0,0)$ to $(1,1)$. We find it easiest to first fix $\lambda_{2}=0$ while smoothly varying $\lambda_{1}$ from 0 to 1 ; then fix $\lambda_{1}=1$ while smoothly varying $\lambda_{2}$ from 0 to 1 . The case $B_{+}<0<B_{-}$is handled similarly.

When $0<B_{+}, B_{-}$, Lemma 3.1.6 gives us the multisets $L_{ \pm}:=\left\{\mu_{j, \pm \infty}(\lambda)\right\}_{j \in \mathbb{Z}}$ for all $\lambda \in[0,1]^{2}$. The only thing left is to pair each element of $L_{-}$with an element of $L_{+}$ (i.e. for each $\mu_{j,-\infty}(\lambda) \in L_{-}$, determine $\left.\mu_{j,+\infty}(\lambda)\right)$. When $\lambda=(0,0)$, this pairing follows immediately from Lemma 3.1.9. That is (using a natural choice of indices),

$$
\begin{aligned}
\ldots, \quad \mu_{-1, \pm \infty}(0,0)=-\sqrt{2\left|B_{ \pm}\right|+m_{0}^{2}}, & \mu_{0, \pm \infty}(0,0)=m_{0} \\
& \mu_{1, \pm \infty}(0,0)=\sqrt{2\left|B_{ \pm}\right|+m_{0}^{2}}, \quad \ldots
\end{aligned}
$$

Since $\mu_{i, \pm \infty}(\lambda) \neq \mu_{j, \pm \infty}(\lambda)$ whenever $i \neq j$, it is straightforward to obtain the limits of each $\mu_{j}$ for $\lambda=(1,1)$ (again by a smooth transition from $\lambda=(0,0)$ to $\lambda=(1,1)$ ). The case $B_{+}, B_{-}<0$ is handled similarly.

Proof of Theorem 3.1.1. Take $m_{0}$ as in Lemma 3.1.9. As in Lemma 3.1.5 the eigenvalues and eigenprojections of $\hat{H}(\zeta ; \lambda)$ are holomorphic in $(\zeta ; \lambda)$. It then follows from Lemma 3.1.5 that

$$
\begin{equation*}
\left|\partial_{\lambda_{1}} \mu(\zeta, \lambda)\right| \leq\left\|m-m_{0}\right\|_{\infty}, \quad\left|\partial_{\lambda_{2}} \mu(\zeta, \lambda)\right| \leq\|V\|_{\infty} \tag{3.1.6}
\end{equation*}
$$

for all eigenvalues $\mu(\zeta ; \lambda)$ of $\hat{H}(\zeta ; \lambda)$. Define

$$
\hat{H}_{ \pm}\left(\zeta ; \lambda_{1}, \lambda_{2}\right):=D_{x} \sigma_{1}+\left(\zeta-x B_{ \pm}\right) \sigma_{2}+m_{0} \sigma_{3}+\lambda_{1}\left(m_{ \pm}-m_{0}\right) \sigma_{3}+\lambda_{2} V_{ \pm} \sigma_{0}
$$

Lemma 3.1.2 states that the spectrum of $\hat{H}_{ \pm}(\zeta ; 0,0)$ consists entirely of eigenvalues and is given by

$$
\begin{aligned}
\sigma\left(\hat{H}_{ \pm}(\zeta ; 0,0)\right) & =\left\{\varepsilon \sqrt{2 k\left|B_{ \pm}\right|+m_{0}^{2}}: \varepsilon \in\{-1,1\}, k \in \mathbb{N}_{+}\right\} \bigcup\left\{m_{0} \operatorname{sgn}\left(B_{ \pm}\right)\right\} \\
& =:\left\{\tilde{\nu}_{k, \pm}: k \in \mathbb{Z}\right\} .
\end{aligned}
$$

Note that $\left|\tilde{\nu}_{k, \pm}\right| \geq\left|m_{0}\right|>0$ for all $k$.

1. Suppose $B_{-}<0<B_{+}$. Recall that $\left\{\mu_{j}(\zeta ; \lambda)\right\}_{j \in \mathbb{Z}}$ denotes the (holomorphic in $(\zeta ; \lambda)$ ) eigenvalues of $\hat{H}(\zeta ; \lambda)$. Then there is a bijection $\iota: \mathbb{Z} \rightarrow \mathbb{Z} \times\{+,-\}$ such that $\lim _{\zeta \rightarrow \infty} \mu_{j}(\zeta ; 0,0)=\tilde{\nu}_{\iota(j)}$ for all $j \in \mathbb{Z}$. Lemma 3.1.7 asserts that $\left|\mu_{j}(\zeta ; 0,0)\right| \rightarrow \infty$ as $\zeta \rightarrow-\infty$. Hence Lemma 3.1.9 implies that $\mu_{j,-\infty}(0,0)= \pm \infty$ if and only if $\pm \tilde{\nu}_{\iota(j)}>0$. That is, the levels $\tilde{\nu}_{\iota(j)}>0$ correspond to branches $\mu_{j}(\cdot ; 0,0)$ that go to $+\infty$ at $-\infty$, while the levels $\tilde{\nu}_{\iota(j)}<0$ correspond to branches $\mu_{j}(\cdot ; 0,0)$ that go to $-\infty$ at $-\infty$. Thus 0 (or any value in $\left.\left(-\left|m_{0}\right|,\left|m_{0}\right|\right)\right)$ separates the branches $\mu_{j}(\cdot ; 0,0)$ that go to $+\infty$ at $-\infty$ from those that go to $-\infty$ at $-\infty$, as demonstrated by Figure 3.1 (bottom right panel).

We now analyze the spectrum of $\hat{H}(\zeta ; \lambda)$ as $\lambda$ is continuously deformed from $(0,0)$ to $(1,0)$ to $(1,1)$, thus separating the effects of $m$ and $V$. Note that we can choose $m_{0}$ such that the spectral flow of $H$ through $\alpha$ is well defined when $\lambda \in\{(0,0),(1,0)\}$ (by definition the spectral flow is well defined when $\lambda=(1,1)$ ). We follow the convention that $\mu_{j}(\zeta ; \lambda)<\mu_{j+1}(\zeta ; \lambda)$ and $\mu_{1, \infty}(0,0)=\left|m_{0}\right|$. This means $\mu_{j, \infty}(0,0)>0$ and $\mu_{j,-\infty}(0,0)=+\infty$ for all $j>0$, while $\mu_{j, \infty}(0,0)<0$ and $\mu_{j,-\infty}(0,0)=-\infty$ for all $j \leq 0$. From (3.1.6), it follows that for all $\lambda \in[0,1]^{2}, \mu_{j,-\infty}(\lambda)=+\infty$
for all $j>0$ and $\mu_{j,-\infty}(\lambda)=-\infty$ for all $j \leq 0$. Moreover, (3.1.6) implies that $\mu_{1, \infty}(1,0)=\max \left\{m_{+}, m_{-}\right\}$and $\mu_{0, \infty}(1,0)=\min \left\{m_{+}, m_{-}\right\}$. By Lemma 3.1.6, the values $\mu_{j, \infty}(1,0)$ can be read off directly from (3.1.3) with $V_{ \pm}=0$, as they are exactly the elements of $\sigma\left(\hat{H}_{+}(0 ; 1,0)\right)+\sigma\left(\hat{H}_{-}(0 ; 1,0)\right.$ (in increasing order with the limit of $\mu_{1}$ already determined above). This confirms Theorem 3.1.1 in the case that $V \equiv 0$.

Now we set $\lambda_{1}=1$ and analyze the transition of $\lambda_{2}$ from 0 to 1 . Let $\left\{\nu_{0, i, \pm}\right\}_{i \in \mathbb{Z}}$ denote the eigenvalues of $\hat{H}_{ \pm}(\zeta ; 1,0)$, where $\nu_{0, i, \pm}<\nu_{0, i+1, \pm}$ for all $i \in \mathbb{Z}$. The eigenvalues of $\hat{H}_{ \pm}\left(\zeta ; 1, \lambda_{2}\right)$ are then given by $\nu_{i, \pm}\left(\lambda_{2}\right):=\nu_{0, i, \pm}+\lambda_{2} V_{ \pm}$for $i \in \mathbb{Z}$. To each $\nu_{i, \pm}\left(\lambda_{2}\right)$ is associated a unique index $j=j\left(\lambda_{2}, i, \pm\right)$ such that $\mu_{j\left(\lambda_{2}, i, \pm\right), \infty}\left(1, \lambda_{2}\right)=\nu_{i, \pm}\left(\lambda_{2}\right)$. Let $\varepsilon_{i, \pm}\left(\lambda_{2}\right) \in\{1,-1\}$ such that $\varepsilon_{i, \pm}\left(\lambda_{2}\right)=1$ if and only if $\mu_{j\left(\lambda_{2}, i, \pm\right),-\infty}\left(1, \lambda_{2}\right)=+\infty$. It follows that

$$
\begin{align*}
& \operatorname{SF}\left(H\left(1, \lambda_{2}\right) ; \alpha\right)= \\
& \qquad \sum_{\delta \in\{+,-\}}\left(\left|\left\{i: \nu_{i, \delta}\left(\lambda_{2}\right)>\alpha, \varepsilon_{i, \delta}\left(\lambda_{2}\right)=-1\right\}\right|-\left|\left\{i: \nu_{i, \delta}\left(\lambda_{2}\right)<\alpha, \varepsilon_{i, \delta}\left(\lambda_{2}\right)=1\right\}\right|\right) . \tag{3.1.7}
\end{align*}
$$

By (3.1.6), we know that if $\nu_{i,+}\left(\lambda_{2}\right) \notin\left\{\nu_{i,-}\left(\lambda_{2}\right)\right\}_{i \in \mathbb{Z}}$ for all $\lambda_{2}$ in an open interval $(a, b)$, then $\varepsilon_{i,+}\left(\lambda_{2}\right)$ is constant over $\lambda_{2} \in(a, b)$. For values $\lambda_{2}^{*}$ and indices $i, j$ such that $\nu_{i,+}\left(\lambda_{2}^{*}\right)=\nu_{j,-}\left(\lambda_{2}^{*}\right)$, Lemma 3.1.3 implies that $\varepsilon_{i,+}\left(\lambda_{2}\right)$ and $\varepsilon_{j,-}\left(\lambda_{2}\right)$ trade signs across $\lambda_{2}^{*}$; that is, $\varepsilon_{i,+}\left(\lambda_{2}^{*}+\eta\right)=\varepsilon_{j,-}\left(\lambda_{2}^{*}-\eta\right)$ and $\varepsilon_{j,-}\left(\lambda_{2}^{*}+\eta\right)=\varepsilon_{i,+}\left(\lambda_{2}^{*}-\eta\right)$ for all $\eta>0$ sufficiently small. But this trade of signs has no effect on the spectral flow (i.e. if $\lim _{\zeta \rightarrow \infty} \tilde{\mu}_{1}(\zeta)=1$ and $\lim _{\zeta \rightarrow \infty} \tilde{\mu}_{0}(\zeta)=-1$ for some smooth branches of spectrum $\tilde{\mu}_{1}$ and $\tilde{\mu}_{0}$, then any spectral flow is independent of whether $\lim _{\zeta \rightarrow-\infty}\left(\tilde{\mu}_{1}(\zeta), \tilde{\mu}_{0}(\zeta)\right)=$ $(+\infty,-\infty)$ or $\left.\lim _{\zeta \rightarrow-\infty}\left(\tilde{\mu}_{1}(\zeta), \tilde{\mu}_{0}(\zeta)\right)=(-\infty,+\infty)\right)$. This means that when evaluat-
ing the right-hand side of (3.1.7), we can replace $\varepsilon_{i, \delta}\left(\lambda_{2}\right)$ by $\varepsilon_{i, \delta}(0)$ to obtain

$$
\operatorname{SF}(H(1,1) ; \alpha)=\sum_{\delta \in\{+,-\}}\left(\left|\left\{i \leq 0: \nu_{0, i, \delta}>\alpha-V_{\delta}\right\}\right|-\left|\left\{i>0: \nu_{0, i, \delta}<\alpha-V_{\delta}\right\}\right|\right),
$$

where we use the convention that $\varepsilon_{i, \pm}(0)=1$ if and only if $i>0$. To verify that the above expression yields Theorem 3.1.1 is a straightforward but tedious exercise. We will do so assuming that $m_{-}<0<m_{+},\left|m_{-}\right| \leq\left|m_{+}\right|$and $\alpha-V_{-}, \alpha-V_{+}>0$, and leave the other cases (which are handled similarly) to the reader.

Under the above assumptions, it follows that

$$
\begin{equation*}
\mathrm{SF}(H(1,1) ; \alpha)=M_{0}-\sum_{\delta \in\{+,-\}}\left|\left\{i>0: \nu_{0, i, \delta}<\alpha-V_{\delta}\right\}\right|=: M_{0}-M_{+}-M_{-}, \tag{3.1.8}
\end{equation*}
$$

where

$$
M_{0}= \begin{cases}1, & \alpha-V_{-}<\left|m_{-}\right| \\ 0, & \text { else }\end{cases}
$$

$$
M_{+}= \begin{cases}0, & \alpha-V_{+}<m_{+} \\ k ; & \sqrt{2(k-1) B_{+}+m_{+}^{2}}<\alpha-V_{+}<\sqrt{2 k B_{+}+m_{+}^{2}}, \quad k \in \mathbb{N}_{+}\end{cases}
$$

and

$$
M_{-}= \begin{cases}0, & \alpha-V_{-}<\sqrt{2\left|B_{-}\right|+m_{-}^{2}} \\ k ; & \sqrt{2 k\left|B_{-}\right|+m_{-}^{2}}<\alpha-V_{-}<\sqrt{2(k+1)\left|B_{-}\right|+m_{-}^{2}}, \quad k \in \mathbb{N}_{+}\end{cases}
$$

We will now show that (3.1.8) agrees with Theorem 3.1.1. Simplifying the formula (3.1.2) from Theorem 3.1.1, we obtain that

$$
N\left(H_{+} ; \alpha\right)= \begin{cases}0 ; & \alpha-V_{+}<\sqrt{2 B_{+}+m_{+}^{2}}, \\ k ; & \sqrt{2 k B_{+}+m_{+}^{2}}<\alpha-V_{+}<\sqrt{2(k+1) B_{+}+m_{+}^{2}}, \quad k \in \mathbb{N}_{+}\end{cases}
$$

and

$$
N\left(H_{-} ; \alpha\right)= \begin{cases}0 ; & \alpha-V_{-}<\sqrt{2\left|B_{-}\right|+m_{-}^{2}}, \\ k ; & \sqrt{2 k\left|B_{-}\right|+m_{-}^{2}}<\alpha-V_{-}<\sqrt{2(k+1)\left|B_{-}\right|+m_{-}^{2}}, \quad k \in \mathbb{N}_{+}\end{cases}
$$

Note from (3.1.1) that in our case,

$$
\begin{aligned}
& I\left(H_{+} ; \alpha\right)=\operatorname{sgn}\left(\alpha-V_{+}-m_{+}\right)\left(N\left(H_{+} ; \alpha\right)+\frac{1}{2}\right), \\
& I\left(H_{-} ; \alpha\right)=-\operatorname{sgn}\left(\alpha-V_{-}+m_{-}\right)\left(N\left(H_{-} ; \alpha\right)+\frac{1}{2}\right) .
\end{aligned}
$$

Observe that $N\left(H_{ \pm} ; \alpha\right)=0$ if $\operatorname{sgn}\left(\alpha-V_{ \pm} \mp m_{ \pm}\right)<0$, hence

$$
I\left(H_{+} ; \alpha\right)= \begin{cases}N\left(H_{+} ; \alpha\right)+\frac{1}{2}, & \alpha-V_{+}-m_{+}>0 \\ -\frac{1}{2}, & \alpha-V_{+}-m_{+}<0\end{cases}
$$

and

$$
I\left(H_{-} ; \alpha\right)= \begin{cases}-N\left(H_{-} ; \alpha\right)-\frac{1}{2}, & \alpha-V_{-}+m_{-}>0 \\ \frac{1}{2}, & \alpha-V_{-}+m_{-}<0\end{cases}
$$

Observe that $N\left(H_{-} ; \alpha\right)=M_{-}$and

$$
N\left(H_{+} ; \alpha\right)= \begin{cases}M_{+}, & \alpha-V_{+}<m_{+} \\ M_{+}-1, & \text { else }\end{cases}
$$

Thus

$$
I\left(H_{+} ; \alpha\right)= \begin{cases}M_{+}-\frac{1}{2}, & \alpha-V_{+}-m_{+}>0 \\ -\frac{1}{2}, & \alpha-V_{+}-m_{+}<0\end{cases}
$$

and

$$
I\left(H_{-} ; \alpha\right)= \begin{cases}-M_{-}-\frac{1}{2}, & \alpha-V_{-}+m_{-}>0 \\ \frac{1}{2}, & \alpha-V_{-}+m_{-}<0\end{cases}
$$

Since $M_{ \pm}=0$ whenever $\alpha-V_{ \pm} \mp m_{ \pm}<0$, we conclude that

$$
I\left(H_{-} ; \alpha\right)-I\left(H_{+} ; \alpha\right)= \begin{cases}-M_{-}-M_{+}, & \alpha-V_{-}+m_{-}>0 \\ -M_{-}-M_{+}+1, & \alpha-V_{-}+m_{-}<0\end{cases}
$$

This agrees with (3.1.8), as desired.
2. Suppose $0<B_{+}, B_{-}$. The argument is similar to case 1 , only now each $\mu(\zeta ; \lambda)$ converges also as $\zeta \rightarrow-\infty$. By Lemmas 3.1.3 and 3.1.9, we know that the branch $\mu(\zeta ; 0,0)$ that converges to $m_{0}$ as $\zeta \rightarrow-\infty$ also converges to $m_{0}$ as $\zeta \rightarrow+\infty$. Thus (3.1.6) implies that the branch $\mu(\zeta ; 1,0)$ that converges to $m_{-}$as $\zeta \rightarrow-\infty$ must converge to $m_{+}$as $\zeta \rightarrow+\infty$. Finally, this implies that the branch $\mu(\zeta ; 1,1)$ that converges to $m_{-}+V_{-}$as $\zeta \rightarrow-\infty$ must converge to $m_{+}+V_{+}$as $\zeta \rightarrow \infty$, and the result follows.

The case $B_{+}<0<B_{-}$is handled similarly to case 1 ; the case $B_{+}, B_{-}<0$ is handled similarly to case 2 . This completes the result.

### 3.2 Physical observable

Let $P(y)=P \in \mathfrak{S}(0,1)$ and $\varphi \in \mathfrak{S}\left(0,1 ; E_{1}, E_{2}\right)$ for some $\left[E_{1}, E_{2}\right] \subset \mathcal{R}$. As in (1.0.1) and Section 2, define the interface conductivity associated to $H$ in (3.0.2) by

$$
\begin{equation*}
\sigma_{I}:=\operatorname{Tr} i[H, P] \varphi^{\prime}(H) \tag{3.2.1}
\end{equation*}
$$

The goal of this section is to relate $\sigma_{I}$ to the spectral flow from Section 3.1, and to prove its stability with respect to perturbations of $H$. To do so, it is convenient to use the framework of pseudo-differential operators ( $\Psi$ DOs) and the Helffer-Sjöstrand formula; the notation and main results we need are summarized in Appendix A.1. Let

$$
\sigma(x, \xi, \zeta):=\xi \sigma_{1}+\left(\zeta-A_{2}(x)\right) \sigma_{2}+m(x) \sigma_{3}+V(x) \sigma_{0}
$$

denote the Weyl symbol of $H$ (so that $H=\mathrm{Op}(\sigma)$ ). Similarly, define

$$
\sigma_{ \pm}(x, \xi, \zeta):=\xi \sigma_{1}+\left(\zeta-x B_{ \pm}\right) \sigma_{2}+m_{ \pm} \sigma_{3}+V_{ \pm} \sigma_{0}
$$

so that $H_{ \pm}=\mathrm{Op}\left(\sigma_{ \pm}\right)$. Note that as opposed to the setting described by (H1), the symbols $\sigma_{ \pm}$still depend on the spatial variable $x$ to model a constant magnetic field.

As mentioned at the beginning of this section, $\sigma$ is not elliptic because its eigenvalues can remain small even as $x$ and $\zeta$ get large. Still, using the fact that

$$
\sigma^{2}(x, \xi, \zeta)=\xi^{2}+\left(\zeta-A_{2}(x)\right)^{2}+m^{2}(x)+V^{2}(x)+2 V(x)\left(\xi \sigma_{1}+\left(\zeta-x B_{ \pm}\right) \sigma_{2}+m_{ \pm} \sigma_{3}\right)
$$

we obtain that

$$
\begin{equation*}
|\operatorname{det} \sigma(x, \xi, \zeta)|^{1 / 2} \geq c_{1}\left\langle\xi, \zeta-A_{2}(x)\right\rangle-c_{2} \tag{3.2.2}
\end{equation*}
$$

for some $0<c_{1}<1$ and $c_{2}>0$. In order to make use of this inequality, we need to verify

Lemma 3.2.1. The map $(x, \xi, \zeta) \mapsto\left\langle\xi, \zeta-A_{2}(x)\right\rangle$ is an order function.

We refer to Section A. 1 for the definitions of an order function and $\langle\cdot\rangle$.

Proof. It is known (see e.g. [89]) that $\mathbb{R}^{2} \ni Y \mapsto\langle Y\rangle$ is an order function. Thus there exist positive constants $C_{0}$ and $N_{0}$ such that

$$
\left\langle\xi_{1}, \zeta_{1}-A_{2}\left(x_{1}\right)\right\rangle \leq C_{0}\left\langle\xi_{2}-\xi_{1}, \zeta_{2}-A_{2}\left(x_{2}\right)-\left(\zeta_{1}-A_{2}\left(x_{1}\right)\right)\right\rangle^{N_{0}}\left\langle\xi_{2}, \zeta_{2}-A_{2}\left(x_{2}\right)\right\rangle
$$

for all $\left(x_{1}, \xi_{1}, \zeta_{1}\right),\left(x_{2}, \xi_{2}, \zeta_{2}\right) \in \mathbb{R}^{3}$. We write

$$
\left(\zeta_{2}-A_{2}\left(x_{2}\right)-\left(\zeta_{1}-A_{2}\left(x_{1}\right)\right)\right)^{2} \leq 2\left(\left(\zeta_{2}-\zeta_{1}\right)^{2}+\left(A_{2}\left(x_{2}\right)-A_{2}\left(x_{1}\right)\right)^{2}\right.
$$

and seek to bound the second term on the above right-hand side. We have

$$
\begin{aligned}
\left(A_{2}\left(x_{2}\right)-A_{2}\left(x_{1}\right)\right)^{2} & =\left(x_{2} B\left(x_{2}\right)-x_{1} B\left(x_{1}\right)\right)^{2}=\left(\left(x_{2}-x_{1}\right) B\left(x_{2}\right)+x_{1}\left(B\left(x_{2}\right)-B\left(x_{1}\right)\right)\right)^{2} \\
& \leq 2\|B\|_{\infty}^{2}\left(x_{2}-x_{1}\right)^{2}+2 x_{1}^{2}\left(B\left(x_{2}\right)-B\left(x_{1}\right)\right)^{2}
\end{aligned}
$$

Since $B\left(x_{2}\right)-B\left(x_{1}\right)$ vanishes whenever $x_{1}$ and $x_{2}$ are sufficiently large and of the same sign, there exist positive constants $C_{1}$ and $C_{2}$ such that $x_{1}^{2}\left(B\left(x_{2}\right)-B\left(x_{1}\right)\right)^{2} \leq C_{1}\left(x_{2}-x_{1}\right)^{2}+C_{2}$. We conclude that

$$
\left\langle\xi_{2}-\xi_{1}, \zeta_{2}-A_{2}\left(x_{2}\right)-\left(\zeta_{1}-A_{2}\left(x_{1}\right)\right)\right\rangle \leq C\left\langle x_{2}-x_{1}, \xi_{2}-\xi_{1}, \zeta_{2}-\zeta_{2}\right\rangle
$$

for some $C>0$, and the result is complete.

We begin by showing that $\sigma_{I}$ is well defined. To do this, we will need two useful decay properties for symbols of related operators.

Lemma 3.2.2. If $z \in \mathbb{C}$ such that $\Im z \neq 0$, then $(z-H)^{-1}=\operatorname{Op}\left(r_{z}\right)$ for some $r_{z} \in$ $S\left(\left\langle\xi, \zeta-A_{2}(x)\right\rangle^{-1}\right)$.

Note that $S\left(\left\langle\zeta, \zeta-A_{2}(x)\right\rangle^{-1}\right)$ is well defined, by Lemma 3.2.1.
Proof. Since $H$ is self-adjoint, we know that $(z-H)^{-1}$ is well defined and bounded. To obtain bounds for the symbol of $(z-H)^{-1}$ (and show that it is a $\Psi D O$ in the first place), we use Beals's criterion presented in [34, Proposition 8.3]. This result states that $(z-H)^{-1}=\mathrm{Op}\left(r_{z}\right)$ for some $r_{z} \in S(1)$ if and only if for any collection of linear forms

$$
\ell_{1}(x, y, \xi, \zeta), \ell_{2}(x, y, \xi, \zeta), \ldots, \ell_{N}(x, y, \xi, \zeta)
$$

on $\mathbb{R}^{4}$, the operator $a d_{L_{1}} \circ \cdots \circ a d_{L_{N}} \circ(z-H)^{-1}$ is bounded in $L^{2}\left(\mathbb{R}^{2}\right) \otimes \mathbb{C}^{2}$, where $L_{j}:=$ $\operatorname{Op}\left(\ell_{j}\right)$ and $a d_{A} B:=[A, B]$. Since $\sigma_{\xi}=\sigma_{1}$ and $\sigma_{\zeta}=\sigma_{2}$ are constant and $\sigma_{x}$ is bounded, it is clear that $\left[L_{j}, H\right]$ is bounded for any such $L_{j}$. Thus the identity $\left[\mathcal{O},(z-H)^{-1}\right]=$ $(z-H)^{-1}[\mathcal{O}, H](z-H)^{-1}$ easily implies that $(z-H)^{-1}=O p\left(r_{z}\right)$ for some $r_{z} \in S(1)$.

By (3.2.2) and the composition calculus, we have

$$
(z-\sigma) \sharp(z-\sigma)^{-1}=1+b_{z},
$$

where $b_{z} \in S\left(\left\langle\xi, \zeta-A_{2}(x)\right\rangle^{-2}\right)$. Indeed, all derivatives of $\sigma$ are bounded and $(z-\sigma)^{-1} \in$ $S\left(\left\langle\xi, \zeta-A_{2}(x)\right\rangle^{-1}\right)$. Letting $G_{z}:=\mathrm{Op}\left((z-\sigma)^{-1}\right)$ and $B_{z}:=\mathrm{Op}\left(b_{z}\right)$, this means

$$
(z-H) G_{z}=1+B_{z}
$$

Applying $(z-H)^{-1}$ to both sides (on the left), we get

$$
(z-H)^{-1}=G_{z}-(z-H)^{-1} B_{z}
$$

The first term on the right-hand side has symbol in $S\left(\left\langle\xi, \zeta-A_{2}(x)\right\rangle^{-1}\right)$ and the second term has symbol in $S\left(\left\langle\xi, \zeta-A_{2}(x)\right\rangle^{-2}\right) \subset S\left(\left\langle\xi, \zeta-A_{2}(x)\right\rangle^{-1}\right)$. Therefore, $r_{z} \in S\left(\left\langle\xi, \zeta-A_{2}(x)\right\rangle^{-1}\right)$ as desired.

Lemma 3.2.3. For any $\Phi \in \mathcal{C}_{c}^{\infty}\left(E_{1}, E_{2}\right)$, we have $\Phi(H) \in \operatorname{Op}\left(S\left(\langle x, \xi, \zeta\rangle^{-\infty}\right)\right)$.

Proof. For any $p>0$, we can write $\Phi(H)=(i-H)^{-p} \Phi_{p}(H)$ with $\Phi_{p} \in \mathcal{C}_{c}^{\infty}\left(E_{1}, E_{2}\right)$. By Lemma 3.2.2 and the composition calculus, this means $\Phi(H) \in \operatorname{Op}\left(S\left(\left\langle\xi, \zeta-A_{2}(x)\right\rangle^{-\infty}\right)\right)$. Since $H_{ \pm}$has a spectral gap in $\left[E_{1}, E_{2}\right]$, we know that $\Phi\left(H_{ \pm}\right)=0$. Thus we can write $\Phi(H)=\phi(x)\left(\Phi(H)-\Phi\left(H_{+}\right)\right)+(1-\phi(x))\left(\Phi(H)-\Phi\left(H_{-}\right)\right)$, for some $\phi \in \mathfrak{S}(0,1)$. The Helffer-Sjöstrand formula (A.1.6) implies that

$$
\Phi(H)-\Phi\left(H_{+}\right)=\frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\Phi}(z)(z-H)^{-1}\left(H-H_{+}\right)\left(z-H_{+}\right)^{-1} d^{2} z
$$

Since $\sigma-\sigma_{+}$vanishes whenever $x$ is sufficiently large, it follows that $\Phi(H)-\Phi\left(H_{+}\right) \in$ $\operatorname{Op}\left(S\left(\left\langle x_{+}\right\rangle^{-\infty}\right)\right)$, where $x_{+}:=\max \{x, 0\}$. Since $\phi$ vanishes whenever $-x$ is sufficiently large, we conclude that $\phi(x)\left(\Phi(H)-\Phi\left(H_{+}\right)\right) \in \operatorname{Op}\left(S\left(\langle x\rangle^{-\infty}\right)\right)$. The same reasoning shows that $(1-\phi(x))\left(\Phi(H)-\Phi\left(H_{-}\right)\right) \in \operatorname{Op}\left(S\left(\langle x\rangle^{-\infty}\right)\right)$. We have thus shown that $\Phi(H) \in$ $\operatorname{Op}\left(S\left(\left\langle\zeta, \zeta-A_{2}(x)\right\rangle^{-\infty}\right) \cap S\left(\langle x\rangle^{-\infty}\right)\right)$. By interpolation, the result is complete.

Lemma 3.2.4. For any $\Phi \in \mathcal{C}_{c}^{\infty}\left(E_{1}, E_{2}\right)$, the operator $[H, P] \Phi(H)$ is trace-class.

Proof. We have $[H, P]=-i P^{\prime}(y) \sigma_{2}$ with $P^{\prime} \in \mathcal{C}_{c}^{\infty}$, hence $[H, P] \in \operatorname{Op}\left(S\left(\langle y\rangle^{-\infty}\right)\right)$. The result then follows from Lemma 3.2.3 and the composition calculus.

Now that we have shown that $\sigma_{I}$ is well defined, we relate it to the spectral flow.

Theorem 3.2.5. For any $\alpha \in\left[E_{1}, E_{2}\right]$, we have $2 \pi \sigma_{I}=S F(H ; \alpha)$.

Combining Theorems 3.1.1 and 3.2.5, we obtain an explicit formula for $\sigma_{I}$. In particular, $\sigma_{I}$ is quantized and independent of compact perturbations in $m, B$, and $V$. Moreover, $\sigma_{I}$ is stable with respect to sufficiently small changes in $m_{ \pm}, B_{ \pm}$and $V_{ \pm}$.

Proof. We follow the arguments presented in [10, Section A]. Recall Lemmas 3.1.3 and 3.1.4, which state that the spectrum of $\hat{H}(\zeta)$ for $\zeta \in \mathbb{R}$ consists entirely of simple eigenvalues $\left\{\mu_{j}(\zeta)\right\}_{j \in \mathbb{Z}}$, where each $\mu_{j}: \mathbb{R} \rightarrow \mathbb{R}$ is analytic. For $j \in \mathbb{Z}$, let $\psi_{j}(\cdot, \zeta)$ denote the normalized eigenfunction (unique up to a phase) of $\hat{H}(\zeta)$ corresponding to eigenvalue $\mu_{j}(\zeta)$. Using Lemma 3.1.8, let $\mathcal{J} \subset \mathbb{Z}$ be the finite set of indices corresponding to branches $\mu_{j}$ that ever enter the interval $\left[E_{1}, E_{2}\right]$. For any $\Phi \in \mathcal{C}_{c}^{\infty}\left(E_{1}, E_{2}\right)$, the Schwartz kernel (see Section A.1) of $\Phi(H)$ is thus given by

$$
\begin{equation*}
k_{\Phi}\left(x, x^{\prime} ; y-y^{\prime}\right)=\int_{\mathbb{R}} \sum_{j \in \mathcal{J}} \Phi\left(\mu_{j}(\zeta)\right) \psi_{j}(x, \zeta) \psi_{j}^{*}\left(x^{\prime}, \zeta\right) \frac{e^{i\left(y-y^{\prime}\right) \zeta}}{2 \pi} d \zeta . \tag{3.2.3}
\end{equation*}
$$

Lemmas 3.1.6 and 3.1.7 imply that each $\mu_{j}(\zeta)$ escapes $\left[E_{1}, E_{2}\right.$ ] whenever $|\zeta|$ is sufficiently large, hence the integrand in (3.2.3) is compactly supported in $\zeta$.

By Lemma 2.1.2, we obtain that $\sigma_{I}=\operatorname{Tr} i[\Psi(H), P] \varphi^{\prime}(H)$ for any $\Psi \in \mathcal{C}_{c}^{\infty}\left(E_{1}, E_{2}\right)$ that satisfies $\Psi(\lambda)=\lambda$ for all $\lambda$ in some open interval containing $\operatorname{supp}\left(\varphi^{\prime}\right)$. The kernel of $[\Psi(H), P] \varphi^{\prime}(H)$ is

$$
t\left(x, x^{\prime} ; y, y^{\prime}\right)=\int_{\mathbb{R}^{2}}\left(P\left(y^{\prime \prime}\right)-P(y)\right) k_{\Psi}\left(x, x^{\prime \prime} ; y-y^{\prime \prime}\right) k_{\varphi^{\prime}}\left(x^{\prime \prime}, x^{\prime} ; y^{\prime \prime}-y^{\prime}\right) d x^{\prime \prime} d y^{\prime \prime}
$$

where $k_{\Phi}$ for $\Phi \in \mathcal{C}_{c}^{\infty}\left(E_{1}, E_{2}\right)$ is given by (3.2.3). It follows from (A.1.7) that

$$
\begin{aligned}
\sigma_{I} & =i \operatorname{tr} \int_{\mathbb{R}^{2}} t(x, x ; y, y) d x d y \\
& =i \operatorname{tr} \int_{\mathbb{R}^{4}}\left(P\left(y^{\prime}\right)-P(y)\right) k_{\Psi}\left(x, x^{\prime} ; y-y^{\prime}\right) k_{\varphi^{\prime}}\left(x^{\prime}, x ; y^{\prime}-y\right) d x^{\prime} d y^{\prime} d x d y .
\end{aligned}
$$

Changing integration variables $\left(y, y^{\prime}\right) \rightarrow\left(z, y^{\prime}\right)$ with $z=y-y^{\prime}$, and using that $\int_{\mathbb{R}} P\left(y^{\prime}\right)-$ $P\left(y^{\prime}+z\right) d y^{\prime}=-z($ which follows from $P \in \mathfrak{S}(0,1)$ ), we obtain

$$
\sigma_{I}=-i \operatorname{tr} \int_{\mathbb{R}^{3}} z k_{\Psi}\left(x, x^{\prime} ; z\right) k_{\varphi^{\prime}}\left(x^{\prime}, x ;-z\right) d x^{\prime} d x d z
$$

By Parseval and using that $k_{\varphi^{\prime}}^{*}\left(x, x^{\prime} ; z\right)=k_{\varphi^{\prime}}\left(x^{\prime}, x ;-z\right)$, we have

$$
\sigma_{I}=\frac{1}{2 \pi} \operatorname{tr} \int_{\mathbb{R}^{3}} \partial_{\zeta} \hat{k}_{\Psi}\left(x, x^{\prime} ; \zeta\right) \hat{k}_{\varphi^{\prime}}^{*}\left(x, x^{\prime}, \zeta\right) d x^{\prime} d x d \zeta
$$

Note that for any $\Phi \in \mathcal{C}_{c}^{\infty}\left(E_{1}, E_{2}\right)$,

$$
\hat{k}_{\Phi}\left(x, x^{\prime} ; \zeta\right)=\sum_{j=1}^{J} \Phi\left(\mu_{j}(\zeta)\right) \psi_{j}(x, \zeta) \psi_{j}^{*}\left(x^{\prime}, \zeta\right)
$$

hence

$$
\begin{equation*}
2 \pi \sigma_{I}=\operatorname{tr} \int_{\mathbb{R}^{3}} \sum_{j, k=1}^{J} \partial_{\zeta}\left(\Psi\left(\mu_{j}(\zeta)\right) \psi_{j}(x, \zeta) \psi_{j}^{*}\left(x^{\prime}, \zeta\right)\right) \varphi^{\prime}\left(\mu_{k}(\zeta)\right) \psi_{k}\left(x^{\prime}, \zeta\right) \psi_{k}^{*}(x, \zeta) d x^{\prime} d x d \zeta \tag{3.2.4}
\end{equation*}
$$

Since

$$
\begin{aligned}
& \operatorname{tr} \int_{\mathbb{R}^{2}} \partial_{\zeta}\left(\psi_{j}(x, \zeta) \psi_{j}^{*}\left(x^{\prime}, \zeta\right)\right) \psi_{k}\left(x^{\prime}, \zeta\right) \psi_{k}^{*}(x, \zeta) d x^{\prime} d x \\
= & \delta_{j k}\left(\int_{\mathbb{R}} \psi_{k}^{*}(x, \zeta) \partial_{\zeta} \psi_{k}(x, \zeta) d x+\int_{\mathbb{R}} \partial_{\zeta} \psi_{k}^{*}(x, \zeta) \psi_{k}(x, \zeta) d x\right) \\
= & \delta_{j k} \partial_{\zeta} \int_{\mathbb{R}} \psi_{k}^{*}(x, \zeta) \psi_{k}(x, \zeta) d x=0,
\end{aligned}
$$

the contribution to (3.2.4) from $\partial_{\zeta}\left(\psi_{j}(x, \zeta) \psi_{j}^{*}\left(x^{\prime}, \zeta\right)\right)$ vanishes, and thus

$$
\begin{aligned}
2 \pi \sigma_{I} & =\operatorname{tr} \int_{\mathbb{R}^{3}} \sum_{j, k=1}^{J} \partial_{\zeta} \Psi\left(\mu_{j}(\zeta)\right) \psi_{j}(x, \zeta) \psi_{j}^{*}\left(x^{\prime}, \zeta\right) \varphi^{\prime}\left(\mu_{k}(\zeta)\right) \psi_{k}\left(x^{\prime}, \zeta\right) \psi_{k}^{*}(x, \zeta) d x^{\prime} d x d \zeta \\
& =\int_{\mathbb{R}} \sum_{j=1}^{J} \partial_{\zeta} \Psi\left(\mu_{j}(\zeta)\right) \varphi^{\prime}\left(\mu_{j}(\zeta)\right) d \zeta=\int_{\mathbb{R}} \sum_{j=1}^{J} \partial_{\zeta} \mu_{j}(\zeta) \varphi^{\prime}\left(\mu_{j}(\zeta)\right) d \zeta=\sum_{j=1}^{J} \int_{\mathbb{R}} \partial_{\zeta} \varphi\left(\mu_{j}(\zeta)\right) d \zeta
\end{aligned}
$$

where we have used orthonormality of the eigenfunctions to justify the second equality above. It follows that

$$
2 \pi \sigma_{I}=\sum_{j=1}^{J}\left(\lim _{\zeta \rightarrow+\infty} \varphi\left(\mu_{j}(\zeta)\right)-\lim _{\zeta \rightarrow-\infty} \varphi\left(\mu_{j}(\zeta)\right)\right)
$$

Since $\varphi \in \mathfrak{S}\left(0,1 ; E_{1}, E_{2}\right)$, the above right-hand side is indeed well defined and the result is complete.

Now we want to analyze the stability of $\sigma_{I}$ with respect to perturbations. Let

$$
H_{\mu}=H+\mu W \quad \text { for } \quad \mu \in[0,1]
$$

where $W$ is a symmetric $\Psi \mathrm{DO}$. We begin with a criterion for stability of $\sigma_{I}$. It will require two preliminary results (Lemmas 3.2.7 and 3.2.8). Below, $U^{\circ}$ denotes the interior of $U$.

Theorem 3.2.6. Let $\Psi \in \mathcal{C}_{c}^{\infty}\left(E_{1}, E_{2}\right)$ such that $\varphi^{\prime} \in \mathcal{C}_{c}^{\infty}\left(\{\Psi(\lambda)=\lambda\}^{\circ}\right)$. If $H_{\mu}$ is selfadjoint and the operators $\left[H_{\mu}, P\right] \varphi^{\prime}\left(H_{\mu}\right), \varphi^{\prime}\left(H_{\mu}\right)-\varphi^{\prime}\left(H_{0}\right)$ and $\Psi\left(H_{\mu}\right)-\Psi\left(H_{0}\right)$ are traceclass, then $\sigma_{I}\left(H_{\mu}\right)=\sigma_{I}\left(H_{0}\right)$.

Proof. By assumption, $\sigma_{I}\left(H_{\mu}\right)$ is well defined. As in the proof of Theorem 3.2.5, it follows that $\sigma_{I}\left(H_{\lambda}\right)=\operatorname{Tr} i\left[\Psi\left(H_{\lambda}\right), P\right] \varphi^{\prime}\left(H_{\lambda}\right)$ for $\lambda \in\{0, \mu\}$. We can thus write the difference of
conductivities as

$$
\sigma_{I}\left(H_{\mu}\right)-\sigma_{I}\left(H_{0}\right)=\operatorname{Tr} i\left[\Psi\left(H_{\mu}\right), P\right]\left(\varphi^{\prime}\left(H_{\mu}\right)-\varphi^{\prime}\left(H_{0}\right)\right)+\operatorname{Tr} i\left[\Psi\left(H_{\mu}\right)-\Psi\left(H_{0}\right), P\right] \varphi^{\prime}\left(H_{0}\right),
$$

where our hypotheses have guaranteed that each trace is well defined. Using Lemma 3.2.7, we can replace $P$ above by $P_{y_{0}}$, where $P_{y_{0}}(y):=P\left(y-y_{0}\right)$. Again applying cyclicity (and linearity) of the trace, we get $\sigma_{I}\left(H_{\mu}\right)-\sigma_{I}\left(H_{0}\right)=\sum_{j=1}^{4} \operatorname{Tr} P_{y_{0}} A_{j}$, where the $A_{j}$ are all trace-class. The result then follows from Lemma 3.2.8.

Lemma 3.2.7. Let $P_{1}, P_{2} \in \mathfrak{S}(0,1)$ with $P_{j}=P_{j}(y)$. Then

$$
\operatorname{Tr} i\left[H, P_{1}\right] \varphi^{\prime}(H)=\operatorname{Tr} i\left[H, P_{2}\right] \varphi^{\prime}(H) .
$$

Proof. With $\Psi$ defined in Theorem 3.2.6, we have

$$
\operatorname{Tr} i\left[H, P_{2}\right] \varphi^{\prime}(H)-\operatorname{Tr} i\left[H, P_{1}\right] \varphi^{\prime}(H)=\operatorname{Tr} i\left[\Psi(H), P_{2}-P_{1}\right] \varphi^{\prime}(H)
$$

Since $P_{2}-P_{1} \in\langle x\rangle^{-\infty}$, Lemma 3.2.3 implies that $\left(P_{2}-P_{1}\right) \varphi^{\prime}(H)$ is trace-class. Therefore,

$$
\begin{aligned}
\operatorname{Tr} i\left[\Psi(H), P_{2}-P_{1}\right] \varphi^{\prime}(H) & =\operatorname{Tr} i \Psi(H)\left(P_{2}-P_{1}\right) \varphi^{\prime}(H)-\operatorname{Tr} i\left(P_{2}-P_{1}\right) \Psi(H) \varphi^{\prime}(H) \\
& =\operatorname{Tr} i\left(P_{2}-P_{1}\right) \varphi^{\prime}(H) \Psi(H)-\operatorname{Tr} i\left(P_{2}-P_{1}\right) \Psi(H) \varphi^{\prime}(H)=0,
\end{aligned}
$$

where we have used cyclicity of the trace to justify the second line, and the fact that $\left[\varphi^{\prime}(H), \Psi(H)\right]=0$ for the last line.

Lemma 3.2.8. Let $P(y)=P \in \mathfrak{S}(0,1)$ and $y_{0} \in \mathbb{R}$, and define $P_{y_{0}}(y):=P\left(y-y_{0}\right)$. Then for any trace-class operator $A$ on $L^{2}\left(\mathbb{R}^{2}\right) \otimes \mathbb{C}^{n}$, we have $\operatorname{Tr} P_{y_{0}} A \rightarrow 0$ as $y_{0} \rightarrow \infty$.

Proof. Writing $A=\frac{1}{2}\left(A+A^{*}\right)+\frac{1}{2}\left(A-A^{*}\right)$, with $\frac{1}{2}\left(A+A^{*}\right)$ and $\frac{i}{2}\left(A-A^{*}\right)$ trace-class and self-adjoint and using the triangle inequality, we may assume that $A$ is self-adjoint. Fix $\varepsilon>0$.

By the spectral theorem, there exists an orthonormal basis $\left\{\psi_{j}\right\}_{j=1}^{\infty}$ of $L^{2}\left(\mathbb{R}^{2}\right) \otimes \mathbb{C}^{n}$ such that $A \psi_{j}=\lambda_{j} \psi_{j}$, with $\left\{\lambda_{j}\right\}_{j=1}^{\infty} \subset \mathbb{R}$ satisfying $\sum_{j=1}^{\infty}\left|\lambda_{j}\right|<\infty$. Thus there exists $N \in \mathbb{N}$ such that $\sum_{j=N}^{\infty}\left|\left(\psi_{j}, P_{y_{0}} A \psi_{j}\right)\right| \leq \sum_{j=N}^{\infty}\left|\lambda_{j}\right|<\varepsilon / 2$ for all $y_{0} \in \mathbb{R}$. Since $\left\{\psi_{j}\right\}_{j=1}^{\infty} \subset L^{2}\left(\mathbb{R}^{2}\right) \otimes \mathbb{C}^{n}$, we know that, for $y_{0}$ sufficiently large, $\left\|P_{y_{0}} \psi_{j}\right\|<\varepsilon /(2 N\|A\|)$ for all $j \in\{1,2, \ldots, N-1\}$. It follows that $\left|\operatorname{Tr} P_{y_{0}} A\right| \leq \sum_{j=1}^{\infty}\left|\left(\psi_{j}, P_{y_{0}} A \psi_{j}\right)\right|<\varepsilon$ for all $y_{0}$ sufficiently large.

We now use Theorem 3.2.6 to prove stability of $\sigma_{I}$ under a large class of perturbations $W$. For the rest of this section, let $\Phi_{0} \in \mathcal{C}_{c}^{\infty}\left(E_{1}, E_{2}\right)$ such that $\Psi \in \mathcal{C}_{c}^{\infty}\left(\left\{\Phi_{0}=1\right\}^{\circ}\right)$, with $\Psi$ as in Theorem 3.2.6.

Theorem 3.2.9. Let $W$ be a symmetric $\Psi D O$, such that $\left(i \pm H_{0}\right)^{-1} W$ is bounded and $\Phi_{0}\left(H_{0}\right) W$ is trace-class. Then $\sigma_{I}\left(H_{\mu}\right)=\sigma_{I}\left(H_{0}\right)$ for all $\mu>0$ sufficiently small.

Proof. We verify the assumptions of Theorem 3.2.6. Since $\left(i \pm H_{0}\right)^{-1} W$ is bounded, $i \pm H_{\mu}$ : $\left(i-H_{0}\right)^{-1} \mathcal{H} \rightarrow \mathcal{H}$ is bijective whenever $\mu<\left\|\left(i \pm H_{0}\right)^{-1} W\right\|^{-1}$, with $\left(i \pm H_{\mu}\right)^{-1}=(1 \pm$ $\left.\mu\left(i \pm H_{0}\right)^{-1} W\right)^{-1}\left(i \pm H_{0}\right)^{-1}$. Hence for all $\mu>0$ sufficiently small, $H_{\mu}$ is self-adjoint with the same domain of definition $\mathcal{D}\left(H_{\mu}\right)=\mathcal{D}\left(H_{0}\right)=\left(i-H_{0}\right)^{-1} \mathcal{H}$.

It remains to verify that $\left[H_{\mu}, P\right] \varphi^{\prime}\left(H_{\mu}\right)$ and $\Phi\left(H_{\mu}\right)-\Phi\left(H_{0}\right)$ are trace-class for $\Phi \in$ $\left\{\varphi^{\prime}, \Psi\right\}$. We start with the difference

$$
\begin{aligned}
\Phi\left(H_{\mu}\right)-\Phi\left(H_{0}\right)=\left(\Phi\left(H_{\mu}\right)\right. & \left.-\Phi\left(H_{0}\right)\right)\left(\Phi_{0}\left(H_{\mu}\right)-\Phi_{0}\left(H_{0}\right)\right) \\
& +\Phi\left(H_{0}\right)\left(\Phi_{0}\left(H_{\mu}\right)-\Phi_{0}\left(H_{0}\right)\right)+\left(\Phi\left(H_{\mu}\right)-\Phi\left(H_{0}\right)\right) \Phi_{0}\left(H_{0}\right)
\end{aligned}
$$

which can be rearranged to give

$$
\begin{align*}
\left(\Phi\left(H_{\mu}\right)-\Phi\left(H_{0}\right)\right)(1 & \left.-\left(\Phi_{0}\left(H_{\mu}\right)-\Phi_{0}\left(H_{0}\right)\right)\right)  \tag{3.2.5}\\
& =\Phi\left(H_{0}\right)\left(\Phi_{0}\left(H_{\mu}\right)-\Phi_{0}\left(H_{0}\right)\right)+\left(\Phi\left(H_{\mu}\right)-\Phi\left(H_{0}\right)\right) \Phi_{0}\left(H_{0}\right)
\end{align*}
$$

By the Helffer-Sjöstrand formula,

$$
\begin{equation*}
\Phi_{0}\left(H_{\mu}\right)-\Phi_{0}\left(H_{0}\right)=\frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\Phi}_{0}(z)\left(z-H_{0}\right)^{-1} \mu W\left(z-H_{\mu}\right)^{-1} d^{2} z \tag{3.2.6}
\end{equation*}
$$

Recall that $\tilde{\Phi}_{0} \in \mathcal{C}_{c}^{\infty}(\mathbb{C})$ with $\left|\bar{\partial} \tilde{\Phi}_{0}(z)\right| \leq C|\Im z|^{2}$, and write

$$
\left(z-H_{0}\right)^{-1}=\left(1+(i-z)\left(z-H_{0}\right)^{-1}\right)\left(i-H_{0}\right)^{-1}
$$

Since $\left\|\left(z-H_{\mu}\right)^{-1}\right\| \leq|\Im z|^{-1}$, we see that the norm of $\Phi_{0}\left(H_{\mu}\right)-\Phi_{0}\left(H_{0}\right)$ is bounded by $C \mu$, meaning that it is less than 1 for all $\mu$ small enough. It thus suffices to show that each term on the right-hand side of (3.2.5) is trace-class, as $1-\left(\Phi_{0}\left(H_{\mu}\right)-\Phi_{0}\left(H_{0}\right)\right)$ has bounded inverse. Using (3.2.6), the first term becomes

$$
\begin{equation*}
\Phi\left(H_{0}\right)\left(\Phi_{0}\left(H_{\mu}\right)-\Phi_{0}\left(H_{0}\right)\right)=\frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\Phi}_{0}(z)\left(z-H_{0}\right)^{-1} \mu \Phi\left(H_{0}\right) W\left(z-H_{\mu}\right)^{-1} d^{2} z \tag{3.2.7}
\end{equation*}
$$

By our assumption that $\Phi\left(H_{0}\right) W$ is trace-class (and again using that all singularities of the resolvent operators are compensated by the decay in $\bar{\partial} \tilde{\Phi}_{0}$ ), we conclude that (3.2.7) is trace-class. Applying the same argument to the second term

$$
\begin{equation*}
\left(\Phi\left(H_{\mu}\right)-\Phi\left(H_{0}\right)\right) \Phi_{0}\left(H_{0}\right)=-\frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\Phi}(z)\left(z-H_{\mu}\right)^{-1} \mu W \Phi_{0}\left(H_{0}\right)\left(z-H_{0}\right)^{-1} d^{2} z \tag{3.2.8}
\end{equation*}
$$

it follows that $\Phi\left(H_{\mu}\right)-\Phi\left(H_{0}\right)$ is trace-class.
For the first operator, we write

$$
\begin{equation*}
\left[H_{\mu}, P\right] \Phi\left(H_{\mu}\right)=\left[H_{\mu}, P\right] \Phi\left(H_{0}\right)+\left[H_{\mu}, P\right]\left(\Phi\left(H_{\mu}\right)-\Phi\left(H_{0}\right)\right) \tag{3.2.9}
\end{equation*}
$$

with the first term on the above right-hand side trace-class by Lemma 3.2.3 and the composition calculus. To show that the second term is also trace-class, we again use (3.2.5).

Namely, it suffices to show that

$$
T_{1}:=\left[H_{0}, P\right] A, \quad T_{2}:=[W, P] A, \quad T_{3}:=\left[H_{0}, P\right] B, \quad T_{4}:=[W, P] B
$$

are trace-class, with $A:=\Phi\left(H_{0}\right)\left(\Phi_{0}\left(H_{\mu}\right)-\Phi_{0}\left(H_{0}\right)\right)$ and $B:=\left(\Phi\left(H_{\mu}\right)-\Phi\left(H_{0}\right)\right) \Phi_{0}\left(H_{0}\right)$. That $T_{1}$ is trace-class follows immediately from (3.2.7) and boundedness of $\left[H_{0}, P\right]=-i P^{\prime}(y) \sigma_{2}$. We know that $P W A$ is trace-class (since $\left(i \pm H_{0}\right)^{-1} W$ is bounded; this is an assumption on $W)$, thus $T_{2}$ is trace-class if $W P A$ is. We write

$$
\begin{aligned}
W P A= & \frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\Phi}_{0}(z) W P\left(z-H_{0}\right)^{-1} \mu \Phi\left(H_{0}\right) W\left(z-H_{\mu}\right)^{-1} d^{2} z \\
= & \frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\Phi}_{0}(z) W\left(z-H_{0}\right)^{-1} P \mu \Phi\left(H_{0}\right) W\left(z-H_{\mu}\right)^{-1} d^{2} z \\
& +\frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\Phi}_{0}(z) W\left[P,\left(z-H_{0}\right)^{-1}\right] \mu \Phi\left(H_{0}\right) W\left(z-H_{\mu}\right)^{-1} d^{2} z
\end{aligned}
$$

Since $\left[P,\left(z-H_{0}\right)^{-1}\right]=-\left(z-H_{0}\right)^{-1}\left[H_{0}, P\right]\left(z-H_{0}\right)^{-1}$, boundedness of $\left[H_{0}, P\right]$ and our assumption on $W$ imply that $W P A$ is trace-class. Using (3.2.8) and the identity

$$
\begin{aligned}
\left(z-H_{\mu}\right)^{-1} & =\left(i-H_{\mu}\right)^{-1}\left(1+(i-z)\left(z-H_{\mu}\right)^{-1}\right) \\
& =\left(i-H_{0}\right)^{-1}\left(1-\mu W\left(i-H_{0}\right)^{-1}\right)^{-1}\left(1+(i-z)\left(z-H_{\mu}\right)^{-1}\right),
\end{aligned}
$$

we similarly conclude that $T_{3}$ and $T_{4}$ are trace-class. This completes the result.

We now prove a stability result that does not require the perturbation to be "small". To do this, we will need the stronger assumption that $\left(i \pm H_{0}\right)^{-1} W$ is compact.

Theorem 3.2.10. Let $W$ be a symmetric $\Psi D O$. Assume that $\left(i \pm H_{0}\right)^{-1} W$ is compact and $\Phi_{0}\left(H_{0}\right) W$ is trace-class. Then $\sigma_{I}\left(H_{1}\right)=\sigma_{I}\left(H_{0}\right)$.

Proof. We again verify the hypotheses of Theorem 3.2.6. The fact that $\left(i \pm H_{0}\right)^{-1} W$ is compact implies that $H_{1}=H_{0}+W$ is self-adjoint with domain of definition $\mathcal{D}\left(H_{1}\right)=\mathcal{D}\left(H_{0}\right)$.

Indeed, the fact that $W$ is symmetric implies that the kernel of $i+H_{1}$ is trivial. The Fredholm alternative and our compactness assumption imply that the dimension of the kernel of $1+(i+$ $\left.H_{0}\right)^{-1} W$ is equal to the codimension of its range. But $1+\left(i+H_{0}\right)^{-1} W=\left(i+H_{0}\right)^{-1}\left(i+H_{1}\right)$ with $\left(i+H_{0}\right)^{-1}: \mathcal{H} \rightarrow \mathcal{D}\left(H_{0}\right)$ a bijection, and thus codim $\operatorname{Ran}\left(i+H_{1}\right)=\operatorname{dim} \operatorname{ker}\left(i+H_{1}\right)=0$. The same argument also shows that codim $\operatorname{Ran}\left(i-H_{1}\right)=\operatorname{dim} \operatorname{ker}\left(i-H_{1}\right)=0$. We conclude by [70, Theorem VIII.3] that $H_{1}$ is self-adjoint.

We now prove the necessary trace-class properties. Let $\Phi \in\left\{\varphi^{\prime}, \Psi\right\}$. By (3.2.5), (3.2.7) and (3.2.8), we have that $\Theta\left(1-\Theta_{0}\right)$ is trace-class, where $\Theta:=\Phi\left(H_{1}\right)-\Phi\left(H_{0}\right)$ and $\Theta_{0}:=$ $\Phi_{0}\left(H_{1}\right)-\Phi_{0}\left(H_{0}\right)$. We know that $\Theta_{0}$ is compact by (3.2.6) and our assumption that $W$ is relatively compact with respect to $H_{0}$. Thus there exist $\Theta_{00}$ and $\Theta_{01}$ such that $\Theta_{00}$ has finite rank, $\left\|\Theta_{01}\right\|<1$, and $\Theta_{0}=\Theta_{00}+\Theta_{01}$. It follows that $\Theta\left(1-\Theta_{01}\right)=\Theta\left(1-\Theta_{0}\right)+\Theta \Theta_{00}$ is trace-class. Applying $\left(1-\Theta_{01}\right)^{-1}$ to both sides (on the right), we conclude that $\Theta$ is trace-class.

We now show that $\left[H_{1}, P\right] \Phi\left(H_{1}\right)$ is trace-class. Using (3.2.9), it suffices to show that $\left[H_{1}, P\right] \Theta$ is trace-class. The proof of Theorem 3.2.9 shows that $\left[H_{1}, P\right] \Theta\left(1-\Theta_{0}\right)$ is traceclass. As above, we write $\left[H_{1}, P\right] \Theta=\left(T+\left[H_{1}, P\right] \Theta \Theta_{00}\right)\left(1-\Theta_{01}\right)^{-1}$ with $T$ and $\left[H_{1}, P\right] \Theta \Theta_{00}$ trace-class (the latter because $\Theta_{00}$ has finite rank) and $\left(1-\Theta_{01}\right)^{-1}$ bounded. This completes the result.

Using Lemmas 3.2.2 and 3.2.3, we see that for any $p>0$,

$$
W \in \mathrm{Op}\left(S\left(\left\langle\xi, \zeta-A_{2}(x)\right\rangle\langle y\rangle^{-1-p}\right)\right)
$$

satisfies the assumptions of Theorem 3.2.9 and $W \in \operatorname{Op}\left(S\left(\left\langle\xi, \zeta-A_{2}(x)\right\rangle^{1-p}\langle x\rangle^{-p}\langle y\rangle^{-1-p}\right)\right)$ satisfies the assumptions of Theorem 3.2.10. Note that the assumption that $\Phi_{0}\left(H_{0}\right) W$ is trace-class forces the symbol of $W$ to decay in $y$, as the symbol of $H_{0}$ is independent of this variable. We now introduce two stability results that no longer require this decay in $y$.

Theorem 3.2.11. Let $W=W(x, y)$ be a symmetric point-wise multiplication operator in $S(1)$ that is compactly supported in $x$. Then $\sigma_{I}(H+W)=\sigma_{I}(H)$.

Proof. The same proofs of Lemmas 3.2.2 and 3.2.3 easily imply that these results hold with $H$ replaced by $H+W$. The fact that $W$ is bounded implies that $H+W$ is self-adjoint. Hence $H^{(\mu)}:=H+\mu W$ satisfies the assumptions of Corollary 2.4.5 for all $\mu \in[0,1]$ and the result follows.

Our next result trades the trace-class assumption of $\Phi_{0}\left(H_{0}\right) W$ in Theorem 2.1.7 for a smoothness property of $W$.

Theorem 3.2.12. Let $W=\operatorname{Op}(w)$ be a symmetric $\Psi D O$ satisfying

$$
w \in S\left(\left\langle\xi, \zeta-A_{2}(x)\right\rangle^{1-\delta}\langle x, y\rangle^{-\delta}\right)
$$

for some $\delta>0$, and $\partial_{\alpha} w \in S(1)$ for all $|\alpha| \geq 1$. Then $\sigma_{I}(H+W)=\sigma_{I}(H)$.
Proof. Since $W$ is relatively compact with respect to $H$ with the latter self-adjoint, we know that $H^{(\mu)}:=H+\mu W$ is self-adjoint for all $\mu \in[0,1]$. Following the proof of Lemma 3.2.2 and using the boundedness of $\partial_{\alpha} w$, we verify that $H^{(\mu)} \in \operatorname{Op}(S(\langle\xi, \zeta-$ $\left.\left.A_{2}(x)\right\rangle^{-1}\right)$ ) for all $\mu$. Thus (3.2.5) implies that $\Theta^{(\mu)}\left(1-\Theta_{0}^{(\mu)}\right) \in \operatorname{Op}\left(S(x, \xi, \zeta)^{-\infty}\right)$, where $\Theta^{(\mu)}:=\Phi\left(H^{(\mu)}\right)-\Phi(H)$ and $\Theta_{0}^{(\mu)}:=\Phi_{0}\left(H^{(\mu)}\right)-\Phi_{0}(H)$. The Helffer-Sjöstrand formula implies that $\Theta_{0}^{(\mu)} \in \operatorname{Op}\left(S\left(\langle x, y, \xi, \zeta\rangle^{-\delta}\right)\right)$, and therefore $\Theta_{0}^{(\mu)}=\Theta_{00}^{(\mu)}+\Theta_{01}^{(\mu)}$, where $\Theta_{00}^{(\mu)} \in \operatorname{Op}\left(S\left(\langle x, y, \xi, \zeta\rangle^{-\infty}\right)\right)$ and $\Theta_{01}^{(\mu)}$ has symbol in $S(1)$ as small as necessary. We conclude that $\Theta^{(\mu)}=\left(T+\Theta^{(\mu)} \Theta_{00}^{(\mu)}\right)\left(1-\Theta_{01}^{(\mu)}\right)^{-1}$ with $T \in \operatorname{Op}\left(S(x, \xi, \zeta)^{-\infty}\right)$, $\Theta^{(\mu)} \Theta_{00}^{(\mu)} \in$ $\operatorname{Op}\left(S\left(\langle x, y, \xi, \zeta\rangle^{-\infty}\right)\right)$ and $\left(1-\Theta_{01}^{(\mu)}\right)^{-1} \in \operatorname{Op}(S(1))$. It follows from the composition calculus that $\Phi\left(H^{(\mu)}\right)=\Phi(H)+\Theta^{(\mu)} \in \operatorname{Op}\left(S\left(\langle x, \xi, \zeta\rangle^{-\infty}\right)\right)$. We have shown that $H^{(\mu)}$ satisfies the assumptions of Corollary 2.4.5 for all $\mu \in[0,1]$, and the proof is complete.

We have provided four stability results, each of which has a different assumption on the perturbation $W$. We now compare these assumptions using four illustrative examples.

Example 1. Suppose $W=W(y)=\eta\langle y\rangle^{-1-\delta}$ for some $\delta>0$, where $\eta \in\{1,-1\}$. Then Lemma 3.2.3 and the composition calculus imply that for any $p>0, \Phi_{0}\left(H_{0}\right) W \in$ $\operatorname{Op}\left(S\left(\langle y\rangle^{-1-\delta}\langle x, \xi, \zeta\rangle^{-p}\right)\right)$. It follows that the symbol of $\Phi_{0}\left(H_{0}\right) W$ is integrable, meaning that $\Phi_{0}\left(H_{0}\right) W$ is trace-class. Since $W$ is bounded, we see that $W$ satisfies the assumptions of Theorem 3.2.9. This means $\sigma_{I}(H+\mu W)=\sigma_{I}(H)$ for all $\mu>0$ sufficiently small. But note that $\left(i+H_{0}\right)^{-1} W$ is not compact, as its symbol does not decay in $\langle x, \zeta\rangle$. Thus Theorem 3.2.10 does not apply. It is also clear that Theorem 3.2.11 does not apply since $W$ is independent of $x$. Theorem 3.2.12 does not apply since the symbol of $(i+H)^{-1} W$ does not decay in $\langle x, \zeta\rangle$.

Example 2. Suppose $W=W(x, y)=a\langle x, y\rangle^{-2-\delta} \operatorname{sgn}(y)$ for some constants $a \in \mathbb{R}$ and $\delta>0$. Define $W_{0}(x, y):=a\langle x, y\rangle^{-2-\delta}$. As above it follows that $\Phi_{0}\left(H_{0}\right) W_{0}$ is traceclass. Since $y \mapsto \operatorname{sgn}(y)$ is bounded, we conclude that $\Phi_{0}\left(H_{0}\right) W$ is trace-class. Moreover, since $W_{0}$ decays in both $x$ and $y$, the composition calculus implies that $\left(i \pm H_{0}\right)^{-1} W_{0} \in$ $\operatorname{Op}\left(S\left(\langle x, y, \xi, \zeta\rangle^{-2}\right)\right)$, meaning that $\left(i \pm H_{0}\right)^{-1} W_{0}$ is compact. Thus $\left(i \pm H_{0}\right)^{-1} W$ is compact, meaning that $W$ satisfies the assumptions of Theorem 3.2.10. Although $W$ also satisfies the assumptions of Theorem 3.2.9, the latter cannot be used to prove $\sigma_{I}(H+W)=\sigma_{I}(H)$ when $a$ is sufficiently large. Again it is clear that Theorem 3.2 .11 does not apply, as $W$ is not compactly supported in $x$. Since $W$ is not continuous, Theorem 3.2.12 does not apply either.

Example 3. Suppose $W(x)=W \in \mathcal{C}_{c}^{\infty}(\mathbb{R})$, so that Theorem 3.2.11 applies. Then the symbol of $\Phi_{0}\left(H_{0}\right) W$ does not decay in $y$, meaning that Theorem 3.2.12 does not apply and that $\Phi_{0}\left(H_{0}\right) W$ is not trace-class. Thus $W$ does not satisfy the assumptions of Theorem 3.2.9 or 3.2.10.

Example 4. Suppose $W=W(x, y)=\langle x, y\rangle^{-\delta}$ for some $0<\delta<1$. It is clear that $W$ satisfies the assumptions of Theorem 3.2.12, but does not decay fast enough to meet the criteria of Theorems 3.2.9-3.2.11.

## CHAPTER 4

## PERIODIC SYSTEMS

This section concerns computational approximations of $\sigma_{I}(H, P)$. A standard method to approximate spectral decompositions of operators in $\mathbb{R}^{2}$ is to consider the restriction of the operator on a box $(-\pi L, \pi L)^{2}$ with periodic boundary conditions and analyze the limit $L \rightarrow \infty$. However, because of the structure of the domain wall with $\sigma(y)$ transitioning from $\sigma_{-}$to $\sigma_{+}$as $y$ crosses 0 , any periodization generates another transition from $\sigma_{+}$to $\sigma_{-}$as $y$ crosses $\pi L$. The asymmetric transport along the $x$ axis when $y$ is close to 0 is compensated by an asymmetric transport along the $x$ axis when $y$ is close to $-\pi L \equiv \pi L$ with opposite chirality, resulting in a (globally) topologically trivial material.

Periodic systems thus no longer enjoy a non-trivial topology. This is a no-go result similar to a fermion doubling or Nelson-Ninomiya theorem [87] ensuring that any domain wall in a mass term on a torus, no matter how large, may continuously be deformed to a constant mass term. The trace in (1.0.1) therefore needs to be modified so the integral focuses on the original domain wall, which as $L \rightarrow \infty$ becomes well separated from the spurious second domain wall. While conductivities may be computed for general pseudodifferential Hamiltonians, as we do in Section 2, see also [11], we restrict our analysis of periodized Hamiltonians to differential systems and avoid complications resulting from nonlocal effects. We therefore consider (unperturbed) infinite-space Hamiltonians that satisfy (H1) and are differential operators of the form

$$
\begin{equation*}
H=M_{0} D_{y}^{m}+M_{m} D_{x}^{m}+\sum_{j=0}^{m} a_{j}(y) D_{x}^{j} D_{y}^{m-j}+\sum_{i+j \leq m-1} a_{i j}(y) D_{x}^{i} D_{y}^{j} \tag{4.0.1}
\end{equation*}
$$

where $D_{x}=-i \partial_{x}$ and $D_{y}=-i \partial_{y}$ and $M_{0}$ and $M_{m}$ are constant nonsingular Hermitian
matrices, while

$$
\left\{M_{0}, M_{m}\right\}\left\{\begin{array}{ll}
=0, & m \text { is odd }  \tag{4.0.2}\\
\geq 0, & m \text { is even }
\end{array}, \quad\left\{M_{i}, a_{j}(y)\right\} \begin{cases}=0, & i+j \text { is odd } \\
\geq 0, & i+j \text { is even }\end{cases}\right.
$$

with the second condition understood to hold for all $i \in\{0, m\}, j \in\{0,1, \ldots, m\}$, and $y \in \mathbb{R}$. The $a_{j}(y)$ and $a_{i j}(y)$ are smooth matrix-valued functions such that $H$ is symmetric. In particular, this means the leading-order coefficients $a_{j}(y)$ are all Hermitian-valued. The anti-commutation relations (4.0.2) with $\left\{O_{1}, O_{2}\right\}:=O_{1} O_{2}+O_{2} O_{1}$ for two operators $O_{1}$ and $O_{2}$ ensure that $H^{2}=M_{0}^{2} D_{y}^{2 m}+M_{m}^{2} D_{x}^{2 m}+A+B$, where $A$ is non-negative and $B$ is a differential operator of order $2 m-1$, so that $H^{2}$ remains a self-adjoint elliptic operator. We need the above structure in some applications to superconductors. We will finally consider perturbed operators $H_{V}=H+V$ where $V=V(x, y)$ is a multiplication operator with support that remains sufficiently close, in an appropriate sense, to the center of the torus.

To approximate $H$ by an operator on $(-\pi L, \pi L)^{2}$, we redefine the coefficients near $y=\pi L$ so that they are smoothly connected there. The resulting periodic Hamiltonian is unitarily equivalent to an operator $H_{\lambda}$ on $(-\pi, \pi)^{2}$, with $\lambda=L^{-1}$ the relevant parameter. The details of the construction of $H_{\lambda}$ are left for section 4.1. For $\varphi \in \mathfrak{S}\left(0,1 ; E_{1}, E_{2}\right)$, we define the periodic interface conductivity by

$$
\begin{equation*}
\tilde{\sigma}_{I}\left(H_{\lambda}\right)=\operatorname{Tr} i Q\left[H_{\lambda}, P\right] \varphi^{\prime}\left(H_{\lambda}\right) \tag{4.0.3}
\end{equation*}
$$

where $P=P(x)$ and $Q=Q(x, y)=Q_{X}(x) Q_{Y}(y)$ are smooth point-wise multiplication operators such that

$$
Q_{\Theta}(\theta)= \begin{cases}1, & |\theta| \leq \pi / 2 \\ 0, & |\theta| \geq \pi / 2+\delta_{Q}\end{cases}
$$

and

$$
P(x)= \begin{cases}0, & x_{1} \leq x \leq x_{2} \\ 1, & -\pi<x_{1}<-3 \pi / 4<-\pi / 2<x_{2} \\ x_{3} \leq x \leq x_{4} & x_{2}<x_{3}<\pi / 2<3 \pi / 4<x_{4}<\pi\end{cases}
$$

for some fixed $0<\delta_{Q}<\pi / 4$. As we may observe, $Q$ is a spatial filter centered at the domain wall of interest where $P^{\prime}>0$ and vanishing in the vicinity of the unwanted domain wall where $P^{\prime}<0$.

Our main result in Section 4.1 shows that $\left|\tilde{\sigma}_{I}\left(H_{\lambda}\right)-\sigma_{I}(H)\right| \lesssim \lambda^{p}$ for any $p \geq 0$ as $\lambda \rightarrow 0$, which corresponds to an almost-exponential rate of convergence. This is based on a careful approximation of the periodic eigenfunctions by truncations of the infinite eigenfunctions corresponding to energies in the bulk spectral gap in the vicinity of domain walls and on a proof that such periodic eigenfunctions have to be negligible away from the domain walls. Estimates of the periodic eigenelements by means of min-max theorems allow us to obtain the above trace estimates.

We next prove in Section 4.2 that for a large class of perturbations $V_{\lambda}$ with appropriate support constraints, $\left|\tilde{\sigma}_{I}\left(H_{\lambda}+V_{\lambda}\right)-\tilde{\sigma}_{I}\left(H_{\lambda}\right)\right| \lesssim \lambda^{p}$ as $\lambda \rightarrow 0$, for all $p \geq 0$. These results of approximate stability of conductivities in a periodic setting parallel the exact stability results in Euclidean space of Section 2.

Section 4.3 finally shows how the above theory applies to the models from solid state physics that were already considered in Section 2.3.

Section 4.4 presents several numerical simulations highlighting our theoretical findings. In particular, we demonstrate for several differential operators that $\tilde{\sigma}_{I}$ is a good numerical approximation of the infinite-space conductivity $\sigma_{I}$ and is stable under perturbations by sufficiently localized point-wise multiplication operators.

We also confirm numerically several theoretical results obtained for a topologically nontrivial $3 \times 3$ system of water wave equations. In particular, it is known that the bulk-interface
correspondence may fail for certain domain wall profiles of the un-regularized system. We confirm numerically results derived in [10] and showing that the conductivity is stable for some $3 \times 3$ perturbations $V$ but not all. We also confirm results obtained in earlier sections that a regularized version of the water wave system does have a stable, well defined, edge conductivity equal to 2 .

### 4.1 Periodic approximations of infinite-space problems

This section addresses approximations of a Hamiltonian in $\mathbb{R}^{2}$ by an operator on the 2-torus $\mathbb{T}^{2}$. Throughout this thesis, $\mathbb{T}$ is identified with $[-\pi, \pi)$. We construct the periodic operator $H_{\lambda}$ introduced above and show that the periodic interface conductivity (4.0.3) is a good approximation of its infinite-space analogue (1.0.1).

Let us introduce the following Hilbert spaces

$$
\mathcal{H}^{j}(X):=\left\{\Psi \in L^{2}(X) \otimes \mathbb{C}^{n} \quad\left|\quad \partial_{\alpha} \Psi \in L^{2}(X) \otimes \mathbb{C}^{n} \quad \forall\right| \alpha \mid \leq j\right\}
$$

where $j \in \mathbb{N}$ and $X \in\left\{\mathbb{R}^{d}, \mathbb{T}^{d}\right\}$ with $d \in\{1,2\}$. For $X=\mathbb{T}^{d}$ and a parameter $\lambda \in(0,1]$, we define on $\mathcal{H}^{j}(X)$ the inner products and norms

$$
\langle f, g\rangle_{\lambda, j}:=\sum_{|\alpha| \leq j}\left\langle\lambda^{|\alpha|} D^{\alpha} f, \lambda^{|\alpha|} D^{\alpha} g\right\rangle, \quad\|f\|_{\lambda, j}:=\langle f, f\rangle_{\lambda, j}^{1 / 2}
$$

with $\langle\cdot, \cdot\rangle$ the standard inner product in $L^{2}\left(\mathbb{T}^{d}\right) \otimes \mathbb{C}^{n}$. For $X=\mathbb{R}^{d}$, the corresponding inner product and norm are

$$
(f, g)_{j}:=\sum_{|\alpha| \leq j}\left(D^{\alpha} f, D^{\alpha} g\right), \quad\| \| f \|_{j}:=(f, f)_{j}^{1 / 2}
$$

with $(\cdot, \cdot)$ the standard inner product in $L^{2}\left(\mathbb{R}^{d}\right) \otimes \mathbb{C}^{n}$. We will use the shorthand

$$
\mathcal{H}(X):=\mathcal{H}^{0}(X), \quad \mathcal{H}:=\mathcal{H}\left(\mathbb{R}^{2}\right), \quad\langle\cdot, \cdot\rangle_{j}:=\langle\cdot, \cdot\rangle_{1, j}, \quad\langle\cdot, \cdot\rangle:=\langle\cdot, \cdot\rangle_{0}, \quad(\cdot, \cdot):=(\cdot, \cdot)_{0}
$$

We will repeatedly need to rescale functions with $\lambda$, embed functions on $\mathbb{R}$ in the torus, and extend functions on $\mathbb{T}$ as functions on $\mathbb{R}$. We define these operations as follows. For $\lambda \in(0,1]$ and $y_{1} \in \mathbb{R}$, define the unitary map $\Lambda_{\lambda, y_{1}}: \mathcal{H}(\mathbb{R}) \rightarrow \mathcal{H}(\mathbb{R})$ by

$$
\left(\Lambda_{\lambda, y_{1}}(u)\right)(y):=\lambda^{1 / 2} u\left(\lambda\left(y-y_{1}\right)+y_{1}\right), \quad y \in \mathbb{R}
$$

Observe that $\Lambda_{\lambda, y_{1}}^{-1}=\Lambda_{\lambda^{-1}, y_{1}}$.
Let $L_{c, 2 \pi}^{\infty}(\mathbb{R})$ denote the space of bounded functions $f \in L_{c}^{\infty}(\mathbb{R})$ such that $\operatorname{supp}(f) \in$ $(\tau, \tau+2 \pi)$ for $\tau \in \mathbb{R}$. We define by $f_{\sharp}(y)=\sum_{q \in \mathbb{Z}} f(2 \pi q+y)$ their periodization (which is smooth when $f$ is smooth) and then $\mathcal{P}: L_{c, 2 \pi}^{\infty}(\mathbb{R}) \rightarrow L^{\infty}(\mathbb{T})$ by

$$
\mathcal{P} u(y)=u_{\sharp}(y)=\sum_{q \in \mathbb{Z}} u(2 \pi q+y) .
$$

Thus, $\mathcal{P}$ is an embedding of functions on $\mathbb{R}$ with sufficiently small support to functions on $\mathbb{T}$.

We define $\tilde{\mathcal{P}}$ as the operator mapping a function in $u \in L^{\infty}(\mathbb{T})$ to $L_{c}^{\infty}(\mathbb{R})$ by

$$
\tilde{\mathcal{P}} u(y)=\chi_{[-\pi, \pi)}(y) u((y+\pi) \bmod 2 \pi-\pi)
$$

with $\chi_{I}$ the indicatrix function of $I \subset \mathbb{R}$.
For $u$ vector-valued, $\mathcal{P} u$ and $\tilde{\mathcal{P}} u$ are defined as above component-wise.

Construction of the periodic operator. We now construct a differential operator on the torus. We start from our original operator $H=\operatorname{Op}(\sigma)$, assumed to be a differential
operator on $\mathcal{H}=L^{2}\left(\mathbb{R}^{2}\right) \otimes \mathbb{C}^{n}$ given by (4.0.1) and (4.0.2) such that $\sigma$ satisfies (H1).
Any operator with smooth coefficients on the torus will have a transition modeling a domain wall between $\sigma_{-}$and $\sigma_{+}$in the vicinity of $y=0$ and another domain wall from $\sigma_{+}$ to $\sigma_{-}$in the vicinity of $y=\pi$. We construct a differential operator $H^{\prime}=\mathrm{Op}\left(\sigma^{\prime}\right)$ with such a domain wall, i.e., such that $\sigma^{\prime}=\sigma_{+}$whenever $y \leq \pi-y_{0}$ and $\sigma^{\prime}=\sigma_{-}$whenever $y \geq \pi+y_{0}$. Assume that $\sigma^{\prime}$ is smooth and independent of $x$, and that for every $y^{\prime} \in \mathbb{R}$, there exists $y \in \mathbb{R}$ such that $\sigma^{\prime}\left(y^{\prime}, \xi, \zeta\right)=\sigma(y, \xi, \zeta)$. We may for instance choose $\sigma^{\prime}(y, \xi, \zeta)=\sigma(\pi-y, \xi, \zeta)$.

The above operators are defined on $\mathbb{R}$ and need to be mapped to a torus and glued together. Define the rescaled versions $\tilde{H}_{\lambda}=\operatorname{Op}\left(\tilde{\sigma}_{\lambda}\right)$ and $\tilde{H}_{\lambda}^{\prime}=\operatorname{Op}\left(\tilde{\sigma}_{\lambda}^{\prime}\right)$, where $\tilde{\sigma}_{\lambda}(y, \xi, \zeta):=$ $\sigma(y / \lambda, \lambda \xi, \lambda \zeta)$ and $\tilde{\sigma}_{\lambda}^{\prime}(y, \xi, \zeta):=\sigma^{\prime}\left(\frac{y-\pi}{\lambda}+\pi, \lambda \xi, \lambda \zeta\right)$. Then there exists $\lambda_{0} \in(0,1]$ such that for all $\lambda \in\left(0, \lambda_{0}\right]$, we can define $H_{\lambda}$ a differential operator on $\mathcal{H}\left(\mathbb{T}^{2}\right)=L^{2}\left(\mathbb{T}^{2}\right) \otimes \mathbb{C}^{n}$ with smooth coefficients that are equal to those of $\tilde{H}_{\lambda}$ when $-\pi / 2 \leq y \leq \pi / 2$ and $\tilde{H}_{\lambda}^{\prime}$ when $\pi / 2 \leq y \leq 3 \pi / 2$ (with the equivalence $-\pi / 2 \equiv 3 \pi / 2$ on $\mathbb{T}$ ). Thus we have

$$
\begin{equation*}
H_{\lambda}=\lambda^{m}\left(M_{0} D_{y}^{m}+M_{m} D_{x}^{m}+\sum_{j=0}^{m} a_{\lambda, j}(y) D_{x}^{j} D_{y}^{m-j}\right)+\sum_{i+j \leq m-1} \lambda^{i+j} a_{\lambda, i j}(y) D_{x}^{i} D_{y}^{j} \tag{4.1.1}
\end{equation*}
$$

where the $a_{\lambda, j}, a_{\lambda, i j} \in \mathcal{C}_{c}^{\infty}(\mathbb{T})$ are constant whenever $\lambda y_{0} \leq|y| \leq \pi-\lambda y_{0}$. By definition, we know that for all $j \in\{0,1, \ldots, m\}$ and $y \in \mathbb{T}$, there exists $y^{\prime} \in \mathbb{R}$ such that $a_{\lambda, j}(y)=a_{j}\left(y^{\prime}\right)$. Hence (4.0.2) holds in the periodic setting, with $a_{j}$ replaced by $a_{\lambda, j}$. Moreover, for any multi-index $\alpha \in \mathbb{N}^{2}$, we have $\lambda^{|\alpha|}\left(\sum_{j}\left\|\partial^{\alpha} a_{\lambda, j}\right\|_{L^{\infty}}+\sum_{i, j}\left\|\partial^{\alpha} a_{\lambda, i j}\right\|_{L^{\infty}}\right) \leq C$ uniformly in $\lambda$.

The Hamiltonian $H-\frac{E_{1}+E_{2}}{2}$ satisfies the same assumptions as $H$, thus we can take $-E_{1}=E_{2}=: E>0$ without loss of generality. To simplify the following calculations, we will assume that $\varphi^{\prime}$ is even, so we can let $\Upsilon \in \mathcal{C}_{c}^{\infty}\left(-1, E^{2}-\delta_{0}\right)$ for some $\delta_{0}>0$, such that $\varphi^{\prime}(x)=\Upsilon\left(x^{2}\right)$ for all $x \in \mathbb{R}$. Recall that by Corollary 2.4.4, the infinite-space conductivity
$\sigma_{I}(H)$ is independent of $\varphi \in \mathfrak{S}(0,1 ;-E, E)$.
The main result of this section is the following, where $\sigma_{I}$ and $\tilde{\sigma}_{I}$ are given by (1.0.1) and (4.0.3), respectively.

Theorem 4.1.1. For any $p>0$, there exists a constant $C_{p}>0$ such that $\left|\sigma_{I}(H)-\tilde{\sigma}_{I}\left(H_{\lambda}\right)\right| \leq$ $C_{p} \lambda^{p}$ as $\lambda \rightarrow 0$.

Hidden in Theorem 4.1.1 is the assertion that $\varphi^{\prime}\left(H_{\lambda}\right)$ is well defined. It suffices to show that $H_{\lambda}$ is self-adjoint [70], which we do by creating a continuous path between $H_{\lambda}$ and a corresponding differential operator with constant coefficients. To this end, define

$$
\begin{aligned}
H_{\lambda, \mu}:=\lambda^{m}\left(M_{0} D_{y}^{m}\right. & \left.+M_{m} D_{x}^{m}\right) \\
& +\mu\left(\lambda^{m} \sum_{j=0}^{m} a_{\lambda, j}(y) D_{x}^{j} D_{y}^{m-j}+\sum_{i+j \leq m-1} \lambda^{i+j} a_{\lambda, i j}(y) D_{x}^{i} D_{y}^{j}\right)
\end{aligned}
$$

for $\mu \in[0,1]$. Note that $H_{\lambda, \mu}$ is symmetric since $H_{\lambda}$ is.
Proposition 4.1.2. There exist positive constants $C_{1}$ and $C_{2}$ such that for all $f \in \mathcal{H}^{m}\left(\mathbb{T}^{2}\right)$,

$$
\left\|H_{\lambda, \mu} f\right\|^{2} \geq C_{1}\|f\|_{\lambda, m}^{2}-C_{2}\|f\|^{2}
$$

uniformly in $\lambda \in\left(0, \lambda_{0}\right]$ and $\mu \in[0,1]$.

In the following proof, we use the Gagliardo-Nirenberg inequality on the Torus which states that for any non-negative integers $i$ and $j$ satisfying $i+j \leq m$,

$$
\begin{equation*}
\left\|\partial_{x}^{i} \partial_{y}^{j} f\right\| \leq\left\|\partial_{x}^{m} f\right\|^{\frac{i}{m}}\left\|\partial_{y}^{m} f\right\|^{\frac{j}{m}}\|f\|^{1-\frac{i+j}{m}} \tag{4.1.2}
\end{equation*}
$$

for all $f \in \mathcal{H}^{m}\left(\mathbb{T}^{2}\right)$.
Proof. From the anti-commutation properties (4.0.2), we see that $H_{\lambda, \mu}^{2}=\lambda^{2 m}\left(M_{0}^{2} D_{y}^{2 m}+\right.$

$$
\begin{aligned}
& \left.M_{m}^{2} D_{x}^{2 m}\right)+A_{\lambda, \mu}+B_{\lambda, \mu}, \text { where } \\
& A_{\lambda, \mu}=\lambda^{2 m}\left(\left\{M_{0}, M_{m}\right\} D_{y}^{m} D_{x}^{m}+\mu \sum_{j=0}^{\left\lfloor\frac{m}{2}\right\rfloor}\left(D_{x}^{2 j} D_{y}^{m-j}\left\{M_{0}, a_{\lambda, 2 j}\right\} D_{y}^{m-j}\right.\right. \\
& \\
& \left.\left.\quad+D_{x}^{2(m-j)} D_{y}^{j}\left\{M_{m}, a_{\lambda, m-2 j}\right\} D_{y}^{j}\right)+\left(\mu \sum_{j=0}^{m} a_{\lambda, j}(y) D_{x}^{j} D_{y}^{m-j}\right)^{2}\right)
\end{aligned}
$$

is non-negative and

$$
B_{\lambda, \mu}=\sum_{i+j \leq 2 m-1} \lambda^{i+j} \tilde{a}_{\lambda, \mu, i j}(y) D_{x}^{i} D_{y}^{j}
$$

is of order $2 m-1$, with $\lambda^{|\alpha|} \sum_{i, j}\left\|\partial^{\alpha} \tilde{a}_{\lambda, \mu, i j}\right\|_{L^{\infty}} \leq C_{\alpha}$ uniformly in $\lambda$ and $\mu$ for all $\alpha \in \mathbb{N}^{2}$. Using that $M_{0}$ and $M_{m}$ are non-singular and Hermitian, we have

$$
\left\|H_{\lambda, \mu} f\right\|^{2} \geq \lambda^{2 m} c\left(\left\|D_{y}^{m} f\right\|^{2}+\left\|D_{x}^{m} f\right\|^{2}\right)+\left(f, B_{\lambda, \mu} f\right)
$$

for some $c>0$. The result then follows from (4.1.2) and the fact that $\left|\left(f, B_{\lambda, \mu} f\right)\right| \leq$ $C\|f\|_{\lambda, m}\|f\|_{\lambda, m-1}$.

Proposition 4.1.2 implies the existence of a positive constant $C$ such that $\|f\|_{\lambda, m} \leq C\|g\|$ for all $f \in \mathcal{H}^{m}\left(\mathbb{T}^{2}\right)$ and $g \in \mathcal{H}\left(\mathbb{T}^{2}\right)$ satisfying $\left(i-H_{\lambda, \mu}\right) f=g$ uniformly in $\lambda$ and $\mu$. Indeed, redefining the constants $C_{1}$ and $C_{2}$, we have

$$
\begin{equation*}
\|f\|_{\lambda, m}^{2} \leq C_{1}\left\|H_{\lambda, \mu} f\right\|^{2}+C_{2}\|f\|^{2}=C_{1}\|g-i f\|^{2}+C_{2}\|f\|^{2} \leq C^{2}\|g\|^{2} \tag{4.1.3}
\end{equation*}
$$

where the last inequality follows from the fact that $\|f\|^{2}=\Im\left(f,\left(i-H_{\lambda, \mu}\right) f\right)=\Im(f, g) \leq$ $\|f\|\|g\|$. Clearly the same bound (4.1.3) holds if $i-H_{\lambda, \mu}$ is replaced by $i+H_{\lambda, \mu}$.

Proposition 4.1.3. For all $\lambda \in\left(0, \lambda_{0}\right], H_{\lambda}$ is self-adjoint on $\mathcal{H}\left(\mathbb{T}^{2}\right)$ with domain of definition $\mathcal{H}^{m}\left(\mathbb{T}^{2}\right)$.

Proof. By (4.1.3) and regularity of the coefficients of $H_{\lambda, \mu}$, we can choose $N \in \mathbb{N}$ sufficiently
large such that $\left\|\left(H_{\lambda,(k+1) / N}-H_{\lambda, k / N}\right) f\right\| \leq\|g\| / 2$ for all $g \in \mathcal{H}\left(\mathbb{T}^{2}\right), f \in \mathcal{H}^{m}\left(\mathbb{T}^{2}\right)$ satisfying $\left(i-H_{\lambda, k / N}\right) f=g$, and $k \in\{0,1, \ldots, N-1\}$. Since $H_{\lambda, 0}$ is symmetric with constant coefficients, we know (or easily verify) that $i-H_{\lambda, 0}$ is a bijection $\mathcal{H}^{m}\left(\mathbb{T}^{2}\right) \rightarrow \mathcal{H}\left(\mathbb{T}^{2}\right)$. Now, suppose $i-H_{\lambda, k / N}$ is a bijection $\mathcal{H}^{m}\left(\mathbb{T}^{2}\right) \rightarrow \mathcal{H}\left(\mathbb{T}^{2}\right)$ for some $k \in\{0,1, \ldots, N-1\}$, meaning that $\left\|\left(H_{\lambda,(k+1) / N}-H_{\lambda, k / N}\right)\left(i-H_{\lambda, k / N}\right)^{-1}\right\| \leq 1 / 2$. Then

$$
i-H_{\lambda,(k+1) / N}=\left(1-\left(H_{\lambda,(k+1) / N}-H_{\lambda, k / N}\right)\left(i-H_{\lambda, k / N}\right)^{-1}\right)\left(i-H_{\lambda, k / N}\right)
$$

is a bijection $\mathcal{H}^{m}\left(\mathbb{T}^{2}\right) \rightarrow \mathcal{H}\left(\mathbb{T}^{2}\right)$, as the first factor on the above right-hand side can be inverted using the Neuman series. We conclude by induction that $i-H_{\lambda, 1}=i-H_{\lambda}$ is a bijection $\mathcal{H}^{m}\left(\mathbb{T}^{2}\right) \rightarrow \mathcal{H}\left(\mathbb{T}^{2}\right)$. The same reasoning implies that $i+H_{\lambda}$ is a bijection $\mathcal{H}^{m}\left(\mathbb{T}^{2}\right) \rightarrow \mathcal{H}\left(\mathbb{T}^{2}\right)$. Since $H_{\lambda}$ is symmetric, it follows from [70, Theorem VIII.3] that $H_{\lambda}$ is self-adjoint with domain of definition $\mathcal{H}^{m}\left(\mathbb{T}^{2}\right)$. This completes the proof.

For $\xi \in \mathbb{R}$, define the differential operator $\hat{H}(\xi)$ on $\mathcal{H}(\mathbb{R})$ by

$$
\hat{H}(\xi):=M_{0} D_{y}^{m}+M_{m} \xi^{m}+\sum_{j=0}^{m} a_{j}(y) \xi^{j} D_{y}^{m-j}+\sum_{i+j \leq m-1} a_{i j}(y) \xi^{i} D_{y}^{j}
$$

Since the coefficients of $H$ are independent of $x$, we have the decomposition

$$
H=\mathcal{F}_{\xi \rightarrow x}^{-1} \int_{\mathbb{R}}^{\oplus} \hat{H}(\xi) d \xi \mathcal{F}_{x \rightarrow \xi},
$$

with $\mathcal{F}$ the one dimensional Fourier transform in the $x$-variable.
For $\lambda \in\left(0, \lambda_{0}\right]$ and $\xi \in \mathbb{R}$, define the differential operator $\hat{H}_{\lambda}(\xi)$ on $\mathcal{H}(\mathbb{T})$ by

$$
\begin{equation*}
\hat{H}_{\lambda}(\xi):=\lambda^{m} M_{0} D_{y}^{m}+M_{m} \xi^{m}+\sum_{j=0}^{m} \lambda^{m-j} a_{\lambda, j}(y) \xi^{j} D_{y}^{m-j}+\sum_{i+j \leq m-1} \lambda^{j} a_{\lambda, i j}(y) \xi^{i} D_{y}^{j} \tag{4.1.4}
\end{equation*}
$$

Noting that the coefficients of $H_{\lambda}$ are independent of $x$, we see that for any $\psi \in \mathcal{H}^{m}(\mathbb{T})$ and $\xi \in \mathbb{Z}$,

$$
\left(\hat{H}_{\lambda}(\lambda \xi) \psi\right)(y)=e^{-i \xi x}\left(H_{\lambda} \psi_{\xi}\right)(x, y)
$$

for all $(x, y) \in \mathbb{T}^{2}$, where $\psi_{\xi} \in \mathcal{H}^{m}\left(\mathbb{T}^{2}\right)$ is given by $\psi_{\xi}(x, y)=e^{i \xi x} \psi(y)$.
One can easily apply the logic from the proofs of Propositions 4.1.2 and 4.1.3 to obtain analogous results for $\hat{H}_{\lambda}(\xi)$. Namely, there exist positive constants $C_{1}$ and $C_{2}$ such that for all $f \in \mathcal{H}^{m}(\mathbb{T})$,

$$
\begin{equation*}
\left\|\hat{H}_{\lambda}(\xi) f\right\|^{2} \geq\left(C_{1} \xi^{2 m}-C_{2}\right)\|f\|^{2}+C_{1}\left\|\lambda^{m} D_{y}^{m} f\right\|^{2} \tag{4.1.5}
\end{equation*}
$$

uniformly in $\lambda \in\left(0, \lambda_{0}\right]$ and $\xi \in \mathbb{R}$. As above, this implies that $\hat{H}_{\lambda}(\xi)$ is self-adjoint with domain of definition $\mathcal{H}^{m}(\mathbb{T})$. In addition, we obtain

Proposition 4.1.4. The operator $\hat{H}_{\lambda}^{2}(\xi)$ is self-adjoint with domain of definition $\mathcal{H}^{2 m}(\mathbb{T})$ and its spectrum consists only of eigenvalues that go to $+\infty$.

Proof. Clearly $\hat{H}_{\lambda}^{2}(\xi)$ is symmetric. Moreover, we see that $i+\hat{H}_{\lambda}^{2}(\xi)=\left(\frac{1-i}{\sqrt{2}}+\hat{H}_{\lambda}(\xi)\right)\left(\frac{-1+i}{\sqrt{2}}+\right.$ $\hat{H}_{\lambda}(\xi)$ ), with each factor on the right-hand side a bijection $\mathcal{H}^{k+2 m}(\mathbb{T}) \rightarrow \mathcal{H}^{k}(\mathbb{T})$ for any $k \in \mathbb{N}$. Thus $i+\hat{H}_{\lambda}^{2}(\xi)$ is a bijection $\mathcal{H}^{2 m}(\mathbb{T}) \rightarrow \mathcal{H}(\mathbb{T})$. The same holds for $i-\hat{H}_{\lambda}^{2}(\xi)$, meaning that $\hat{H}_{\lambda}^{2}(\xi)$ is self-adjoint with domain $\mathcal{H}^{2 m}(\mathbb{T})$.

To prove the spectral property, we write $\left(1+\hat{H}_{\lambda}^{2}(\xi)\right)^{-1}=\left(1+\hat{H}_{\lambda}^{2}(\xi)\right)^{-1}\left(1+\hat{H}_{\lambda, 0}^{2}(\xi)\right)(1+$ $\left.\hat{H}_{\lambda, 0}^{2}(\xi)\right)^{-1}$ for $\hat{H}_{\lambda, 0}(\xi):=M_{0} \lambda^{m} D_{y}^{m}+M_{m} \xi^{m}$. Since $\hat{H}_{\lambda, 0}(\xi)$ has constant coefficients, it is clear that $\left(1+\hat{H}_{\lambda, 0}^{2}(\xi)\right)^{-1}$ is Hilbert-Schmidt. Since $\left(1+\hat{H}_{\lambda}^{2}(\xi)\right)^{-1}\left(1+\hat{H}_{\lambda, 0}^{2}(\xi)\right)$ is bounded, we conclude that $\left(1+\hat{H}_{\lambda}^{2}(\xi)\right)^{-1}$ is Hilbert-Schmidt, hence compact. Therefore the spectrum of $\left(1+\hat{H}_{\lambda}^{2}(\xi)\right)^{-1}$ consists entirely of eigenvalues that converge to 0 , meaning that the spectrum of $\hat{H}_{\lambda}^{2}(\xi)$ consists only of eigenvalues tending to $+\infty$.

We now state the following well known result characterizing the spectrum of self-adjoint operators, which can be found in, e.g., [71, 81].

Theorem 4.1.5. Let $A: \mathcal{H}_{0} \rightarrow \mathcal{H}_{0}$ be self-adjoint with $\mathcal{H}_{0}$ a separable Hilbert space, and let $E_{1} \leq E_{2} \leq E_{3} \ldots$ be the eigenvalues of $A$ below the essential spectrum (counted with multiplicity), respectively, the infimum of the essential spectrum, once there are no more eigenvalues left. Let $v_{j}$ denote the eigenfunction corresponding to $E_{j}$ (when the latter is an eigenvalue), such that $\left\langle v_{i}, v_{j}\right\rangle=\delta_{i j}$ for all $i$ and $j$. Then for any $\ell \in \mathbb{N}_{+}$and $u \in$ $\operatorname{span}\left(v_{1}, v_{2}, \ldots, v_{\ell-1}\right)^{\perp},\langle u, A u\rangle \geq E_{\ell}\|u\|^{2}$. Moreover, we have

$$
\begin{aligned}
E_{\ell} & =\max _{\phi_{1}, \ldots, \phi_{\ell-1}} \min \left\{\langle\phi, A \phi\rangle: \phi \in \operatorname{span}\left(\phi_{1}, \ldots, \phi_{\ell-1}\right)^{\perp},\|\phi\|=1\right\} \\
& =\min _{\phi_{1}, \ldots, \phi_{\ell}} \max \left\{\langle\phi, A \phi\rangle: \phi \in \operatorname{span}\left(\phi_{1}, \ldots, \phi_{\ell}\right),\|\phi\|=1\right\} .
\end{aligned}
$$

We now gather useful properties of the infinite-space operator $\hat{H}(\xi)$ and its spectrum.
Proposition 4.1.6. Take $H=\operatorname{Op}(\sigma)$ as above, and fix $\delta>0$. Then for each $\xi \in \mathbb{R}, \hat{H}(\xi)$ and $\hat{H}^{2}(\xi)$ are self-adjoint with respective domains of definition $\mathcal{H}^{m}(\mathbb{R})$ and $\mathcal{H}^{2 m}(\mathbb{R})$. The spectrum of $\hat{H}^{2}(\xi)$ in the interval $\left[0, E^{2}-\delta\right)$ consists of a finite number of eigenvalues, each with finite multiplicity. These eigenvalues and corresponding eigenfunctions can be chosen so that they are analytic in $\xi$. Moreover, the rank of $\hat{H}^{2}(\xi)$ in $\left[0, E^{2}-\delta\right)$ is uniformly bounded in $\xi \in \mathbb{R}$ and vanishes whenever $|\xi|$ is sufficiently large.

Proof. Fix $\xi \in \mathbb{R}$. Since $\hat{H}(\xi)=\operatorname{Op}\left(\sigma_{\xi}\right)$ with $\sigma_{\xi} \in S_{1,0}^{m}$ elliptic, it follows from [20, 56] that $\hat{H}(\xi)$ is self-adjoint with domain of definition $\mathcal{H}^{m}(\mathbb{R})$. We know that $\hat{H}^{2}(\xi)=\operatorname{Op}\left(\tau_{\xi}\right)$, where $\tau_{\xi}(y, \zeta)=(\sigma \sharp \sigma)(y, \xi, \zeta)$. Applying the same argument to the elliptic symbol $\tau_{\xi} \in S_{1,0}^{2 m}$, wee see that $\hat{H}^{2}(\xi)$ is self-adjoint with domain of definition $\mathcal{H}^{2 m}(\mathbb{R})$. Define $\sigma_{ \pm, \xi}(\zeta):=\sigma_{ \pm}(\xi, \zeta)$ and $\tau_{ \pm, \xi}:=\sigma_{ \pm, \xi}^{2}$, and observe that $\tau_{\xi}=\tau_{+, \xi}$ whenever $y \geq y_{0}$, and $\tau_{\xi}=\tau_{-, \xi}$ whenever
$y \leq-y_{0}$. Since the $\tau_{ \pm, \xi}$ are independent of $y$,

$$
T_{ \pm, \xi}:=\operatorname{Op}\left(\tau_{ \pm, \xi}\right)=\left(\operatorname{Op}\left(\sigma_{ \pm, \xi}\right)\right)^{2} \geq 0
$$

Let $\chi \in \mathfrak{S}(0, \pi / 2,-1,1)$, and define $\chi_{\varepsilon}(y):=\chi(\varepsilon y)$ and

$$
\mathrm{Op}\left(\tau_{\varepsilon, \xi}\right)=T_{\varepsilon, \xi}:=\sin \left(\chi_{\varepsilon}(y)\right) T_{+, \xi} \sin \left(\chi_{\varepsilon}(y)\right)+\cos \left(\chi_{\varepsilon}(y)\right) T_{-, \xi} \cos \left(\chi_{\varepsilon}(y)\right)
$$

for $\varepsilon \in(0,1]$. We see that $\tau_{\varepsilon, \xi}=\sin ^{2}\left(\chi_{\varepsilon}(y)\right) \tau_{+, \xi}+\cos ^{2}\left(\chi_{\varepsilon}(y)\right) \tau_{-, \xi}+\varepsilon r_{\varepsilon, \xi}$, with $r_{\varepsilon, \xi} \in$ $S\left(\langle\zeta\rangle^{2 m}\right)$ uniformly in $\varepsilon$, and $\tau_{ \pm, \xi} \geq \max \left\{E^{2}, c\langle\zeta\rangle^{2 m}\right\}$. It follows that

$$
\tau_{\varepsilon, \xi} \geq \max \left\{E^{2}, c\langle\zeta\rangle^{2 m}\right\}-c^{\prime} \varepsilon\langle\zeta\rangle^{2 m}
$$

for some positive constant $c^{\prime}$, hence $\tau_{\varepsilon, \xi} \geq E^{2}-\delta / 4$ for $\varepsilon$ sufficiently small. Using the fact that $T_{\varepsilon, \xi} \geq 0$, we apply Lemma 4.1.7 below (with a unitary transformation) to conclude that $T_{\varepsilon, \xi}$ has no spectrum in the interval $\left(-\infty, E^{2}-\delta / 2\right)$ for $\varepsilon$ sufficiently small.

We now observe that

$$
\begin{aligned}
\left(i-T_{\varepsilon, \xi}\right)^{-1}- & \left(i-\hat{H}^{2}(\xi)\right)^{-1}= \\
& \left(i-T_{\varepsilon, \xi}\right)^{-1}\left(\hat{H}^{2}(\xi)-T_{\varepsilon, \xi}\right)\left(i-\hat{H}^{2}(\xi)\right)^{-1} \in \operatorname{Op}\left(S\left(\langle y\rangle^{-\infty}\langle\zeta\rangle^{-2 m}\right)\right)
\end{aligned}
$$

as $\tau_{\xi}-\tau_{\varepsilon, \xi} \in S\left(\langle y\rangle^{-\infty}\langle\zeta\rangle^{2 m}\right)$ and both resolvents have symbol in $S\left(\langle\zeta\rangle^{-2 m}\right)$. Since $\langle\zeta\rangle^{-2 m} \in$ $L^{2}(\mathbb{R})$, it follows that $\left(i-T_{\varepsilon, \xi}\right)^{-1}-\left(i-\hat{H}^{2}(\xi)\right)^{-1}$ is Hilbert-Schmidt, hence compact. By [81, Theorem 6.18], we conclude that the essential spectra of $T_{\varepsilon, \xi}$ and $\hat{H}^{2}(\xi)$ are the same, meaning that the spectrum of $\hat{H}^{2}(\xi)$ in $\left[0, E^{2}-\delta / 2\right)$ consists of eigenvalues, each with finite multiplicity. The number of these eigenvalues must be finite, as we can verify from the proof of Proposition 2.1.10 that $\Phi_{0}\left(\hat{H}^{2}(\xi)\right)$ is trace-class for all $\Phi_{0} \in \mathcal{C}_{c}^{\infty}\left(-1, E^{2}\right)$.

Fix $\xi \in \mathbb{R}$. Let $\psi \in \mathcal{C}_{c}^{\infty}(\mathbb{R}) \otimes \mathbb{C}^{n}$ such that $\|\psi\| \|=1$. Define $\tilde{\psi}:=\Lambda_{\lambda_{1}, 0}^{-1} \psi$, where
$\lambda_{1} \in\left(0, \lambda_{0}\right]$ is sufficiently small so that $\operatorname{supp}(\tilde{\psi}) \subset(-\pi / 2, \pi / 2)$. Then by (4.1.5),

$$
\|\hat{H}(\xi) \psi \mid\|=\left\|\hat{H}_{\lambda_{1}}(\xi) \mathcal{P} \tilde{\psi}\right\| \geq C_{1} \xi^{m}-C_{2}
$$

Since $\mathcal{C}_{c}^{\infty}(\mathbb{R})$ is dense in $\mathcal{H}^{m}(\mathbb{R})$ with respect to the $\mathcal{H}^{m}$-norm, we have proved that $\hat{H}^{2}(\xi)$ has no spectrum in $\left[0, E^{2}\right)$ whenever $|\xi|$ is sufficiently large.

Let $\nu_{1}(\xi) \leq \nu_{2}(\xi) \leq \cdots \leq \nu_{\tilde{N}(\xi)}(\xi)$ denote the eigenvalues of $\hat{H}(\xi)$ below $E^{2}-\delta / 2$. We showed above that $\tilde{N}(\xi)$ is finite for all $\xi \in \mathbb{R}$. Let $N(\xi)$ be the number of eigenvalues of $\hat{H}^{2}(\xi)$ in $\left[0, E^{2}-\delta\right)$. To show that $N(\xi)$ is bounded uniformly in $\xi$, we will first prove that the eigenvalues are uniformly equicontinuous in $\xi$. By "uniformly equicontinuous", we mean that for all $\varepsilon>0$, there exists $\eta>0$ such that for any $\left|\xi_{1}-\xi_{2}\right|<\eta$ and any $j$ satisfying $j \leq \tilde{N}\left(\xi_{1}\right)$, it follows that $\left|\nu_{j}\left(\xi_{2}\right)-\nu_{j}\left(\xi_{1}\right)\right|<\varepsilon$.

Fix $C_{0}>0$. Let $\bar{\xi}>0$ such that $\hat{H}^{2}(\xi)$ has no spectrum in $\left[0, E^{2}-\delta / 2\right)$ whenever $|\xi|>\bar{\xi}$. Since $\hat{H}^{2}(\xi+\rho)-\hat{H}^{2}(\xi)$ is a differential operator of order $2 m-1$ with coefficients bounded by $C|\rho|$, the Gagliardo-Nirenberg inequality (4.1.2) on $\mathbb{R}^{2}$ implies that

$$
\begin{equation*}
\left|\left(u,\left(\hat{H}^{2}(\xi+\rho)-\hat{H}^{2}(\xi)\right) u\right)\right| \leq C|\rho|\|u\|\left\|^{\frac{2 m-1}{m}}\right\|\|u\|^{\frac{1}{m}} \leq C C_{0}^{\frac{2 m-1}{m}}|\rho|\|u\|^{2} \tag{4.1.6}
\end{equation*}
$$

uniformly in $\rho \in[-1,1]$ and $\xi \in[-\bar{\xi}, \bar{\xi}]$, for all $u \in \mathcal{H}^{m}(\mathbb{R})$ satisfying $\|u\|_{m} \leq C_{0}\|u u\|$. By ellipticity of $\hat{H}(\xi)$, we can choose $C_{0}$ sufficiently large so that $\|\|\hat{H}(\xi) u\|\| \geq E\|u\|$ whenever $\|u\|_{m}>C_{0}\|u\|$ and $\xi \in \mathbb{R}$. Hence given any $\ell$-dimensional subspace $S_{\ell} \subset \mathcal{H}^{m}(\mathbb{R})$, the bound (4.1.6) holds uniformly in $u \in S_{\ell}$ and $\xi \in[-\bar{\xi}, \bar{\xi}]$ satisfying $\|\hat{H}(\xi) u\|<E\|u\|$. It follows that

$$
\left|\max _{u \in S_{\ell},\|u\|=1}\left(u, \hat{H}^{2}(\xi+\rho) u\right)-\max _{u \in S_{\ell},\|u\| \|=1}\left(u, \hat{H}^{2}(\xi) u\right)\right| \leq C C_{0}^{\frac{2 m-1}{2 m}}|\rho| ; \quad \rho \in[-1,1]
$$

uniformly in $\xi \in[-\bar{\xi}, \bar{\xi}]$ satisfying $\max _{u \in S_{\ell},\|u\|=1}\| \| \hat{H}(\xi) u\| \|<E$. Therefore by Theorem
4.1.5,

$$
\begin{aligned}
\left|\nu_{\ell}(\xi+\rho)-\nu_{\ell}(\xi)\right| & =\left|\min _{S_{\ell}} \max _{u \in S_{\ell},\|u\|=1}\left(u, \hat{H}^{2}(\xi+\rho) u\right)-\min _{S_{\ell}} \max _{u \in S_{\ell},\|u\|=1}\left(u, \hat{H}^{2}(\xi) u\right)\right| \\
& \leq C C_{0}^{\frac{2 m-1}{2 m}}|\rho|
\end{aligned}
$$

uniformly in $\rho \in[-1,1]$ and $\ell \leq \tilde{N}(\xi)$. Thus we have proven uniform equicontinuity of the eigenvalues.

Suppose by contradiction that $N(\xi)$ is not bounded uniformly in $\xi$. Since $N(\xi)$ vanishes outside $[-\bar{\xi}, \bar{\xi}]$, there exists a sequence $\left(\xi_{k}\right) \subset[-\bar{\xi}, \bar{\xi}]$ and a number $\xi_{*} \in[-\bar{\xi}, \bar{\xi}]$ such that $\xi_{k} \rightarrow \xi_{*}$ and $N\left(\xi_{k}\right) \rightarrow \infty$ as $k \rightarrow \infty$. Uniform equicontinuity of the eigenvalues implies that for all $k$ sufficiently large, all eigenvalues $\nu_{j}$ satisfying $\nu_{j}\left(\xi_{k}\right)<E^{2}-\delta$ must also satisfy $\nu_{j}\left(\xi_{*}\right)<E^{2}-\delta / 2$. This implies

$$
\tilde{N}\left(\xi_{*}\right) \geq N\left(\xi_{k}\right) \rightarrow \infty,
$$

which is a contradiction since we already proved that $\tilde{N}\left(\xi_{*}\right)$ must be finite.
It remains to show that the eigenvalues and eigenfunctions can be chosen analytic in $\xi$ when they lie in $\left[0, E^{2}-\delta\right)$. We have shown that there exists $\tilde{N}_{0}<\infty$ such that $\tilde{N}(\xi) \leq \tilde{N}_{0}$ for all $\xi \in \mathbb{R}$. Thus for all $\xi$ such that $\nu_{1}(\xi)<E^{2}-\delta$, there exists $k(\xi) \in\left\{1,2, \ldots, \tilde{N}_{0}\right\}$ such that $\nu_{k(\xi)+1}(\xi) \geq E^{2}-\delta$ and $\nu_{k(\xi)+1}(\xi)-\nu_{k(\xi)}(\xi) \geq \frac{\delta}{2 \tilde{N}_{0}}$. Uniform equicontinuity implies that there exists $\eta>0$ such that $\nu_{k(\xi)+1}(\hat{\xi})-\nu_{k(\xi)}(\hat{\xi}) \geq \frac{\delta}{4 \tilde{N}_{0}}$ whenever $|\hat{\xi}-\xi|<\eta$. Now write $[-\bar{\xi}, \bar{\xi}] \subset \cup_{j} B_{j}$, where $\left\{B_{j}\right\}$ is a finite collection of overlapping open intervals ( $B_{j} \cap$ $B_{j+1} \neq \emptyset$ ) of length $2 \eta$. Let $\xi_{j}$ be the midpoint of $B_{j}$. Then for every $j$, we can apply [58, Theorems VII.1.7 and VII.1.8] to the finite system of eigenvalues $\nu_{1}(\xi), \nu_{2}(\xi), \ldots, \nu_{k\left(\xi_{j}\right)}(\xi)$ and corresponding eigenfunctions to conclude that they can be chosen analytic in $\xi \in B_{j}$. Since the $B_{j}$ form an open cover of $[-\bar{\xi}, \bar{\xi}]$, this proves that all eigenvalues (and corresponding eigenfunctions) in the energy interval $\left[0, E^{2}-\delta\right)$ can be chosen analytic in $\xi$. This completes
the result.

Lemma 4.1.7. Suppose $A_{h}=\operatorname{Op}_{h}(a)$ is non-negative for all $h \in(0,1]$, where $a \in S_{1,0}^{m}$ is Hermitian-valued and elliptic. Then for any constant $c \geq 0$ satisfying $a_{\min } \geq c$, there exist constants $\beta>0$ and $0<h_{0} \leq 1$ such that $\left\langle\psi, A_{h} \psi\right\rangle \geq c-\beta h$ for all $\|\psi\|=1$ and $h \in\left(0, h_{0}\right]$.

Proof. We know that $A_{h}$ is self-adjoint by our assumptions on $a$. Since $A_{h} \geq 0$, it follows that $\left(1+A_{h}\right)^{-1}=\mathrm{Op}_{h}\left(r_{h}\right)$, where $S(1) \ni r_{h}=(1+a)^{-1}+O(h)$. By the semi-classical sharp Gårding inequality [34, Theorem 7.12], there exist constants $\beta_{0}>0$ and $0<h_{0} \leq 1$ such that

$$
\begin{equation*}
0<\left\langle\phi,\left(1+A_{h}\right)^{-1} \phi\right\rangle \leq\left((1+c)^{-1}+\beta_{0} h\right)\|\phi\|^{2} \tag{4.1.7}
\end{equation*}
$$

for all nonzero $\phi \in \mathcal{H}$ and $h \in\left(0, h_{0}\right]$. Let $\psi \in \mathcal{D}\left(A_{h}\right)$ such that $\|\psi\|=1$, and define $\phi=\left(1+A_{h}\right)^{1 / 2} \psi$. Then

$$
\left\langle\psi,\left(1+A_{h}\right) \psi\right\rangle=\|\phi\|^{2}=\frac{\|\phi\|^{2}}{\|\psi\|^{2}}=\frac{\|\phi\|^{2}}{\left\langle\phi,\left(1+A_{h}\right)^{-1} \phi\right\rangle} \geq 1+c-\beta h, \quad h \in\left(0, h_{0}\right]
$$

for some fixed constant $\beta$. This completes the proof.

Spectral approximations. Recall that $H^{\prime}=\operatorname{Op}\left(\sigma^{\prime}\right)$, where for every $y^{\prime} \in \mathbb{R}$, there exists $y \in \mathbb{R}$ such that $\sigma^{\prime}\left(y^{\prime}, \xi, \zeta\right)=\sigma(y, \xi, \zeta)$. Since $\sigma$ satisfies (H1), it follows that $\sigma^{\prime}$ is elliptic: $\left|\sigma_{\min }^{\prime}\right| \geq c\langle\xi, \zeta\rangle^{m}-1$. Hence the fact that $\sigma^{\prime}=\sigma_{+}$whenever $y \leq \pi-y_{0}$ and $\sigma^{\prime}=\sigma_{-}$ whenever $y \geq \pi+y_{0}$ implies that $\sigma^{\prime}$ also satisfies (H1), with the roles of $\sigma_{+}$and $\sigma_{-}$reversed. We conclude that Proposition 4.1.6 still holds if $H$ is replaced by $H^{\prime}$, where we would instead define $\tilde{\psi}:=\Lambda_{\lambda_{1}, \pi}^{-1} \psi$ in the third paragraph of the proof.

The approximation result of Theorem 4.1.1 requires a detailed analysis of the spectrum of $H_{\lambda}$ constructed in (4.1.1), which it inherits from that of the operators $H$ and $H^{\prime}$. For $\xi \in \mathbb{R}$, let $\mu_{1}(\xi) \leq \mu_{2}(\xi) \leq \mu_{3}(\xi) \ldots$ denote the combined eigenvalues of $\hat{H}^{2}(\xi)$ and $\hat{H}^{\prime 2}(\xi)$
below $\Sigma_{\text {ess }}\left(\hat{H}^{2}(\xi)\right) \cup \Sigma_{\text {ess }}\left(\hat{H}^{\prime 2}(\xi)\right)$, respectively, the infimum of $\Sigma_{\text {ess }}\left(\hat{H}^{2}(\xi)\right) \cup \Sigma_{\text {ess }}\left(\hat{H}^{\prime 2}(\xi)\right)$, once there are no more eigenvalues left. Let $R_{\xi}\left(\right.$ resp. $\left.R_{\xi}^{\prime}\right)$ be the set consisting of the indices $j$ for which $\mu_{j}(\xi)$ is an eigenvalue of $\hat{H}^{2}(\xi)$ (resp. $\hat{H}^{\prime 2}(\xi)$ ). Thus $R_{\xi}$ and $R_{\xi}^{\prime}$ form a partition of the indices below $\Sigma_{\text {ess }}\left(\hat{H}^{2}(\xi)\right) \cup \Sigma_{\text {ess }}\left(\hat{H}^{\prime 2}(\xi)\right)$. When $\mu_{j}(\xi) \in R_{\xi} \cup R_{\xi}^{\prime}$, we denote the corresponding normalized eigenfunction by $\psi_{j, \xi}(y)$. We choose the eigenfunctions so that $\left(\psi_{i, \xi}, \psi_{j, \xi}\right)=\delta_{i j}$ whenever $i, j \in R_{\xi}$ or $i, j \in R_{\xi}^{\prime}$.

Let $\tilde{s}(\xi)$ be the total number of eigenvalues of $\hat{H}^{2}$ and $\hat{H}^{\prime 2}$ lying in $\left[0, E^{2}-\frac{\delta_{0}}{2}\right.$ ). By Proposition 4.1.6 (and the above paragraphs), we know that $\tilde{s}(\xi)$ is indeed finite and that the eigenvalues $\mu_{1}(\xi) \ldots \mu_{\tilde{s}(\xi)}(\xi)$ make up the entire spectrum of $\hat{H}^{2}(\xi)$ and $\hat{H}^{\prime 2}(\xi)$ in $\left[0, E^{2}-\frac{\delta_{0}}{2}\right)$. Moreover, there exists $s_{0} \in \mathbb{N}$ such that $\tilde{s}(\xi) \leq s_{0}$ for all $\xi \in \mathbb{R}$, and there exists a compact interval $I \subset \mathbb{R}$ such that $\tilde{s}(\xi)=0$ for all $\xi \notin I$.

Let $\mu_{\lambda, 1}(\xi) \leq \mu_{\lambda, 2}(\xi) \leq \mu_{\lambda, 3}(\xi) \ldots$ denote the eigenvalues of the operator $\hat{H}_{\lambda}^{2}(\xi)$ defined on the torus, which we know are well defined and go to infinity by Proposition 4.1.4. Let $\theta_{\lambda, 1, \xi}, \theta_{\lambda, 2, \xi}, \theta_{\lambda, 3, \xi}, \ldots$ be the corresponding (orthonormalized) eigenfunctions.

We now prove a result of (collective) exponential decay of the eigenfunctions of an operator with constant coefficients outside a bounded interval. This verifies that the low-energy eigenfunctions are localized to the vicinity of the domain walls, in both the periodic and infinite-space settings.

Proposition 4.1.8. For any $\alpha \in \mathbb{N}$, there are positive constants $C$ and $r$ such that

$$
\left|\partial_{y}^{\alpha} \psi_{j, \xi}\right|(y) \leq C e^{-r|y|}
$$

uniformly in $j \in\{1,2, \ldots, \tilde{s}(\xi)\}$ and $\xi \in I$. Similarly, given any closed interval $T \subset \mathbb{T}$ not containing 0 or $\pi,\left\|\lambda^{\alpha} \partial_{y}^{\alpha} \theta_{\lambda, j, \xi}\right\|_{T} \leq C e^{-r / \lambda}$ uniformly in $j$ and $\xi$ satisfying $\mu_{\lambda, j}(\xi)<$ $E^{2}-\delta_{0} / 2$.

Here and below, we use the shorthand $\|\cdot\|_{T}:=\|\cdot\|_{L^{2}(T) \otimes \mathbb{C}^{n}}$ if $T \subset \mathbb{T}$.

Proof. We first prove the claim regarding the infinite-space eigenfunctions and assume $y>0$ for concreteness. Restricted to the region $y>y_{0}$, we have $\hat{H}(\xi)=\operatorname{Op}\left(\sigma_{+, \xi}(\zeta)\right)$, where $\sigma_{+, \xi}(\zeta)=\sigma_{+}(\xi, \zeta)$. Our spectral gap assumption in (H1) is that $\sigma_{+}^{2}(\xi, \zeta)$ does not have eigenvalues inside $\left[0, E^{2}\right.$ ) while here by construction, $\mu_{j} \in\left[0, E_{0}^{2}\right]$ with $E_{0}:=\sqrt{E^{2}-\frac{\delta_{0}}{2}}<E$. By construction of $\psi_{j, \xi}$, the statement thus follows if we can prove the bound for any solution $\psi$ of

$$
\begin{equation*}
\left(\sigma_{+, \xi}^{2}\left(D_{y}\right)-\mu\right) \psi(y)=0, \quad y>y_{0} \tag{4.1.8}
\end{equation*}
$$

for an eigenvalue $\mu \in\left[0, E_{0}^{2}\right]$. We replace the above system by the first-order system

$$
\frac{d}{d y} u=A u
$$

for $A=A_{\xi}$ a $N \times N$ matrix with $N=2 m n$ and $u=\left(\psi, \psi^{\prime}, \ldots, \psi^{(2 m-1)}\right)^{t}$. We then write $A=P J P^{-1}$ with $J$ in Jordan form and $u=P v$. Solving $J v=\frac{d}{d y} v$, we see that each eigenvalue $\nu$ of $J$ must correspond to an eigenfunction of the form $v=e^{\nu y} v_{0}$ so that $u=e^{\nu y} P v_{0}$. This means that for every such $\nu$, there exists a solution $\psi$ of (4.1.8) that is of the form $\psi(y)=e^{\nu y} \bar{\psi}_{0}$, where $\bar{\psi}_{0} \in \mathbb{C}^{n}$ is independent of $y$. Since $\psi \neq 0$ by definition, this shows that

$$
\begin{equation*}
\operatorname{det}\left(\sigma_{+, \xi}^{2}(-i \nu)-\mu\right)=0 \tag{4.1.9}
\end{equation*}
$$

that is, $\sigma_{+, \xi}^{2}(-i \nu)$ must have an eigenvalue that is equal to $\mu$. We will now show that $\nu$ must be bounded away from the imaginary axis, uniformly in all variables. First, observe that since $\sigma_{+, \xi}(\zeta)$ is bounded above by $C\langle\xi, \zeta\rangle^{m}$ and elliptic in $\zeta$, there exist positive constants $\nu_{1}$ and $\delta_{1}$ such that whenever $|\Im \nu|>\nu_{1}$ and $|\Re \nu|<\delta_{1}$, all eigenvalues of $\sigma_{+, \xi}(-i \nu)$ lie outside $(-E, E)$. Since $\sigma_{+, \xi}(\zeta)$ is continuous in $\zeta$ and $\sigma_{+, \xi}(-i \nu)$ does not have spectrum in
$(-E, E)$ when $\Re \nu=0$, there exists $\delta_{2}>0$ such that whenever $|\Im \nu| \leq \nu_{1}$ and $|\Re \nu|<\delta_{2}$, all eigenvalues of $\sigma_{+, \xi}(-i \nu)$ lie outside $\left[-E_{0}, E_{0}\right]$. Thus, taking $\tilde{\delta}:=\min \left\{\delta_{1}, \delta_{2}\right\}$, we have shown that all $\nu$ satisfying (4.1.9) must also satisfy $|\Re \nu| \geq \tilde{\delta}$.

It is clear that $\tilde{\delta}=\tilde{\delta}(\xi, \mu)$ is continuous in both variables. Since $\sigma_{+, \xi}(\zeta)$ is bounded above by $C\langle\xi, \zeta\rangle^{m}$ and elliptic in $\xi$, there exist positive constants $\xi_{0}$ and $\delta_{0}$ such that for all $|\xi| \geq \xi_{0}$, any $\nu$ solving (4.1.9) must also satisfy $|\Re \nu| \geq \delta_{0}$. Thus, setting $K:=\left[-\xi_{0}, \xi_{0}\right] \times\left[-E_{0}, E_{0}\right]$ and $\delta:=\min \left\{\delta_{0}, \min _{(\xi, \mu) \in K} \tilde{\delta}(\xi, \mu)\right\}>0$, we have shown that all solutions $\nu$ of (4.1.9) satisfy $|\Re \nu| \geq \delta>0$ uniformly in $\xi \in \mathbb{R}$ and $\mu \in\left[0, E_{0}^{2}\right]$.

Note that every solution $\psi$ of (4.1.8) is a finite linear combination of terms of the form $y^{k} e^{\nu y} \bar{\psi}_{k}$, with $\nu$ satisfying (4.1.9). Since $\psi$ is square integrable in $y$ (and thus cannot be exponentially increasing), this shows that $\psi$ and all of its derivatives are exponentially decreasing as $y \rightarrow \infty$, uniformly in $\xi$ and $\mu$. Thus we have proven the desired decay property for eigenfunctions of $\hat{H}(\xi)$ as $y \rightarrow \infty$. The argument for exponential decay as $y \rightarrow-\infty$ is identical, as is the proof for eigenfunctions of $\hat{H}^{\prime}(\xi)$.

The same ideas are used for the periodic eigenfunctions, only now we must handle the $\lambda$-dependence. By definition of $T$, there exists a closed interval $T_{1} \subset \mathbb{T}$, also not containing 0 or $\pi$, such that $T \subset T_{1}^{\circ}$. Suppose that $T_{1} \subset(0, \pi)$ for concreteness. The defining equation for the eigenfunctions $\theta_{\lambda, j, \xi}$ in $T_{1}$ is

$$
\sigma_{+}^{2}\left(\xi, \lambda D_{y}\right) \theta_{\lambda, j, \xi}(y)=\mu_{\lambda, j}(\xi) \theta_{\lambda, j, \xi}(y)
$$

This is equivalent to solving an equation of the form

$$
\begin{equation*}
\left(\sigma_{+}^{2}\left(D_{y}\right)-\mu\right) f(y)=0, \quad f(y)=\theta_{\lambda}(\lambda y) \tag{4.1.10}
\end{equation*}
$$

where $y \in T_{\lambda, 1}:=\lambda^{-1} T_{1}$ and $0 \leq \mu \leq E_{0}^{2}<E^{2}$. The differential equation in (4.1.10) is identical to (4.1.8), with $f$ replacing $\psi$. Hence $\theta_{\lambda}(y)=f(y / \lambda)$ is exponentially increasing or
decreasing in $T_{1}$ with a rate that is proportional to $\lambda^{-1}$. Since $\theta_{\lambda}(y)$ has bounded $L^{2}-$ norm on $T_{1}$, its $L^{2}$-norm on $T$ is bounded by $C e^{-r / \lambda}$ for $r>0$. Derivatives of $\theta_{\lambda}$ also satisfy such an estimate. This completes the proof.

What remains before proving Theorem 4.1.1 is to show that the combined spectrum (including eigenspaces) of the infinite-space Hamiltonians is a good approximation of the spectrum of the periodic Hamiltonian. We will truncate the infinite-space eigenfunctions $\psi_{j, \xi}$ so that they have compact support and can be embedded in the torus. Due to the rapid decay (in $y$ ) of these eigenfunctions and their derivatives, the truncated versions almost solve the eigenproblem, admitting a small residual term that goes to zero with $\lambda$. The fact that the eigenvalues for the periodic problem are themselves well approximated is the statement of Proposition 4.1.10, and follows from Lemma 4.1.9 and Theorem 4.1.5. Once the error in eigenvalues is controlled, we use Lemma 4.1.12 to show that the periodic eigenfunctions are well approximated by their infinite-space analogues, which will allow us to directly bound the error $\left|\tilde{\sigma}_{I}\left(H_{\lambda}\right)-\sigma_{I}(H)\right|$.

For the following results, $\xi \in \mathbb{R}$ is assumed to be arbitrary and is dropped from the notation for brevity. It will be clear that the below estimates are all uniform in $\xi$, given the uniform bounds from Proposition 4.1.8.

Eigenvalue approximation. Let $s$ be the total number of eigenvalues of $\hat{H}^{2}$ and $\hat{H}^{\prime 2}$ lying in $\left[0, E^{2}-\delta_{0}\right)$. For $\ell \in\{0,1, \ldots, s\}$, define $\tilde{R}_{\ell}:=\{1,2, \ldots, \ell\} \cap R$ and $\tilde{R}_{\ell}^{\prime}:=\{1,2, \ldots, \ell\} \cap R^{\prime}$. 1 We will need to approximate the eigenfunctions of the infinite-space Hamiltonian by functions that live on the torus. Let $\hat{\chi}_{0} \in \mathcal{C}_{c}^{\infty}(-\pi / 4, \pi / 4)$ such that $0 \leq \hat{\chi}_{0} \leq 1$ on $\mathbb{R}$ and $\hat{\chi}_{0}=1$ in $[-\pi / 8, \pi / 8]$. Similarly, let $\hat{\chi}_{\pi} \in \mathcal{C}_{c}^{\infty}(\pi-\pi / 4, \pi+\pi / 4)$ such that $0 \leq \hat{\chi}_{\pi} \leq 1$ on $\mathbb{R}$ and $\hat{\chi} \pi=1$ in $[\pi-\pi / 8, \pi+\pi / 8]$. Define $\psi_{\lambda, j}^{0}:=\hat{\chi}_{j} \Lambda_{\lambda, t_{j}}^{-1} \psi_{j}$, where $t_{j}=0$ if $j \in R$ and

[^1]$t_{j}=\pi$ if $j \in R^{\prime}$. Set $\psi_{\lambda, 1}:=N_{\lambda, 1} \psi_{\lambda, 1}^{0}$, and for all $j>1$ define
$$
\psi_{\lambda, j}=N_{\lambda, j}\left(\psi_{\lambda, j}^{0}-\sum_{i=1}^{j-1}\left\langle\psi_{\lambda, i}, \psi_{\lambda, j}^{0}\right\rangle \psi_{\lambda, i}\right)
$$
with $N_{\lambda, j}>0$ such that $\left\|\psi_{\lambda, j}\right\|=1$. Thus the $\psi_{\lambda, j}$ form an orthonormal set of functions on $\mathbb{R}^{2}$. Let $\tilde{\psi}_{\lambda, j}:=\mathcal{P} \psi_{\lambda, j}$, which is well defined since the support of each $\psi_{\lambda, j}$ is contained in an interval of length $\pi / 2$.

In words, what we have done is embed the infinite-space eigenfunctions on the torus by rescaling in $\lambda$ as necessary, truncating them to have compact support and orthonormalizing at the end. By the rapid decay of the $\psi_{j, \xi}$ (Proposition 4.1.8), the error we get from truncating and orthonormalizing goes to zero exponentially in $\lambda$. Namely, we have that for any $\alpha \in \mathbb{N}$, there exist positive constants $C$ and $r$ such that

$$
\begin{equation*}
\left\|\left\|\lambda^{|\alpha|} \partial_{y}^{\alpha}\left(\psi_{\lambda, j}-\Lambda_{\lambda, t_{j}}^{-1} \psi_{j}\right)\right\| \mid \leq C e^{-r / \lambda}\right. \tag{4.1.11}
\end{equation*}
$$

uniformly in $\lambda \in\left(0, \lambda_{0}\right]$ for all $j \in\{1,2, \ldots, s\}$.

Lemma 4.1.9. There exist positive constants $C$ and $r$ such that for all $\lambda \in\left(0, \lambda_{0}\right], \ell \in$ $\{0,1, \ldots, s\}$, and $u_{0} \in\left(\operatorname{span}\left\{\tilde{\psi}_{\lambda, j}\right\}_{j=1}^{\ell}\right)^{\perp}$ such that $\operatorname{supp}\left(u_{0}\right) \subset\left(-\frac{3 \pi}{4}, \frac{3 \pi}{4}\right)$ or $\operatorname{supp}\left(u_{0}\right) \subset$ $\left(\frac{\pi}{4}, \frac{7 \pi}{4}\right)$,

$$
\begin{equation*}
\left\langle u_{0}, \hat{H}_{\lambda}^{2} u_{0}\right\rangle \geq\left(\min \left\{\mu_{\ell+1}, E^{2}+1\right\}-C e^{-r / \lambda}\right)\left\|u_{0}\right\|^{2} . \tag{4.1.12}
\end{equation*}
$$

Proof. Suppose $\operatorname{supp}\left(u_{0}\right) \subset\left(-\frac{3 \pi}{4}, \frac{3 \pi}{4}\right)$ for concreteness and $\left\|u_{0}\right\|=1$ without loss of generality. Then for $u:=\Lambda_{\lambda, 0} \tilde{\mathcal{P}} u_{0}$, we have $\left\langle u_{0}, \hat{H}_{\lambda}^{2} u_{0}\right\rangle=\left(u, \hat{H}^{2} u\right)$. Define $\tilde{u}:=u-$ $\sum_{j \in \tilde{R}_{\ell}}\left(\psi_{j}, u\right) \psi_{j}$, so that $\left(\psi_{j}, \tilde{u}\right)=0$ for all $j \in R_{\ell}$, and hence

$$
\left(\tilde{u}, \hat{H}^{2} \tilde{u}\right) \geq\|\tilde{u}\|^{2} \min \left\{\mu_{j}: j \in\left(R \cup\left\{s_{0}+1\right\}\right) \cap\left\{\ell+1, \ell+2, \ldots, s_{0}+1\right\}\right\} \geq \mu_{\ell+1}\|\tilde{u}\|^{2}
$$

by Theorem 4.1.5. Since $\|\tilde{u}\| \|^{2}=1-\sum_{j \in \tilde{R}_{\ell}}\left|\left(\psi_{j}, u\right)\right|^{2}$, it follows that

$$
\left(u, \hat{H}^{2} u\right)=\left(\tilde{u}, \hat{H}^{2} \tilde{u}\right)+\sum_{j \in \tilde{R}_{\ell}} \mu_{j}\left|\left(\psi_{j}, u\right)\right|^{2} \geq \mu_{\ell+1}-\sum_{j \in \tilde{R}_{\ell}}\left(\mu_{\ell+1}-\mu_{j}\right)\left|\left(\psi_{j}, u\right)\right|^{2}
$$

Using that $\left(\Lambda_{\lambda, 0} \psi_{\lambda, j}, u\right)=0$, we see by (4.1.11) that

$$
\left|\left(\psi_{j}, u\right)\right|=\left|\left(\Lambda_{\lambda, 0} \psi_{\lambda, j}-\psi_{j}, u\right)\right| \leq\left|\left\|\Lambda_{\lambda, 0} \psi_{\lambda, j}-\psi_{j}|\|| \leq C e^{-r / \lambda}\right.\right.
$$

for all $j \in \tilde{R}_{\ell}$. We conclude that

$$
\left(u, \hat{H}^{2} u\right) \geq \mu_{\ell+1}\left(1-C e^{-r / \lambda}\right) \geq \min \left\{\mu_{\ell+1}, E^{2}+1\right\}-C e^{-r / \lambda}
$$

and the result is complete.

We will now show that the eigenvalues for the periodic problem are well approximated by those for the infinite-space problem. The lower bounds for the $\mu_{\lambda, j}(\xi)$ will be obtained using Theorem 4.1.5 and Lemma 4.1.9.

Proposition 4.1.10. There exist positive constants $C$ and $r$ such that

$$
\mu_{\ell}-C e^{-r / \lambda} \leq \mu_{\lambda, \ell} \leq \mu_{\ell}+C e^{-r / \lambda}
$$

uniformly in $\lambda \in\left(0, \lambda_{0}\right]$ for all $\ell \in\{1,2, \ldots, s\}$. Moreover, $\mu_{\lambda, s+1} \notin \operatorname{supp}(\Upsilon)$ for all $\lambda$ sufficiently small.

Proof. We recall that $\varphi^{\prime}(x)=\Upsilon\left(x^{2}\right)$. We first prove the upper bound. Fix $\lambda \in\left(0, \lambda_{0}\right]$ and $\ell \in\{1, \ldots, s\}$. Let $u=\sum_{j=1}^{\ell} a_{j} \tilde{\psi}_{\lambda, j}$ with the $a_{j} \in \mathbb{C}$ such that $\sum_{j=1}^{\ell}\left|a_{j}\right|^{2}=1$. Then

$$
\left\langle u, \hat{H}_{\lambda}^{2} u\right\rangle=\sum_{j=1}^{\ell} a_{j}\left\langle u, \hat{H}_{\lambda}^{2} \tilde{\psi}_{\lambda, j}\right\rangle=\sum_{j=1}^{\ell}\left|a_{j}\right|^{2} \mu_{j}+\sum_{j=1}^{\ell} a_{j}\left\langle u,\left(\hat{H}_{\lambda}^{2}-\mu_{j}\right) \tilde{\psi}_{\lambda, j}\right\rangle,
$$

with the first term on the right-hand side bounded above by $\mu_{\ell}$. To control the second term, observe that

$$
\left\|\left(\hat{H}_{\lambda}^{2}-\mu_{j}\right) \tilde{\psi}_{\lambda, j}\right\|=\left\|\left(A_{j}-\mu_{j}\right) \Lambda_{\lambda, t_{j}} \psi_{\lambda, j}\right\| \|,
$$

with $\left(A_{j}, t_{j}\right)=\left(\hat{H}^{2}, 0\right)$ if $j \in R$ and $\left(A_{j}, t_{j}\right)=\left(\hat{H}^{\prime 2}, \pi\right)$ if $j \in R^{\prime}$. Using that $\left(A_{j}-\mu_{j}\right) \psi_{j}=0$, we conclude that

$$
\left\langle u, \hat{H}_{\lambda}^{2} u\right\rangle \leq \mu_{\ell}+\sum_{j=1}^{\ell} \mid\left\|\left(A_{j}-\mu_{j}\right)\left(\Lambda_{\lambda, t_{j}} \psi_{\lambda, j}-\psi_{j}\right)\right\| \| \leq \mu_{\ell}+C e^{-r / \lambda}
$$

where the last inequality follows from (4.1.11). Since $u$ was an arbitrary function in a subspace of dimension $\ell$, Theorem 4.1.5 implies that $\mu_{\lambda, \ell} \leq \mu_{\ell}+C e^{-r / \lambda}$.

We will now prove the significantly more challenging lower bound. Let $T_{0}=\left(-\frac{3 \pi}{4},-\frac{\pi}{4}\right) \cup$ $\left(\frac{\pi}{4}, \frac{3 \pi}{4}\right)$ and $T_{1}=\left(-\frac{7 \pi}{8},-\frac{\pi}{8}\right) \cup\left(\frac{\pi}{8}, \frac{7 \pi}{8}\right)$ be subsets of the torus $\mathbb{T}$, so that $T_{0} \subset \subset T_{1} \subset \subset \mathbb{T}$. Fix $\chi_{0}>0$, and let $\tilde{\chi}: \mathbb{T} \rightarrow\left[0, \chi_{0}\right]$ be a smooth function supported in $T_{1}$ such that $\tilde{\chi}=\chi_{0}$ in $\bar{T}_{0}$. Let

$$
B_{\lambda}=\tilde{\chi} R_{m} \tilde{\chi}
$$

with $R_{m}$ an elliptic operator $1+\lambda^{2 m} D_{y}^{2 m}$. This is an operator that is large on $T_{0}$ as well as non-negative.

Let $\chi: \mathbb{T} \rightarrow[0,1]$ be a smooth function supported in $\left(-\frac{3 \pi}{4}, \frac{3 \pi}{4}\right)$ such that $\chi=1$ in $\left[-\frac{\pi}{4}, \frac{\pi}{4}\right]$. Note that $\operatorname{supp}\left(\chi^{\prime}\right) \subset T_{0}$. Let $u$ be an arbitrary function in $\mathcal{H}^{m}(\mathbb{T})$. Then,

$$
\begin{equation*}
\left\langle(1-\chi) u, \hat{H}_{\lambda}^{2} \chi u\right\rangle=\left\langle(1-\chi) \hat{H}_{\lambda} u, \chi \hat{H}_{\lambda} u\right\rangle+\left\langle(1-\chi) \hat{H}_{\lambda} u,\left[\hat{H}_{\lambda}, \chi\right] u\right\rangle-\left\langle\left[\hat{H}_{\lambda}, \chi\right] u, \hat{H}_{\lambda} \chi u\right\rangle . \tag{4.1.13}
\end{equation*}
$$

The first term on the above right-hand side is non-negative since both $\chi$ and $1-\chi$ are. We now control the remaining terms. Observe that all operators involved are differential
operators, and the coefficients of $\left[\hat{H}_{\lambda}, \chi\right]$ vanish wherever $\chi^{\prime}$ does. Thus

$$
\left\langle(1-\chi) \hat{H}_{\lambda} u,\left[\hat{H}_{\lambda}, \chi\right] u\right\rangle=\frac{1}{\chi_{0}^{2}}\left\langle(1-\chi) \hat{H}_{\lambda} \tilde{\chi} u,\left[\hat{H}_{\lambda}, \chi\right] \tilde{\chi} u\right\rangle
$$

We can similarly insert factors of $\tilde{\chi} / \chi_{0}$ in the third term of (4.1.13). Since $\hat{H}_{\lambda}$ is a differential operator of order $m$ and the commutators $\left[\hat{H}_{\lambda}, \chi\right]$ introduce an extra factor of $\lambda$, there exists a constant $C_{0}>0$ such that

$$
\begin{equation*}
\Re\left\langle(1-\chi) u, \hat{H}_{\lambda}^{2} \chi u\right\rangle \geq-\frac{C_{0} \lambda}{\chi_{0}^{2}}\left(\left\|\lambda^{m} D_{y}^{m} \tilde{\chi} u\right\|^{2}+\|\tilde{\chi} u\|^{2}\right) \tag{4.1.14}
\end{equation*}
$$

uniformly in $u \in \mathcal{H}^{m}(\mathbb{T})$. Since $\left\langle u, B_{\lambda} u\right\rangle=\|\tilde{\chi} u\|^{2}+\left\|\lambda^{m} D_{y}^{m} \tilde{\chi} u\right\|^{2}$, we can choose $\chi_{0}$ sufficiently large so that

$$
\begin{equation*}
\left\langle u, B_{\lambda} u\right\rangle+2 \Re\left\langle(1-\chi) u, \hat{H}_{\lambda}^{2} \chi u\right\rangle \geq \frac{C}{\chi_{0}^{2}}\left(\left\|\lambda^{m} D_{y}^{m} \tilde{\chi} u\right\|^{2}+\|\tilde{\chi} u\|^{2}\right) \geq C\|u\|_{T_{0}}^{2} \tag{4.1.15}
\end{equation*}
$$

for $C>0$ as large as necessary. Recall that $\|\cdot\|_{T_{0}}:=\|\cdot\|_{L^{2}\left(T_{0}\right) \otimes \mathbb{C}^{n}}$.
By Proposition 4.1.8, we know that the low-energy spectrum of $\hat{H}_{\lambda}^{2}$ satisfies

$$
\begin{equation*}
\max _{\psi \in \tilde{S}_{s} ;\|\psi\|=1}\left\langle B_{\lambda} \psi, \psi\right\rangle=\varepsilon \leq e^{-r / \lambda} \tag{4.1.16}
\end{equation*}
$$

for some $c>0$, with $\tilde{S}_{s}=\operatorname{span}\left(\theta_{\lambda, 1}, \ldots \theta_{\lambda, s}\right)$. Recall from the definition of $s$ that $\mu_{\lambda, \ell}<E^{2}$ remains in the (infinite domain) spectral gap for all $\ell \leq s$.

Let us consider the operator $\hat{H}_{\lambda}^{2}+B_{\lambda}$, also self-adjoint (this is proved as for $\hat{H}_{\lambda}^{2}$ ), with eigenvalues $\nu_{\lambda, j}$. For $1 \leq \ell \leq s$, we have by the min-max principle,

$$
\nu_{\lambda, \ell}=\min _{S_{\ell}} \max _{\psi \in S_{\ell},\|\psi\|=1}\left\langle\hat{H}_{\lambda}^{2} \psi, \psi\right\rangle+\left\langle B_{\lambda} \psi, \psi\right\rangle \leq \max _{\psi \in \tilde{S}_{\ell},\|\psi\|=1}\left\langle\hat{H}_{\lambda}^{2} \psi, \psi\right\rangle+\left\langle B_{\lambda} \psi, \psi\right\rangle \leq \mu_{\lambda, \ell}+\varepsilon
$$

We may therefore obtain a lower bound for $\nu_{\lambda, \ell}$ now thanks to the regularization $B_{\lambda}$. We
have

$$
\nu_{\lambda, \ell} \geq \min _{u \in S_{\ell-1}^{\perp},\|\psi\|=1}\left\langle\left(\hat{H}_{\lambda}^{2}+B_{\lambda}\right) u, u\right\rangle
$$

with now $S_{\ell-1}$ the span of $\tilde{\psi}_{\lambda, j}$ for $1 \leq j \leq \ell-1$ using both families of infinite-domain eigenfunctions properly embedded in the torus.

It follows from Lemma 4.1.9 and (4.1.15) that for $u \in S_{\ell-1}^{\perp}$,

$$
\begin{aligned}
\left\langle u,\left(\hat{H}_{\lambda}^{2}+B_{\lambda}\right) u\right\rangle & =\left\langle\hat{H}_{\lambda}^{2} \chi u, \chi u\right\rangle+\left\langle\hat{H}_{\lambda}^{2}(1-\chi) u,(1-\chi) u\right\rangle+2 \Re\left\langle\hat{H}_{\lambda}^{2} \chi u,(1-\chi) u\right\rangle+\left\langle B_{\lambda} u, u\right\rangle \\
& \geq\left(\mu_{\ell}-\eta\right)\|\chi u\|^{2}+\left(\mu_{\ell}-\eta\right)\|(1-\chi) u\|^{2}+C\|u\|_{T_{0}}^{2}
\end{aligned}
$$

with $\eta \leq C e^{-r / \lambda}$. So, with $C$ large, we have $\left\langle u,\left(\hat{H}_{\lambda}^{2}+B_{\lambda}\right) u\right\rangle \geq\left(\mu_{\ell}-\eta\right)\|u\|^{2}$ and thus get

$$
\mu_{\lambda, \ell} \geq \nu_{\lambda, \ell}-\varepsilon \geq \mu_{\ell}-\varepsilon-\eta \geq \mu_{\ell}-C e^{-r / \lambda}
$$

for all $\lambda$ sufficiently small.
It remains to prove that $\mu_{\lambda, s+1}$ lies above the support of $\Upsilon$ when $\lambda$ is small enough. By definition of $\Upsilon$, there exists $\delta^{\prime}>0$ such that $\Upsilon \in \mathcal{C}_{c}^{\infty}\left(-1, E^{2}-\delta_{0}-\delta^{\prime}\right)$. Let $u \in \mathcal{H}^{m}(\mathbb{T})$ such that $u \in S_{s}^{\perp}$ and $\|u\|=1$. Take $\chi$ as above, with the additional requirement that $\operatorname{supp}\left(\chi^{\prime}\right) \subset T_{0}^{\prime}$ for some $T_{0}^{\prime} \subset T_{0}$ satisfying $\|u\|_{T_{0}^{\prime}}^{2} \leq \frac{\delta^{\prime}}{E^{2}+1}$. By (4.1.3), which can easily be shown to hold in this one-dimensional setting (see (4.1.5)), it follows that

$$
\left\|\left[\hat{H}_{\lambda}, \chi\right]\left(i-\hat{H}_{\lambda}\right)^{-1}\right\| \leq C \lambda, \quad \lambda \in\left(0, \lambda_{0}\right]
$$

Writing $\left[\hat{H}_{\lambda}, \chi\right]=\left[\hat{H}_{\lambda}, \chi\right]\left(i-\hat{H}_{\lambda}\right)^{-1}\left(i-\hat{H}_{\lambda}\right)$ and $\hat{H}_{\lambda} \chi=\left[\hat{H}_{\lambda}, \chi\right]+\chi \hat{H}_{\lambda}$, we use (4.1.13) to conclude that

$$
\left\langle(1-\chi) u, \hat{H}_{\lambda}^{2} \chi u\right\rangle \geq-C \lambda\left(\left\|\hat{H}_{\lambda} u\right\|^{2}+\|u\|^{2}\right)=-C \lambda\left(\left\|\hat{H}_{\lambda} u\right\|^{2}+1\right)
$$

Hence by Lemma 4.1.9, we have

$$
\begin{aligned}
\left\langle u, \hat{H}_{\lambda}^{2} u\right\rangle & =\left\langle\chi u, \hat{H}_{\lambda}^{2} \chi u\right\rangle+\left\langle(1-\chi) u, \hat{H}_{\lambda}^{2}(1-\chi) u\right\rangle+2 \Re\left\langle(1-\chi) u, \hat{H}_{\lambda}^{2} \chi u\right\rangle \\
& \geq\left(\check{\mu}_{s+1}-C e^{-r / \lambda}\right)\|\chi u\|^{2}+\left(\check{\mu}_{s+1}-C e^{-r / \lambda}\right)\|(1-\chi) u\|^{2}-C \lambda\left(\left\|\hat{H}_{\lambda} u\right\|^{2}+1\right),
\end{aligned}
$$

where $\check{\mu}_{s+1}:=\min \left\{\mu_{s+1}, E^{2}+1\right\}$. It follows that $\left\langle u, \hat{H}_{\lambda}^{2} u\right\rangle \geq \check{\mu}_{s+1}\left(\|\chi u\|^{2}+\|(1-\chi) u\|^{2}\right)-$ $C \lambda$. Since

$$
1=\|u\|^{2}=\|\chi u\|^{2}+\|(1-\chi) u\|^{2}+2\langle(1-\chi) u, \chi u\rangle \leq\|\chi u\|^{2}+\|(1-\chi) u\|^{2}+\frac{1}{2}\|u\|_{T_{0}^{\prime}}^{2}
$$

we have shown that

$$
\left\langle u, \hat{H}_{\lambda}^{2} u\right\rangle \geq \check{\mu}_{s+1}-\frac{1}{2} \check{\mu}_{s+1}\|u\|_{T_{0}^{\prime}}^{2}-C \lambda \geq \check{\mu}_{s+1}-\frac{1}{2} \delta^{\prime}-C \lambda .
$$

Since $u$ was arbitrary, Theorem 4.1.5 implies that $\mu_{\lambda, s+1} \geq \check{\mu}_{s+1}-\delta^{\prime}$ for all $\lambda$ sufficiently small. We know that $\mu_{s+1} \geq E^{2}-\delta_{0}$ by definition, hence $\mu_{\lambda, s+1} \geq E^{2}-\delta_{0}-\delta^{\prime}$. This completes the proof.

Eigenspace approximation. For the following lemma, fix $\lambda \in(0,1]$ and let

$$
\left\{\phi_{1}, \phi_{2}, \ldots, \phi_{s}\right\} \subset \mathcal{H}^{m}(\mathbb{T})
$$

such that $\left\langle\phi_{i}, \phi_{j}\right\rangle=\delta_{i j}$. For any integers $j, k$, and $\ell$ satisfying $1 \leq j \leq k \leq \ell \leq s$, define

$$
\begin{equation*}
\alpha_{j, k, \ell}:=\left(1-\sum_{i=j}^{\ell}\left|\left\langle\theta_{\lambda, i}, \phi_{k}\right\rangle\right|^{2}\right)^{1 / 2}, \quad \alpha_{k, \ell}:=\alpha_{1, k, \ell}, \quad r_{j}:=\left(\hat{H}_{\lambda}^{2}-\mu_{\lambda, j}\right) \phi_{j} . \tag{4.1.17}
\end{equation*}
$$

Note that if we denote by $\Pi_{j, \ell}$ the orthogonal projector onto the span of $\left\{\theta_{\lambda, j}, \ldots, \theta_{\lambda, \ell}\right\}$, then we have

$$
\begin{equation*}
\alpha_{j, k, \ell}=\left\|\left(I-\Pi_{j, \ell}\right) \phi_{k}\right\| . \tag{4.1.18}
\end{equation*}
$$

Lemma 4.1.11. Let $j, k, \ell$ be integers satisfying $1 \leq j \leq k \leq \ell \leq s$. Then

$$
\begin{equation*}
\alpha_{j, k, \ell}^{2} \leq\left\|r_{k}\right\|^{2}\left(\mu_{\lambda, \ell+1}-\mu_{\lambda, k}\right)^{-2}+\sum_{i=1}^{j-1}\left\|r_{i}\right\|^{2}\left(\mu_{\lambda, j}-\mu_{\lambda, i}\right)^{-2} \tag{4.1.19}
\end{equation*}
$$

(4.1.18) means $\phi_{k}$ lives approximately in $\operatorname{span}\left(\theta_{\lambda, j}, \theta_{\lambda, j+1}, \ldots, \theta_{\lambda, \ell}\right)$ when the right-hand side of (4.1.19) is small.

Proof. For brevity, set $\theta_{i}:=\theta_{\lambda, i}$. We see that $\alpha_{j, k, \ell}^{2}=\alpha_{k, \ell}^{2}+\sum_{i=1}^{j-1}\left|\left\langle\theta_{i}, \phi_{k}\right\rangle\right|^{2}$, with

$$
\sum_{i=1}^{j-1}\left|\left\langle\theta_{i}, \phi_{k}\right\rangle\right|^{2} \leq \sum_{i=1}^{j-1}\left(1-\sum_{h=1}^{j-1}\left|\left\langle\theta_{i}, \phi_{h}\right\rangle\right|^{2}\right)=\sum_{h=1}^{j-1}\left(1-\sum_{i=1}^{j-1}\left|\left\langle\theta_{i}, \phi_{h}\right\rangle\right|^{2}\right)=\sum_{h=1}^{j-1} \alpha_{h, j-1}^{2}
$$

It remains to bound $\alpha_{k, \ell}^{2}$ and the $\alpha_{h, j-1}^{2}$. Define $\tilde{\phi}_{k}:=\phi_{k}-\sum_{i=1}^{\ell}\left\langle\theta_{i}, \phi_{k}\right\rangle \theta_{i}$, so that $\left\langle\theta_{i}, \tilde{\phi}_{k}\right\rangle=$ 0 for all $i \in\{1, \ldots, \ell\}$, and thus

$$
\begin{equation*}
\left\langle\tilde{\phi}_{k}, \hat{H}_{\lambda}^{2} \tilde{\phi}_{k}\right\rangle \geq \mu_{\lambda, \ell+1}\left\|\tilde{\phi}_{j}\right\|^{2}=\mu_{\lambda, \ell+1} \alpha_{k, \ell}^{2} \tag{4.1.20}
\end{equation*}
$$

by Theorem 4.1.5. We also see that $\hat{H}_{\lambda}^{2} \tilde{\phi}_{k}=\mu_{\lambda, k} \phi_{k}-\sum_{i=1}^{\ell} \mu_{\lambda, i}\left\langle\theta_{i}, \phi_{k}\right\rangle \theta_{i}+r_{k}$, and hence

$$
\begin{align*}
\left\langle\tilde{\phi}_{k}, \hat{H}_{\lambda}^{2} \tilde{\phi}_{k}\right\rangle & =\mu_{\lambda, k}-\mu_{\lambda, k} \sum_{i=1}^{\ell}\left|\left\langle\theta_{i}, \phi_{k}\right\rangle\right|^{2}+\left\langle\tilde{\phi}_{k}, r_{k}\right\rangle  \tag{4.1.21}\\
& =\mu_{\lambda, k} \alpha_{k, \ell}^{2}+\left\langle\tilde{\phi}_{k}, r_{k}\right\rangle \leq \mu_{\lambda, k} \alpha_{k, \ell}^{2}+\left\|r_{k}\right\| \alpha_{k, \ell}
\end{align*}
$$

Combining (4.1.20) and (4.1.21), we obtain that $\alpha_{k, \ell} \leq\left\|r_{k}\right\|\left(\mu_{\lambda, \ell+1}-\mu_{\lambda, k}\right)^{-1}$. The same
argument proves that $\alpha_{h, j-1} \leq\left\|r_{h}\right\|\left(\mu_{\lambda, j}-\mu_{\lambda, h}\right)^{-1}$ for all $h \in\{1,2, \ldots, j-1\}$, and the result is complete.

We will apply Lemma 4.1 .11 to the functions $\phi_{j}=\tilde{\psi}_{\lambda, j}$. We write $r_{j}=\left(\hat{H}_{\lambda}^{2}-\mu_{j}\right) \tilde{\psi}_{\lambda, j}+$ $\left(\mu_{j}-\mu_{\lambda, j}\right) \tilde{\psi}_{\lambda, j}$, meaning that $\left\|r_{j}\right\| \leq\left\|\left(\hat{H}_{\lambda}^{2}-\mu_{j}\right) \tilde{\psi}_{\lambda, j}\right\|+\left|\mu_{j}-\mu_{\lambda, j}\right| \leq C e^{-r / \lambda}$, where the last inequality follows from Proposition 4.1.10 and its proof. Thus if $\min \left\{\mu_{\lambda, \ell+1}-\mu_{\lambda, k}, \mu_{\lambda, j}-\right.$ $\left.\mu_{\lambda, j-1}\right\} \geq \delta$ for some $\delta>0$, then $\alpha_{j, k, \ell}^{2} \leq C \delta^{-2} e^{-r / \lambda}$. In addition, we have

$$
\begin{align*}
\beta_{j, k, \ell}^{2}: & =\left\|\theta_{\lambda, k}-\sum_{i=j}^{\ell}\left\langle\tilde{\psi}_{\lambda, i}, \theta_{\lambda, k}\right\rangle \tilde{\psi}_{\lambda, i}\right\|^{2}=1-\sum_{i=j}^{\ell}\left|\left\langle\tilde{\psi}_{\lambda, i}, \theta_{\lambda, k}\right\rangle\right|^{2} \\
& \leq \sum_{i^{\prime}=j}^{\ell}\left(1-\sum_{i=j}^{\ell}\left|\left\langle\tilde{\psi}_{\lambda, i}, \theta_{\lambda, i^{\prime}}\right\rangle\right|^{2}\right)=\sum_{i=j}^{\ell}\left(1-\sum_{i^{\prime}=j}^{\ell}\left|\left\langle\tilde{\psi}_{\lambda, i}, \theta_{\lambda, i^{\prime}}\right\rangle\right|^{2}\right)  \tag{4.1.22}\\
& =\sum_{i=j}^{\ell} \alpha_{j, i, \ell}^{2} \leq C(\ell+1-j) \delta^{-2} e^{-r / \lambda}
\end{align*}
$$

Lemma 4.1.12. Let $A_{\lambda}:=\sum_{i=0}^{m-1} a_{\lambda, i}(y) \lambda^{i} D_{y}^{i}$ be an operator on $\mathcal{H}(\mathbb{T})$, for some smooth coefficients $a_{\lambda, i}$ satisfying

$$
\sum_{i=0}^{m-1}\left\|\lambda^{k} D_{y}^{k} a_{\lambda, i}\right\|_{L^{\infty}(\mathbb{T})} \leq C^{\prime}
$$

uniformly in $\lambda$ for all $k \in\{0,1, \ldots, m\}$. Let $j \leq \ell$ be positive integers such that $\ell \leq s$ and define $\delta:=\min \left\{\mu_{\lambda, \ell+1}-\mu_{\lambda, \ell}, \mu_{\lambda, j}-\mu_{\lambda, j-1}\right\}$, where the second argument is ignored if $j=1$. Then there exist positive constants $C$ and $r$ such that

$$
\left|\sum_{i=j}^{\ell}\left(\left\langle\theta_{\lambda, i}, A_{\lambda} \theta_{\lambda, i}\right\rangle-\left\langle\tilde{\psi}_{\lambda, i}, A_{\lambda} \tilde{\psi}_{\lambda, i}\right\rangle\right)\right| \leq C \delta^{-1} e^{-r / \lambda}
$$

Proof. First, we observe that for all $i \in\{1, \ldots, s\}$ and $k \in\{0,1, \ldots, m\}$,

$$
\begin{equation*}
\left\|\lambda^{k} D_{y}^{k} \nu_{\lambda, i}\right\|^{2} \leq\left\|\lambda^{m} D_{y}^{m} \nu_{\lambda, i}\right\|^{\frac{2 k}{m}} \leq C_{0}\left(\left\|\hat{H}_{\lambda} \nu_{\lambda, i}\right\|^{2}+C_{1}\right)^{\frac{k}{m}} \leq C_{0}\left(E^{2}+C_{1}\right)^{\frac{k}{m}}, \quad \nu=\theta, \tilde{\psi} \tag{4.1.23}
\end{equation*}
$$

uniformly in $\lambda$, with the second inequality following from ellipticity of $\hat{H}_{\lambda}$. We write

$$
\begin{aligned}
& \sum_{i=j}^{\ell}\left(\left\langle\theta_{\lambda, i}, A_{\lambda} \theta_{\lambda, i}\right\rangle-\left\langle\tilde{\psi}_{\lambda, i}, A_{\lambda} \tilde{\psi}_{\lambda, i}\right\rangle\right)= \\
& \quad \sum_{i=j}^{\ell}\left(\left\langle\theta_{\lambda, i}-\sum_{i^{\prime}=j}^{\ell}\left\langle\tilde{\psi}_{\lambda, i^{\prime}}, \theta_{\lambda, i}\right\rangle \tilde{\psi}_{\lambda, i^{\prime}}, A_{\lambda} \theta_{\lambda, i}\right\rangle-\left\langle\tilde{\psi}_{\lambda, i}, A_{\lambda}\left(\tilde{\psi}_{\lambda, i}-\sum_{i^{\prime}=j}^{\ell}\left\langle\theta_{\lambda, i^{\prime}}, \tilde{\psi}_{\lambda, i}\right\rangle \theta_{\lambda, i^{\prime}}\right)\right\rangle\right),
\end{aligned}
$$

which implies that

$$
\left|\sum_{i=j}^{\ell}\left(\left\langle\theta_{\lambda, i}, A_{\lambda} \theta_{\lambda, i}\right\rangle-\left\langle\tilde{\psi}_{\lambda, i}, A_{\lambda} \tilde{\psi}_{\lambda, i}\right\rangle\right)\right| \leq \sum_{i=j}^{\ell}\left(\beta_{j, i, \ell}\left\|A_{\lambda} \theta_{\lambda, i}\right\|+\alpha_{j, i, \ell}\left\|A_{\lambda}^{*} \tilde{\psi}_{\lambda, i}\right\|\right)
$$

where $A_{\lambda}^{*}:=\sum_{i=0}^{m-1} \lambda^{i} D_{y}^{i} a_{\lambda, i}^{*}(y)$ is the formal adjoint of $A_{\lambda}$. By (4.1.23), we know that the norms of $A_{\lambda} \nu_{\lambda, i}$ and $A_{\lambda}^{*} \nu_{\lambda, i}$ are bounded uniformly in $\lambda$, for $\nu=\theta, \tilde{\psi}$. Using (4.1.22) and the corresponding bound for $\alpha_{j, i, \ell}^{2}$, we conclude that $\left|\sum_{i=j}^{\ell}\left(\left\langle\theta_{\lambda, i}, A_{\lambda} \theta_{\lambda, i}\right\rangle-\left\langle\tilde{\psi}_{\lambda, i}, A_{\lambda} \tilde{\psi}_{\lambda, i}\right\rangle\right)\right| \leq$ $C \delta^{-1} e^{-r / \lambda}$. This completes the result.

Conductivity approximation. We are now ready to prove Theorem 4.1.1. We will write $\tilde{\sigma}_{I}$ as a sum of inner products over $\xi \in \lambda \mathbb{Z}$, thus it is helpful to reintroduce the $\xi$-dependent notation.

Proof of Theorem 4.1.1. Since the coefficients of $H_{\lambda}$ are independent of $x$, we have

$$
2 \pi \tilde{\sigma}_{I}=2 \pi \operatorname{Tr} i Q\left[H_{\lambda}, P\right] \Upsilon\left(H_{\lambda}^{2}\right)=\sum_{\xi \in \lambda \mathbb{Z}} \sum_{j \in \mathbb{N}}\left\langle\theta_{\lambda, j, \xi}, i \lambda Q_{Y} G_{\lambda}(\xi) \Upsilon\left(\hat{H}_{\lambda}^{2}(\xi)\right) \theta_{\lambda, j, \xi}\right\rangle
$$

where

$$
G_{\lambda}(\xi):=m M_{m} \xi^{m-1}+\sum_{j=1}^{m} j \lambda^{m-j} a_{\lambda, j}(y) \xi^{j-1} D_{y}^{m-j}+\sum_{i+j \leq m-1} i \lambda^{j} a_{\lambda, i j}(y) \xi^{i-1} D_{y}^{j}
$$

To obtain $G_{\lambda}$, we used that the contributions from $Q_{X}\left[H_{\lambda}, P\right]$ of all second and higher order derivatives of $P$ must vanish, as $\int_{\mathbb{T}} Q_{X} P^{(q)}(x) d x=P^{(q-1)}(\pi / 2)-P^{(q-1)}(-\pi / 2)=0$ for all $q>1$. Using that the $\theta_{\lambda, j, \xi}$ are eigenfunctions of $\hat{H}_{\lambda}^{2}(\xi)$, we obtain that

$$
2 \pi \tilde{\sigma}_{I}=\sum_{\xi \in \lambda \mathbb{Z} \cap I} \sum_{j=1}^{s_{0}}\left\langle\theta_{\lambda, j, \xi}, i \lambda Q_{Y} G_{\lambda}(\xi) \theta_{\lambda, j, \xi}\right\rangle \Upsilon\left(\mu_{\lambda, j}(\xi)\right)
$$

The next step is to partition the eigenvalues into clusters so that eigenvalues that are nearby belong to the same cluster and the separation between clusters is controlled (bounded from below). We will choose the clusters with diameter sufficiently small so that we can approximate all eigenvalues in a given cluster by the smallest eigenvalue in the cluster, with negligible error. Once $\Upsilon$ is constant over each cluster, we will approximate $\tilde{\sigma}_{I}$ by a corresponding sum involving the $\tilde{\psi}_{\lambda, j}$ and apply Lemma 4.1.12 to control the error.

By Proposition 4.1.10, the number of eigenvalues of $\hat{H}_{\lambda}^{2}(\xi)$ in $\operatorname{supp}(\Upsilon)$ is bounded by $s(\xi)$ uniformly in $\lambda>0$ sufficiently small and $\xi \in \lambda \mathbb{Z}$. Recall that $s(\xi)$ is the total number of eigenvalues of $\hat{H}^{2}(\xi)$ and $\hat{H}^{\prime 2}(\xi)$ lying in $\left[0, E^{2}-\delta_{0}\right)$, which is itself bounded by $s_{0}$ uniformly in $\xi$ and vanishes whenever $\xi \notin I$, with $I \subset \mathbb{R}$ a compact interval; see Proposition 4.1.6. Let $\xi \in \lambda \mathbb{Z} \cap I$ and $\delta:=\delta(\lambda)$. Define $k_{0}(\xi):=1$, and for all $i \geq 1$ define $k_{i}(\xi):=\inf \left\{\ell>k_{i-1}(\xi):\right.$ $\mu_{\lambda, \ell}(\xi)-\mu_{\lambda, \ell-1}(\xi) \geq \delta \quad$ or $\left.\quad \ell>s(\xi)\right\}$. We see that the $k_{i}(\xi)$ form an increasing sequence of integers, with $k_{J(\xi)}(\xi)=s(\xi)+1$ for some $J(\xi) \in \mathbb{N}$. Define $S_{\xi, j}:=\left\{k_{j}, \ldots, k_{j+1}-1\right\}$ for all $j \in\{0,1, \ldots, J(\xi)-1\}$, so that $S_{\xi, 0}, S_{\xi, 1}, \ldots, S_{\xi, J(\xi)-1}$ forms a partition of $\{1,2, \ldots, s(\xi)\}$.

Thus we have

$$
2 \pi \tilde{\sigma}_{I}=\sum_{\xi \in \lambda \mathbb{Z} \cap I} \sum_{i=0}^{J(\xi)-1} \sum_{j \in S_{\xi, i}}\left\langle\theta_{\lambda, j, \xi}, i \lambda Q_{Y} G_{\lambda}(\xi) \theta_{\lambda, j, \xi}\right\rangle \Upsilon\left(\mu_{\lambda, j}(\xi)\right)
$$

By (4.1.23), we see that $\left|\left\langle\theta_{\lambda, j, \xi}, G_{\lambda}(\xi) \theta_{\lambda, j, \xi}\right\rangle\right| \leq C$ uniformly in $\lambda$ and $\xi$. Defining

$$
2 \pi \tilde{\sigma}_{I, 1}:=\sum_{\xi \in \lambda \mathbb{Z} \cap I} \sum_{i=0}^{J(\xi)-1} \sum_{j \in S_{\xi, i}}\left\langle\theta_{\lambda, j, \xi}, i \lambda Q_{Y} G_{\lambda}(\xi) \theta_{\lambda, j, \xi}\right\rangle \Upsilon\left(\mu_{\lambda, k_{i}}(\xi)\right),
$$

which replaces each eigenvalue by the smallest eigenvalue in its cluster, we thus obtain that

$$
\begin{equation*}
\left|\tilde{\sigma}_{I, 1}-\tilde{\sigma}_{I}\right| \leq s_{0}(|I|+1) C \sup _{|x-y|<\delta}|\Upsilon(x)-\Upsilon(y)| \leq C \delta \tag{4.1.24}
\end{equation*}
$$

by regularity of $\Upsilon$, with $|I|$ the length of the interval $I$. We will now control the error from replacing the periodic eigenfunctions $\theta_{\lambda, j, \xi}$ in $\tilde{\sigma}_{I, 1}$ by the truncations $\tilde{\psi}_{\lambda, j, \xi}$. Namely, it follows from Lemma 4.1.12 that

$$
2 \pi \tilde{\sigma}_{I, 2}:=\sum_{\xi \in \lambda \mathbb{Z} \cap I} \sum_{i=0}^{J(\xi)-1} \sum_{j \in S_{\xi, i}}\left\langle\tilde{\psi}_{\lambda, j, \xi}, i \lambda Q_{Y} G_{\lambda}(\xi) \tilde{\psi}_{\lambda, j, \xi}\right\rangle \Upsilon\left(\mu_{\lambda, k_{i}}(\xi)\right)
$$

satisfies $\left|\tilde{\sigma}_{I, 2}-\tilde{\sigma}_{I, 1}\right| \leq C \delta^{-1} e^{-r / \lambda}$. Here, we have used the extra factor of $\lambda$ that appears in the inner product to cancel the $\lambda^{-1}$ scaling of the number of terms in the sum over $\xi$. We can simplify $\tilde{\sigma}_{I, 2}$ as

$$
2 \pi \tilde{\sigma}_{I, 2}=\sum_{\xi \in \lambda \mathbb{Z} \cap I} \sum_{i=0}^{J(\xi)-1} \sum_{j \in S_{\xi, i}^{\prime}}\left\langle\tilde{\psi}_{\lambda, j, \xi}, i \lambda G_{\lambda}(\xi) \tilde{\psi}_{\lambda, j, \xi}\right\rangle \Upsilon\left(\mu_{\lambda, k_{i}}(\xi)\right),
$$

with $S_{\xi, i}^{\prime}$ containing only the indices $j$ in $S_{\xi, i}$ such that $\operatorname{supp}\left(\tilde{\psi}_{j, \xi, \lambda}\right) \cap \operatorname{supp} Q_{Y} \neq \emptyset$. That is, $S_{\xi, i}^{\prime} \subseteq S_{\xi, i}$ and $\bigcup_{i=0}^{J(\xi)-1} S_{\xi, i}^{\prime}=R_{\xi}$. Here, we used the fact that $Q_{Y}=1$ on $\operatorname{supp}\left(\tilde{\psi}_{\lambda, j, \xi}\right)$
for all $j \in S_{\xi, i}^{\prime}$.
Note that for all $\xi \in \lambda \mathbb{Z} \cap I, i \in\{1, \ldots, J(\xi)-1\}$ and $j \in S_{\xi, i}^{\prime}$, we have

$$
\left|\mu_{j}(\xi)-\mu_{\lambda, k_{i}}(\xi)\right| \leq\left|\mu_{j}(\xi)-\mu_{\lambda, j}(\xi)\right|+\left|\mu_{\lambda, j}(\xi)-\mu_{\lambda, k_{i}}(\xi)\right| \leq C\left(e^{-r / \lambda}+\delta\right)
$$

by Proposition 4.1.10. Thus, by the same logic used to justify (4.1.24), we know that

$$
2 \pi \tilde{\sigma}_{I, 3}:=\sum_{\xi \in \lambda \mathbb{Z} \cap I} \sum_{i=0}^{J(\xi)-1} \sum_{j \in S_{\xi, i}^{\prime}}\left\langle\tilde{\psi}_{\lambda, j, \xi}, i \lambda G_{\lambda}(\xi) \tilde{\psi}_{\lambda, j, \xi}\right\rangle \Upsilon\left(\mu_{j}(\xi)\right)
$$

satisfies $\left|\tilde{\sigma}_{I, 3}-\tilde{\sigma}_{I, 2}\right| \leq C\left(e^{-r / \lambda}+\delta\right)$.
We now express $\tilde{\sigma}_{I, 3}$ in terms of functions in $\mathcal{H}(\mathbb{R})$ as

$$
2 \pi \tilde{\sigma}_{I, 3}=\lambda \sum_{\xi \in \lambda \mathbb{Z} \cap I} \sum_{i=0}^{J(\xi)-1} \sum_{j \in S_{\xi, i}^{\prime}}\left(\Lambda_{\lambda, 0} \psi_{\lambda, j, \xi}, \hat{G}(\xi) \Lambda_{\lambda, 0} \psi_{\lambda, j, \xi}\right) \Upsilon\left(\mu_{j}(\xi)\right)
$$

where $\hat{G}(\xi)=\operatorname{Op}\left(\tau_{\xi}\right)$ with $\tau_{\xi}(y, \zeta):=\partial_{\xi} \sigma(y, \xi, \zeta)$ Hermitian-valued. Defining

$$
2 \pi \tilde{\sigma}_{I, 4}:=\lambda \sum_{\xi \in \lambda \mathbb{Z} \cap I} \sum_{i=0}^{J(\xi)-1} \sum_{j \in S_{\xi, i}^{\prime}}\left(\psi_{j, \xi}, \hat{G}(\xi) \psi_{j, \xi}\right) \Upsilon\left(\mu_{j}(\xi)\right)
$$

it follows that

$$
\begin{aligned}
2 \pi\left(\tilde{\sigma}_{I, 4}-\tilde{\sigma}_{I, 3}\right)=\lambda \sum_{\xi \in \lambda \mathbb{Z} \cap I} & \sum_{i=0}^{J(\xi)-1} \sum_{j \in S_{\xi, i}^{\prime}}\left(\left(\psi_{j, \xi}-\Lambda_{\lambda, 0} \psi_{\lambda, j, \xi}, \hat{G}(\xi) \psi_{j, \xi}\right)\right. \\
& \left.+\left(\Lambda_{\lambda, 0} \psi_{\lambda, j, \xi}, \hat{G}(\xi)\left(\psi_{j, \xi}-\Lambda_{\lambda, 0} \psi_{\lambda, j, \xi}\right)\right)\right) \Upsilon\left(\mu_{j}(\xi)\right)
\end{aligned}
$$

so that

$$
\begin{aligned}
2 \pi\left|\tilde{\sigma}_{I, 4}-\tilde{\sigma}_{I, 3}\right| \leq \lambda & \sum_{\xi \in \lambda \mathbb{Z} \cap I}
\end{aligned} \sum_{i=0}^{J(\xi)-1} \sum_{j \in S_{\xi, i}^{\prime}}\left(\| \| \psi_{j, \xi}-\Lambda_{\lambda, 0} \psi_{\lambda, j, \xi}\| \|\left\|\hat{G}(\xi) \psi_{j, \xi}\right\| \|\right.
$$

Note that $\hat{G}(\xi) \in \operatorname{Op}\left(S_{1,0}^{m}\right)$ and $(i-\hat{H}(\xi))^{-1} \in \mathrm{Op}\left(S_{1,0}^{-m}\right)$ by $(\mathrm{H} 1)$, hence $\hat{G}(\xi)(i-\hat{H}(\xi))^{-1}$ is bounded for all $\xi \in I$. The norm is continuous as a function of $\xi$ in $I$ a compact interval, meaning that $\left\|\hat{G}(\xi)(i-\hat{H}(\xi))^{-1}\right\|$ is bounded uniformly in $\xi \in I$. Writing

$$
\hat{G}(\xi) \psi_{j, \xi}=\hat{G}(\xi)\left(1+\hat{H}^{2}(\xi)\right)^{-1}\left(1+\hat{H}^{2}(\xi)\right) \psi_{j, \xi}
$$

with $\left\|\left\|\left(1+\hat{H}^{2}(\xi)\right) \psi_{j, \xi}\right\|\right\| \leq 1+E^{2}$ for all $j \leq s(\xi)$, it follows from (4.1.11) that $\left|\tilde{\sigma}_{I, 4}-\tilde{\sigma}_{I, 3}\right| \leq$ $C e^{-r / \lambda}$.

So far, we have shown that $\left|\tilde{\sigma}_{I, 4}-\tilde{\sigma}_{I}\right| \leq C\left(\delta^{-1} e^{-r / \lambda}+e^{-r / \lambda}+\delta\right)$, for some positive constants $C$ and $r$. Therefore, setting $\delta:=e^{-r / 2 \lambda}$, we obtain that $\left|\tilde{\sigma}_{I, 4}-\tilde{\sigma}_{I}\right| \leq C e^{-r / 2 \lambda}$.

Finally, we recall from [7, 10] that

$$
2 \pi \sigma_{I}=\int_{\mathbb{R}} \sum_{i=0}^{J(\xi)-1} \sum_{j \in S_{\xi, i}^{\prime}}\left(\psi_{j, \xi}, \hat{G}(\xi) \psi_{j, \xi}\right) \Upsilon\left(\mu_{j}(\xi)\right) d \xi
$$

with the above integrand a smooth, compactly supported function of $\xi$ by Proposition 4.1.6. It is well known (see, e.g. [60]) that such integrals can be approximated by sampling over a uniform grid of size $\lambda$, with convergence faster than any power of $\lambda$. We conclude that $\left|\tilde{\sigma}_{I, 4}-\sigma_{I}\right| \leq C \lambda^{p}$, and the proof is complete.

Observe that $\left|\tilde{\sigma}_{I, 4}-\sigma_{I}\right|$ is the only error term that does not converge exponentially in $\lambda$. Thus if it were not for the approximation of the integral over $\xi$ by a discrete sum, we would get exponential convergence of the periodic conductivity to its corresponding infinite-space
value.

### 4.2 Stability of physical observable

The aim of this section is to prove stability of $\tilde{\sigma}_{I}\left(H_{\lambda}\right)$ under perturbations in the limit $\lambda \rightarrow 0$, with $H_{\lambda}$ given by (4.1.1) and $\tilde{\sigma}_{I}$ by (4.0.3). We consider perturbations of the form $V_{\lambda}:=$ $\sum_{i+j \leq m-1} \lambda^{i+j} v_{\lambda, i j}(x, y) D_{x}^{i} D_{y}^{j}$, where the $v_{\lambda, i j}$ are smooth matrix-valued functions that make $V_{\lambda}$ a symmetric differential operator on $L^{2}\left(\mathbb{T}^{2}\right) \otimes \mathbb{C}^{n}$. Assume that $\sum_{i, j}\left\|v_{\lambda, i j}\right\|_{L^{\infty}} \leq C$ uniformly in $\lambda$. Suppose there exists a closed set $S_{1} \subset \mathbb{T}^{2}$ such that $\cup_{i, j} \operatorname{supp}\left(v_{\lambda, i j}\right) \subset S_{1}$ for all $\lambda \in\left(0, \lambda_{0}\right]$ and $\operatorname{supp}\left((1-Q) P^{\prime}\right) \cap S_{1}=\emptyset$.

The condition on the sup-norm of $v_{\lambda, i j}$ guarantees that the coefficients of $H_{\lambda}+V_{\lambda}$ are bounded uniformly in $\lambda$. As in section 4.1 and combined with the assumption that $V_{\lambda}$ is symmetric, this ensures that $H_{\lambda}+\mu V_{\lambda}$ is uniformly elliptic (in the sense of Proposition 4.1.2) and hence self-adjoint. Note that $L^{\infty}$-bounds on derivatives of the $v_{\lambda, i j}$ are not necessary since $V_{\lambda}$ has no leading order terms. The condition on the support of $v_{\lambda, i j}$ is natural as it ensures that the perturbation is close to the domain wall of interest and well separated from the spurious domain wall introduced by the periodization. It allows us to express $\tilde{\sigma}_{I}\left(H_{\lambda}+V_{\lambda}\right)-\tilde{\sigma}_{I}\left(H_{\lambda}\right)$ as the integral over a product of operators, two of which are differential operators with disjoint support (see the proof of Theorem 4.2.1 below). Note that the assumption $\operatorname{supp}\left((1-Q) P^{\prime}\right) \cap S_{1}=\emptyset$ could easily be replaced by $\operatorname{supp}\left(Q P^{\prime}\right) \cap S_{1}=\emptyset$.

Our main stability result is the following

Theorem 4.2.1. Take $H_{\lambda}$ and $V_{\lambda}$ as above. Then for all $N \in \mathbb{N}$,

$$
\left|\tilde{\sigma}_{I}\left(H_{\lambda}+V_{\lambda}\right)-\tilde{\sigma}_{I}\left(H_{\lambda}\right)\right| \leq C_{N} \lambda^{N} .
$$

This result, combined with that of Theorem 4.1.1, shows that the periodic approximation of the conductivity enjoys a spectral (almost exponential) convergence property with $\sigma_{I}(H)-$
$\tilde{\sigma}_{I}\left(H_{\lambda}+V_{\lambda}\right)$ being of order $\lambda^{p}$ for any $p \geq 0$.
The rest of this section is devoted to proving Theorem 4.2.1. For $\varepsilon \in\{0,1\}$, define $H_{\lambda, \varepsilon}:=H_{\lambda}+\varepsilon V_{\lambda} .{ }^{2}$ The arguments from section 4.1 can easily be adapted to prove that the $H_{\lambda, \varepsilon}$ are self-adjoint for all $\lambda$, with

$$
\begin{equation*}
\left\|H_{\lambda, \varepsilon} f\right\|^{2} \geq C_{1}\|f\|_{\lambda, m}^{2}-C_{2}\|f\|^{2}, \quad\left\|\lambda^{|\alpha|} D^{\alpha}\left(i-H_{\lambda, \varepsilon}\right)^{-1}\right\| \leq C \tag{4.2.1}
\end{equation*}
$$

for all $f \in \mathcal{H}^{m}\left(\mathbb{T}^{2}\right)$ and $|\alpha| \leq m$ uniformly in $\lambda \in\left(0, \lambda_{0}\right]$.
We first show that $\tilde{\sigma}_{I}$ is unchanged for $Q$ replaced by $Q-1$, using $1=Q+1-Q$ and the following result:

Lemma 4.2.2. Let $H$ be a self-adjoint linear operator and $\Phi \in \mathcal{C}_{c}^{\infty}(\mathbb{R})$ such that $[H, P] \Phi(H)$, $H \Phi(H)$, and $\Phi(H)$ are trace-class. Then $\operatorname{Tr} i[H, P] \Phi(H)=0$.

Proof. Since $P$ is bounded, $P H \Phi(H)$ is trace class, and thus so is $H P \Phi(H)=[H, P] \Phi(H)+$ $P H \Phi(H)$. Using Lemma 2.1.2 (which still applies in the periodic setting) and cyclicity of the trace, we have

$$
\begin{aligned}
\operatorname{Tr} i[H, P] \Phi(H)=\operatorname{Tr} i[\Psi(H), P] \Phi(H) & =\operatorname{Tr} i \Psi(H) P \Phi(H)-\operatorname{Tr} i P \Psi(H) \Phi(H) \\
& =\operatorname{Tr} i P \Phi(H) \Psi(H)-\operatorname{Tr} i P \Psi(H) \Phi(H)=0,
\end{aligned}
$$

as desired.

Clearly, and thankfully, the assumptions of Lemma 4.2.2 were not satisfied in the infinitespace setting. However, due to compactness of the torus, these assumptions hold in the periodic setting for the Hamiltonians we consider as we show below. The filter $Q$ that appears in $\tilde{\sigma}_{I}$ thus allows for non-vanishing conductivities.

We now bound the trace-norm of appropriate functionals of the $H_{\lambda, \varepsilon}$.

[^2]Proposition 4.2.3. For all $\Phi \in \mathcal{C}_{c}^{\infty}(\mathbb{R})$ and $\varepsilon \in\{0,1\}$, there is a positive constant $C$ such that $\left\|\Phi\left(H_{\lambda, \varepsilon}\right)\right\|_{1} \leq C \lambda^{-2}$ uniformly in $\lambda \in\left(0, \lambda_{0}\right]$.

Proof. Let $M>0$ such that $\Phi \in \mathcal{C}_{c}^{\infty}(-M, M)$. For $(\xi, \zeta) \in \mathbb{Z}^{2}$ and $j \in\{1,2, \ldots, n\}$, define $\phi_{\xi, \zeta, j}(x, y)=\frac{1}{2 \pi} e^{i(\xi x+\zeta y)} v_{j}$, where $v_{j}$ is the $j$ th column of the $n \times n$ identity matrix. Thus the $\phi_{\xi, \zeta, j}$ form an orthonormal basis for $\mathcal{H}\left(\mathbb{T}^{2}\right)$. Let $k \in \mathbb{N}$ and define $S_{k}:=(\mathbb{Z} \cap[-k+$ $1, k-1])^{2} \times(\mathbb{N} \cap[1, n])$ and $T_{k}:=\left(\mathbb{Z}^{2} \times(\mathbb{N} \cap[1, n])\right) \backslash S_{k}$. Let $u=\sum_{(\xi, \zeta, j) \in T_{k}} c_{\xi, \zeta, j} \phi_{\xi, \zeta, j}$, with $\left(c_{\xi, \zeta, j}\right) \subset \mathbb{C}$ chosen so that $u \in \mathcal{H}^{m}\left(\mathbb{T}^{2}\right)$. It follows that

$$
\begin{aligned}
\|u\|_{\lambda, m}^{2} & =\sum_{|\alpha| \leq m} \sum_{(\xi, \zeta, j) \in T_{k}}\left|c_{\xi, \zeta, j}\right|^{2}(\lambda \xi)^{2 \alpha_{1}}(\lambda \zeta)^{2 \alpha_{2}} \\
& \geq(\lambda k)^{2 m} \sum_{(\xi, \zeta, j) \in T_{k}}\left|c_{\xi, \zeta, j}\right|^{2}=(\lambda k)^{2 m}\|u\|^{2} .
\end{aligned}
$$

To justify the inequality, we sum over $\alpha \in\{(m, 0),(0, m)\}$ and use the fact that for all $(\xi, \zeta, j) \in T_{k}, \xi \geq k$ or $\zeta \geq k$. Thus if $(\lambda k)^{2 m} \geq C_{1}^{-1}\left(M^{2}+C_{2}\right)$, then (4.2.1) implies that $\left\|H_{\lambda, \varepsilon} u\right\|^{2} \geq M^{2}\|u\|^{2}$. Since $u$ is an arbitrary function in $\operatorname{span}\left(\phi_{\xi, \zeta, j}:(\xi, \zeta, j) \in S_{k}\right)^{\perp}$, Theorem 4.1.5 implies that the spectrum of $H_{\lambda, \varepsilon}^{2}$ in $\left(-\infty, M^{2}\right)$ consists entirely of eigenvalues, and the number of these eigenvalues is bounded by $C \lambda^{-2}$ for some $C>0$. Hence the number of eigenvalues of $H_{\lambda, \varepsilon}$ in $(-M, M)$ is also bounded by $C \lambda^{-2}$, and the result follows.

We are now ready to prove the main result of this section. Below we bound the difference of conductivities $\left|\tilde{\sigma}_{I}\left(H_{\lambda, 1}\right)-\tilde{\sigma}_{I}\left(H_{\lambda, 0}\right)\right|$ by the product of a trace-norm and operator-norm. Proposition 4.2.3 provides a ( $\lambda$-dependent) bound on the involved trace-norm. Combined with (4.2.1), this bound verifies the assumptions of Lemma 4.2.2. We conclude by showing that the operator norm goes to zero faster than any power of $\lambda$.

Proof of Theorem 4.2.1. Let $W=W(x, y)$ be a smooth point-wise multiplication operator
proportional to the identity matrix. By (4.2.1), we have

$$
\begin{align*}
\left\|\left[H_{\lambda, \varepsilon^{\prime}}, W\right]\left(z-H_{\lambda, \varepsilon}\right)^{-1}\right\| & =\left\|\left[V_{\lambda}, W\right]\left(i-H_{\lambda, \varepsilon}\right)^{-1}\left(i-H_{\lambda, \varepsilon}\right)\left(z-H_{\lambda, \varepsilon}\right)^{-1}\right\| \\
& \leq\left\|\left[V_{\lambda}, W\right]\left(i-H_{\lambda, \varepsilon}\right)^{-1}\right\|\left(1+|i-z|\left\|\left(z-H_{\lambda, \varepsilon}\right)^{-1}\right\|\right)  \tag{4.2.2}\\
& \leq C|\Im z|^{-1} \lambda
\end{align*}
$$

for all $\left(\varepsilon, \varepsilon^{\prime}\right) \in\{0,1\}^{2}$ uniformly in $\lambda \in\left(0, \lambda_{0}\right]$ and $z \in Z \backslash\{\Im z=0\}$, where $Z:=\left[E_{1}, E_{2}\right] \times$ $[-2,2] \subset \mathbb{C}$. Similarly,

$$
\begin{equation*}
\left\|V_{\lambda}\left(z-H_{\lambda, \varepsilon}\right)^{-1}\right\| \leq C|\Im z|^{-1} \tag{4.2.3}
\end{equation*}
$$

uniformly in $\lambda \in\left(0, \lambda_{0}\right]$ and $z \in Z \backslash\{\Im z=0\}$. Let $\Phi \in \mathcal{C}_{c}^{\infty}\left(E_{1}, E_{2}\right)$ such that $\Phi \varphi^{\prime}=\varphi^{\prime}$. Applying Proposition 4.2 .3 to the compactly supported function $x \mapsto(i-x) \Phi(x)$, we obtain by (4.2.2) that

$$
\begin{equation*}
\left\|\left[H_{\lambda, \varepsilon}, W\right] \Phi\left(H_{\lambda, \varepsilon}\right)\right\|_{1} \leq\left\|\left[H_{\lambda, \varepsilon}, W\right]\left(i-H_{\lambda, \varepsilon}\right)^{-1}\right\|\left\|\left(i-H_{\lambda, \varepsilon}\right) \Phi\left(H_{\lambda, \varepsilon}\right)\right\|_{1} \leq C \lambda^{-1} \tag{4.2.4}
\end{equation*}
$$

uniformly in $\lambda$.
Fix $N \in \mathbb{N}$ and define $\tilde{Q}:=Q-1$. With Proposition 4.2.3 and (4.2.4), we have verified that $H_{\lambda, \varepsilon}$ satisfies the assumptions of Lemma 4.2.2 for $\varepsilon \in\{0,1\}$, hence $\tilde{\sigma}_{I}\left(H_{\lambda, \varepsilon}\right)=$ $\operatorname{Tr} i \tilde{Q}\left[H_{\lambda, \varepsilon}, P\right] \varphi^{\prime}\left(H_{\lambda, \varepsilon}\right)$. By assumption, $\tilde{Q}\left[V_{\lambda}, P\right]=0$, so that

$$
\tilde{\sigma}_{I}\left(H_{\lambda, \varepsilon}\right)=\operatorname{Tr} i \tilde{Q}\left[H_{\lambda, 0}, P\right] \varphi^{\prime}\left(H_{\lambda, \varepsilon}\right)
$$

Moreover, there exists a ( $\lambda$-independent) smooth point-wise multiplication operator $R$ proportional to the identity matrix, such that $R=1$ on $\operatorname{supp}\left(\tilde{Q} P^{\prime}\right)$ and $\operatorname{supp}(R) \cap S_{1}=\emptyset$. It
follows that

$$
\begin{aligned}
\tilde{\sigma}_{I}\left(H_{\lambda, 0}\right)-\tilde{\sigma}_{I}\left(H_{\lambda, 1}\right)= & \operatorname{Tr} i R \tilde{Q}\left[H_{\lambda, 0}, P\right]\left(\varphi^{\prime}\left(H_{\lambda, 0}\right)-\varphi^{\prime}\left(H_{\lambda, 1}\right)\right) \\
= & \operatorname{Tr} i \tilde{Q}\left[H_{\lambda, 0}, P\right] \Phi\left(H_{\lambda, 0}\right)\left(\varphi^{\prime}\left(H_{\lambda, 0}\right)-\varphi^{\prime}\left(H_{\lambda, 1}\right)\right) R \\
& +\operatorname{Tr} i \tilde{Q}\left[H_{\lambda, 0}, P\right] R\left(\Phi\left(H_{\lambda, 0}\right)-\Phi\left(H_{\lambda, 1}\right)\right) \varphi^{\prime}\left(H_{\lambda, 1}\right)=: \Delta_{0}+\Delta_{1}
\end{aligned}
$$

where we have used cyclicity of the trace to move $R$ to the right-most position on the second line. By the Helffer-Sjöstrand formula, we see that

$$
\left(\varphi^{\prime}\left(H_{\lambda, 0}\right)-\varphi^{\prime}\left(H_{\lambda, 1}\right)\right) R=\frac{1}{\pi} \int_{Z} \bar{\partial}_{\varphi^{\prime}}(z)\left(z-H_{\lambda, 1}\right)^{-1} V_{\lambda}\left(z-H_{\lambda, 0}\right)^{-1} R d^{2} z
$$

Since $\operatorname{supp}(R)$ and $S_{1}$ are closed and disjoint sets, there exists a ( $\lambda$-independent) collection $\left\{W_{0}, W_{1}, \ldots, W_{N}\right\}$ of smooth, point-wise multiplication operators proportional to the identity matrix such that $W_{j} V_{\lambda}=0$ and $W_{j+1} W_{j}=W_{j}$ for all $j$ and $\lambda$, with $W_{0}:=R$. Using the fact that $\left[\left(z-H_{\lambda, 0}\right)^{-1}, W_{j}\right]=\left(z-H_{\lambda, 0}\right)^{-1}\left[H_{\lambda, 0}, W_{j}\right]\left(z-H_{\lambda, 0}\right)^{-1}$, we obtain that

$$
\begin{aligned}
V_{\lambda}\left(z-H_{\lambda, 0}\right)^{-1} R & =V_{\lambda}\left[\left(z-H_{\lambda, 0}\right)^{-1}, W_{0}\right]=V_{\lambda}\left(z-H_{\lambda, 0}\right)^{-1}\left[H_{\lambda, 0}, W_{0}\right]\left(z-H_{\lambda, 0}\right)^{-1} \\
& =V_{\lambda}\left(z-H_{\lambda, 0}\right)^{-1} W_{1}\left[H_{\lambda, 0}, W_{0}\right]\left(z-H_{\lambda, 0}\right)^{-1} \\
& =V_{\lambda}\left(z-H_{\lambda, 0}\right)^{-1}\left[H_{\lambda, 0}, W_{1}\right]\left(z-H_{\lambda, 0}\right)^{-1}\left[H_{\lambda, 0}, W_{0}\right]\left(z-H_{\lambda, 0}\right)^{-1} \\
& =V_{\lambda}\left(z-H_{\lambda, 0}\right)^{-1} W_{2}\left[H_{\lambda, 0}, W_{1}\right]\left(z-H_{\lambda, 0}\right)^{-1}\left[H_{\lambda, 0}, W_{0}\right]\left(z-H_{\lambda, 0}\right)^{-1} \\
& =\ldots \\
= & V_{\lambda}\left(z-H_{\lambda, 0}\right)^{-1}\left[H_{\lambda, 0}, W_{N}\right]\left(z-H_{\lambda, 0}\right)^{-1}\left[H_{\lambda, 0}, W_{N-1}\right]\left(z-H_{\lambda, 0}\right)^{-1} \\
& \ldots\left(z-H_{\lambda, 0}\right)^{-1}\left[H_{\lambda, 0}, W_{0}\right]\left(z-H_{\lambda, 0}\right)^{-1}
\end{aligned}
$$

By (4.2.2), $\left\|\left[H_{\lambda, 0}, W_{j}\right]\left(z-H_{\lambda, 0}\right)^{-1}\right\| \leq C_{j}|\Im z|^{-1} \lambda$ for every $j$. Thus by (4.2.3) and the rapid decay of $\bar{\partial} \tilde{\varphi}^{\prime}$ near the real axis, it follows that $\left\|\left(\varphi^{\prime}\left(H_{\lambda, 0}\right)-\varphi^{\prime}\left(H_{\lambda, 1}\right)\right) R\right\| \leq C \lambda^{N+1}$.

Using (4.2.4), we conclude that

$$
\left|\Delta_{0}\right| \leq\left\|\tilde{Q}\left[H_{\lambda, 0}, P\right] \Phi\left(H_{\lambda, 0}\right)\right\|_{1}\left\|\left(\varphi^{\prime}\left(H_{\lambda, 0}\right)-\varphi^{\prime}\left(H_{\lambda, 1}\right)\right) R\right\| \leq C \lambda^{N}
$$

By cyclicity of the trace, $\Delta_{1}=\operatorname{Tr} i \varphi^{\prime}\left(H_{\lambda, 1}\right) \tilde{Q}\left[H_{\lambda, 0}, P\right] R\left(\Phi\left(H_{\lambda, 0}\right)-\Phi\left(H_{\lambda, 1}\right)\right)$, so that the above argument can be repeated to obtain the same bound for $\left|\Delta_{1}\right|$. We conclude that $\left|\tilde{\sigma}_{I}\left(H_{\lambda, 0}\right)-\tilde{\sigma}_{I}\left(H_{\lambda, 1}\right)\right| \leq C \lambda^{N}$, and the proof is complete.

### 4.3 Applications

We apply the main results from Sections 4.1 and 4.2 to the Hamiltonians from Section 2.3 that are differential operators. We have already verified that these systems satisfy (H1), thus the only thing left to check is the anti-commutation relations of the leading order terms (4.0.2).
$2 \times 2$ Dirac system (2.3.1). We have $H=D_{x} \sigma_{1}+D_{y} \sigma_{2}+m(y) \sigma_{3}$, meaning that we can take $M_{0}=\sigma_{2}, M_{1}=\sigma_{1}$, and $a_{00}(y)=m(y) \sigma_{3}$. Since the $\sigma_{j}$ are nonsingular and $\left\{\sigma_{1}, \sigma_{2}\right\}=0$, (4.0.2) holds. It follows that $\left|\tilde{\sigma}_{I}\left(H_{\lambda}\right)-\sigma_{I}(H)\right| \leq C_{p} \lambda^{p}$ as $\lambda \rightarrow 0$, for any p. Moreover, $\left|\tilde{\sigma}_{I}\left(H_{\lambda}\right)-\tilde{\sigma}_{I}\left(H_{\lambda}+V_{\lambda}\right)\right| \leq C_{p} \lambda^{p}$ for all smooth Hermitian-valued point-wise multiplication operators $V_{\lambda}=V_{\lambda}(x, y)$ such that $\left\|V_{\lambda}\right\|_{L^{\infty}} \leq C$ uniformly in $\lambda$, and that satisfy $S_{1} \cap \operatorname{supp}(1-Q)=\emptyset$ for some closed set $S_{1}$ containing $\cup_{\lambda} \operatorname{supp}\left(V_{\lambda}\right)$.
$p$-wave superconductor model (2.3.2). We can take $M_{0}=M_{2}=\frac{1}{2 m} \sigma_{1}$, which takes care of all leading order terms. Since $\sigma_{1}$ is non-singular, (4.0.2) holds. It follows that $\left|\tilde{\sigma}_{I}\left(H_{\lambda}\right)-\sigma_{I}(H)\right| \leq C_{p} \lambda^{p}$ for any $p$. We also have $\left|\tilde{\sigma}_{I}\left(H_{\lambda}\right)-\tilde{\sigma}_{I}\left(H_{\lambda}+V_{\lambda}\right)\right| \leq C_{p} \lambda^{p}$ for all Hermitian-valued first-order differential operators $V_{\lambda}:=\sum_{i+j \leq 1} \lambda^{i+j} v_{\lambda, i j}(x, y) D_{x}^{i} D_{y}^{j}$ with smooth coefficients $v_{\lambda, i j}$ such that $\sum_{i, j}\left\|v_{\lambda, i j}\right\|_{L^{\infty}} \leq C$ and $\cup_{i, j} \operatorname{supp}\left(v_{\lambda, i j}\right) \subset S_{1}$ uniformly in $\lambda$, for some closed set $S_{1}$ satisfying $S_{1} \cap \operatorname{supp}(1-Q)=\emptyset$.
$d$-wave superconductor model (2.3.3). Take $M_{0}=M_{2}=\frac{1}{2 m} \sigma_{1}, a_{0}=c_{0} \sigma_{2}, a_{1}(y)=$
$c(y) \sigma_{3}, a_{2}=-c_{0} \sigma_{2}$. Since $\sigma_{1}$ is nonsingular and $\left\{\sigma_{1}, \sigma_{j}\right\}=0$ for $j \in\{2,3\}$, (4.0.2) holds. The following conclusions are the same as those for the $p$-wave superconductor model above.

### 4.4 Numerical simulations

To illustrate the stability (or not) of interface conductivities, we present numerical simulations of the $2 \times 2$ Dirac and $p$-wave superconductor models from section 2.3 , as well as models of equatorial waves analyzed in section 2.3 (see also [10]). We discretize a grid of size $L_{x} \times L_{y}=24 \times 24$ uniformly into $N_{x} \times N_{y}=64 \times 64$ grid points and impose periodic boundary conditions. Point-wise multiplication operators are represented as diagonal matrices, and derivatives $\tilde{D}_{\alpha} \approx \frac{1}{i} \partial_{\alpha}(\alpha=x, y)$ are computed in the Fourier basis $\tilde{D}_{\alpha}=F_{\alpha}^{-1} \Lambda_{\alpha} F_{\alpha}$, where $F_{\alpha}$ is the discrete Fourier transform with respect to direction $\alpha$, and

$$
\Lambda_{\alpha}=\frac{2 \pi}{N_{\alpha} L_{\alpha}} \operatorname{diag}\left(-\frac{N_{\alpha}}{2},-\frac{N_{\alpha}}{2}+1, \ldots, \frac{N_{\alpha}}{2}-1\right) .
$$

Note that $F_{\alpha}$ is unitary, so that $F_{\alpha}^{-1}=F_{\alpha}^{*}$. We refer the reader to [84] for more details on spectral methods and their accuracy.

As in section 4.1, we define the conductivity by

$$
\tilde{\sigma}_{I}(H):=\operatorname{Tr} i Q[H, P] \varphi^{\prime}(H)
$$

where $P=P(x)$ and $Q=Q(x, y)=Q_{X}(x) Q_{Y}(y)$ are point-wise multiplication operators satisfying

$$
P(x)=\left\{\begin{array}{ll}
1, & \delta_{x} \leq x \leq 3 L_{x} / 8 \\
0, & -3 L_{x} / 8 \leq x \leq-\delta_{x}
\end{array} \quad \text { and } \quad Q_{\Theta}(\theta)= \begin{cases}1, & |\theta| \leq L_{\theta} / 4 \\
0, & |\theta| \geq L_{\theta} / 4+\delta_{\theta}\end{cases}\right.
$$

for some $0<\delta_{\theta} \ll L_{\theta} / 4$. The Hamiltonians are periodized so that the $y$ dependent terms


Figure 4.1: Eigenfunctions of the periodic $2 \times 2$ Dirac Hamiltonian (top) and $p$-wave superconductor (bottom) models. First three columns: edge states localized at the $y=0$ and $y= \pm L_{y} / 2$ interfaces; far right: bulk states that do not contribute to the conductivity. The plotted vector components and corresponding eigenvalues are labeled.
( $m(y)$ for the $2 \times 2$ Dirac system and $c(y)$ for the $p$-wave superconductor) are equal to 1 whenever $\delta^{\prime} \leq y \leq L_{y} / 2-\delta^{\prime}$ and -1 for $-L_{y} / 2+\delta^{\prime} \leq y \leq-\delta^{\prime}$, and are smoothly connected in between. Here, $\delta^{\prime}$ is a positive constant that is small relative to $L_{y}$, but large enough for the interval $\left(-\delta^{\prime}, \delta^{\prime}\right)$ to contain at least 3 grid points. The eigenvalues of $H$ in the support of $\varphi^{\prime}$ (and the corresponding eigenvectors) are computed using the "eigs" command in MATLAB. Since the number of these eigenvalues is typically much smaller than the dimension of $H$, we avoid the computational expense of a full spectral decomposition.

For the $2 \times 2$ models, we consider perturbations of the form

$$
V(x, y)=(\hat{n} \cdot \vec{\sigma}) v(x, y), \quad v(x, y)=\left\{\begin{array}{l}
r \exp \left(-a^{2} /\left(a^{2}-x^{2}-y^{2}\right)\right), \quad x^{2}+y^{2}<a^{2} \\
0, \quad \text { else }
\end{array}\right.
$$

where $\hat{n} \in \mathbb{R}^{4}$ is a unit normal vector and $\vec{\sigma}=\left(\sigma_{0}, \sigma_{1}, \sigma_{2}, \sigma_{3}\right)$. Here, the $\sigma_{j}$ are the Pauli matrices (with $\sigma_{0}$ the $2 \times 2$ identity matrix). We observe that $\tilde{\sigma}_{I}$ is stable with respect to such perturbations. Namely, if we fix the support of $v$ to lie well within $\operatorname{supp}(Q)$ while increasing the amplitude of that perturbation, the conductivity does not change by much (even if the


Figure 4.2: Perturbed eigenfunctions of the $2 \times 2$ Dirac (top) and $p$-wave superconductor (bottom) models. For all plots, $r=2$ and $a=7.5$. We see a mix of propagating and evanescent modes. Despite the qualitative differences between the perturbed and unperturbed eigenfunctions, the conductivity remains approximately the same.
amplitude is very large). Similarly, for a fixed amplitude, the conductivity remains close to its original value until $v$ becomes highly delocalized. See Figure 4.3 (left), where we have plotted the dependence of the conductivity on the perturbation strength and localization for the $2 \times 2$ Dirac and $p$-wave superconductor models. Note that for many other choices of parameters, the conductivity remains within $2 \%$ of its unperturbed value for values of $r$ and $a$ much larger than 10 . The empirical stability of $\tilde{\sigma}_{I}$ is entirely consistent with the theoretical results from earlier sections.

We present examples of computed eigenfunctions for the $2 \times 2$ Dirac and $p$-wave superconductor models in Figures 4.1 and 4.2. We observe that perturbations $V$ of the above form strongly alter the edge states. They are no longer a superposition of terms of the form $e^{i \xi x} \psi(y)$ and we even recognise localized modes (see top right of Figure 4.2) that do not contribute to the conductivity. Thus the robustness of $\tilde{\sigma}_{I}$ observed in Figure 4.3 occurs in spite of the instability of the eigenfunctions with respect to perturbations.

One may also consider filters $Q_{\Delta_{X}, \Delta_{Y}}(x, y)=Q_{\Delta_{X}}(x) Q_{\Delta_{Y}}(y)$ with different centers,


Figure 4.3: The left panel demonstrates the numerical stability of $\tilde{\sigma}_{I}$ for the $2 \times 2$ Dirac (solid line) and $p$-wave superconductor (dashed line) models. For the solid line, we fixed $a=7.5$ (with $r$ varying) and the dashed line corresponds to $r=10$ (with $a$ varying). The center panel shows the conductivity for the $2 \times 2$ Dirac (solid line) and $p$-wave superconductor (dashed line) models as a function of the center of $Q_{Y}$. As expected, we get $-1,-2$ when the filter selects the increasing domain wall and 1,2 when the filter selects the decreasing domain wall, with a sharp transition in between. The right panel plots the conductivity for the $3 \times 3$ model as a function of perturbation strength, for perturbations in four distinct matrix elements. The conductivity is stable with respect to perturbations of the other five matrix elements; see text.
where

$$
Q_{\Delta_{\Theta}}(\theta)= \begin{cases}1, & \theta \in\left[-\frac{L_{\theta}}{4}-\Delta_{\Theta}, \frac{L_{\theta}}{4}-\Delta_{\Theta}\right] \\ 0, & \theta \notin\left(-\frac{L_{\theta}}{4}-\Delta_{\Theta}-\delta_{\theta}, \frac{L_{\theta}}{4}-\Delta_{\Theta}+\delta_{\theta}\right)\end{cases}
$$

When $\Delta_{Y}=L_{y} / 2, Q$ would instead test the conductivity associated to the opposite, spurious, domain wall. As the center $\Delta_{Y}$ of the filter increases, we expect the numerical conductivity $\sigma_{I}$ to shift from the conductivity of one domain wall to the that of the next domain wall. This is confirmed by the plot in Figure 4.3 (center) obtained for the $2 \times 2$ Dirac and $p$-wave superconductor models. Translating the center of $Q$ in the $x$-direction would have the same effect, as the sign of $P^{\prime}$ on the support of $Q$ changes when $\Delta_{X}=L_{x} / 2$.

We finally carry out the above numerical simulations on the model of equatorial waves presented in section 2.3; see also [10]. The unperturbed Hamiltonian with symbol given in

$$
\begin{equation*}
H_{0}=\left(D_{x}, D_{y},-f(y)\right) \cdot \Gamma \tag{4.4.1}
\end{equation*}
$$

where $\Gamma=\left(\gamma_{1}, \gamma_{4}, \gamma_{7}\right)$ with the $\gamma_{i}$ denoting the $(3 \times 3)$ Gell-Mann matrices. The Coriolis force is given by $f(y)$, which changes signs across the equator. As shown in [10], the conductivity of such a system depends on the choice of $f$ when $\mu=0$ whereas the theory presented in section 2.3 shows that the conductivity equals 2 as soon as $\mu \neq 0$. For example, $2 \pi \sigma_{I}=1$ if $f=f_{0} \operatorname{sgn}(y)$, and $2 \pi \sigma_{I}=2$ if $f(y)=\beta y$, with $f_{0}, \beta>0$.

Moreover, it was shown in [10] that the conductivity is stable under perturbations of the form $V=\operatorname{diag}\left(V_{11}, 0,0\right)$, but not $V=\operatorname{diag}\left(0, V_{22}, V_{33}\right)$. We have verified these results numerically for $f(y)$ a smooth periodic function satisfying the same conditions as $m(y)$ and $c(y)$ above. For example, we see that if $V_{0}=5 g_{L_{y} / 4}(y)$ (where $g_{\sigma}(y)$ is the pdf of a Gaussian with mean zero and standard deviation $\sigma$ ), then $2 \pi \tilde{\sigma}_{I}=1.9950$ if $V=0,2 \pi \tilde{\sigma}_{I}=1.9941$ if $V=\operatorname{diag}\left(V_{0}, 0,0\right), 2 \pi \tilde{\sigma}_{I}=-0.5394$ if $V=\operatorname{diag}\left(0, V_{0}, 0\right)$, and $2 \pi \tilde{\sigma}_{I}=0.1430$ if $V=$ $\operatorname{diag}\left(0,0, V_{0}\right)$. More generally, we have verified numerically that $\tilde{\sigma}_{I}$ is stable, as predicted by theory, under perturbations of the form $V(y)$ times any of the matrices

$$
\left(\begin{array}{lll}
1 & 0 & 0  \tag{4.4.2}\\
0 & 0 & 0 \\
0 & 0 & 0
\end{array}\right),\left(\begin{array}{lll}
0 & 1 & 0 \\
1 & 0 & 0 \\
0 & 0 & 0
\end{array}\right),\left(\begin{array}{lll}
0 & 0 & 1 \\
0 & 0 & 0 \\
1 & 0 & 0
\end{array}\right),\left(\begin{array}{ccc}
0 & i & 0 \\
-i & 0 & 0 \\
0 & 0 & 0
\end{array}\right),\left(\begin{array}{ccc}
0 & 0 & i \\
0 & 0 & 0 \\
-i & 0 & 0
\end{array}\right),\left(\begin{array}{ccc}
0 & 0 & 0 \\
0 & 0 & i \\
0 & -i & 0
\end{array}\right),
$$

and unstable under all other Hermitian perturbations, see Figure 4.3 (right).
Given that $H$ is translationally invariant in $x$, we can approximate its continuous branches of spectrum by computing the eigenvalues of $\hat{H}[\xi]:=\left(\xi, D_{y},-f(y)\right) \cdot \Gamma$ as a function of

$$
\xi \in\left\{\frac{2 \pi j}{N_{x}}: j \in\left\{-\frac{N_{x}}{2},-\frac{N_{x}}{2}+1, \ldots, \frac{N_{x}}{2}-1\right\}\right\},
$$



Figure 4.4: Branches of continuous spectrum for the $3 \times 3$ equatorial wave Hamiltonian, for different values of regularization parameter $\mu$. The top line corresponds to $f(y)=\operatorname{sgn}(y)$ while the bottom line corresponds to $f(y)=\tanh (\beta y)$. The solid curves represent the nontrivial increasing branches (and the two flat bands at $\pm 1$ when $\mu=0$ and $f(y)=\operatorname{sgn}(y)$ ). For $\mu=0$, we omit many eigenvalues approximately equal to 0 , as they correspond to essential spectrum for the continuous problem. When $\mu>0$ (resp. $\mu<0$ ), these eigenvalues populate the region $\{E(\xi)>0\}$ (resp. $\{E(\xi)<0\}$ ), making it difficult to identify branches of spectrum there.
see Figure $4.4(\mu=0)$. Consistent with theory [10], the number of branches passing through $E= \pm 1 / 2$ depends on the profile of $f$. Namely, we obtain two nondecreasing branches of continuous spectrum passing through $E=1 / 2$ when using $f(y)=\tanh (\beta y)$ and only one with $f(y)=\operatorname{sgn}(y)$. (The qualitative behavior of the branches is independent of the above perturbations under which $\tilde{\sigma_{I}}$ is stable.) Here, $\beta>0$ is sufficiently small so that the transition of $f(y)$ from values near -1 to values near 1 occurs over at least several grid points. Of course, $f(y)$ is periodically wrapped, so that it in fact equals $-\tanh \left(\beta\left(y \mp L_{y} / 2\right)\right)$ in the vicinity of $y= \pm L_{y} / 2$.

Given [10, Theorem 2.1 and Appendix B] and the fact that $\sigma_{I}(H):=\operatorname{Tr} i[H, P] \varphi^{\prime}(H)=0$ for all matrices $H$ and $P$, we would expect the number of signed crossings of $E=1 / 2$ to
be 0 . That is, every nondecreasing branch of spectrum passing through $E=1 / 2$ should be accompanied by a nonincreasing branch that also passes through $E=1 / 2$, as we observe. But these nonincreasing branches correspond to eigenfunctions that localize at $y= \pm L_{y} / 2$ and thus become insignificant when we apply $Q$ to compute $\tilde{\sigma}_{I}$.

The plots in Figure 4.4 with $\mu \neq 0$ correspond to the regularized model from Section 2.3. As expected, we find two nondecreasing branches of spectrum passing through the energy interval of interest when $f(y)$ is smooth or $\mu \neq 0$. Again, the qualitative behavior of the branches is robust to perturbations of the form (4.4.2). We see that the branches corresponding to $f(y)=\operatorname{sgn}(y)$, particularly $E(\xi)= \pm 1$ for $\mu=0$, are more sensitive to variations in $\mu$ than those corresponding to smooth $f$. The flat bands are eliminated when $\mu \neq 0$, resulting in two non-trivial increasing branches of spectrum.

## CHAPTER 5

## INTEGRAL FORMULATION OF KLEIN-GORDON SINGULAR WAVEGUIDES

The time-harmonic Klein-Gordon equation,

$$
-\Delta u+m^{2} u=E^{2} u
$$

arises naturally in a wide variety of contexts, including condensed matter and particle physics, classical mechanics, and optics. When $|E|<m$ it models an insulating medium: solutions decay exponentially quickly away from a source. In the last forty years there has been particular interest in the case in which two different insulators are brought together, meeting at an interface. In such situations, depending on the physical parameters, it is possible to generate surface waves which are localized near, and propagate along, the interface.

In two dimensions, this can be modelled by the following set of partial differential equations (PDEs)

$$
\begin{cases}-\Delta u(x)+m_{2}^{2} u(x)-E^{2} u(x)=f_{2}(x), & x \in \Omega_{2}  \tag{5.0.1}\\ -\Delta u(x)+m_{1}^{2} u(x)-E^{2} u(x)=f_{1}(x), & x \in \Omega_{1} \\ \lim _{y \rightarrow x \in \Omega_{2}} u(y)=\lim _{y \rightarrow x \in \Omega_{1}} u(y), & x \in \Gamma \\ \lim _{y \rightarrow x \in \Omega_{2}} \hat{n}(x) \cdot \nabla u(y)-\lim _{y \rightarrow x \in \Omega_{1}} \hat{n}(x) \cdot \nabla u(y)=-\left(m_{1}+m_{2}\right) u(x), & x \in \Gamma\end{cases}
$$

where $\hat{n}(x)$ denotes the unit normal to $\Gamma$ at $x \in \Gamma$ pointing in the direction of $\Omega_{2}$, the domains $\Omega_{1}$ and $\Omega_{2}$ denote the supports of the first and second insulators, respectively, $m_{1}$ and $m_{2}$ we refer to as their 'masses', $E$ is an energy, and $f_{1}$ and $f_{2}$ are source terms. In the sequel we will always assume that $|E|^{2}<m_{1}^{2}, m_{2}^{2}$, that $\overline{\Omega_{1} \cup \Omega_{2}}=\mathbb{R}^{2}$ is the entire plane,
and that $\Omega_{1}$ and $\Omega_{2}$ meet along an interface $\Gamma=\partial \Omega_{1}=\partial \Omega_{2}$. Moreover, we assume that $\Gamma$ is a single smooth simple curve which is asymptotically flat in both directions and has a positive opening angle at infinity (see Section 5.1.2 for precise definitions and more detailed discussions). Intuitively, we require $\Gamma$ has a smooth parameterization $\gamma: \mathbb{R} \rightarrow \Gamma$ which asymptotically approach two different rays as $t$ goes to $-\infty$ and $\infty$. In order to have unique solutions, one should also supplement these equations with suitable boundary conditions at infinity and radiation boundary conditions along the interface.

In this section we construct a novel system of boundary integral equations for solving the above equation, establishing bounded invertibility for a range of masses and interfaces (see Section 5.2). Our approach is based on introducing an auxiliary variable defined via the fundamental solution of a certain time-harmonic wave equation on the interface $\Gamma$. This formulation easily lends itself to implementation. In Section 5.3 we describe an algorithm based on our boundary integral equations, and in Section 5.4 present several numerical examples.

This section focuses on solving a source problem for singular Schrödinger equations with outgoing radiation conditions for an energy range $E \in\left(-m_{0}, m_{0}\right)$. A complete spectral analysis of the problem seems quite challenging. In a subsequent work, we plan to extend the current analysis of Dirac operators and more complex geometries for interfaces.

### 5.1 Mathematical preliminaries

### 5.1.1 Detailed formulation of the problem

In this section we give a more precise statement of the problem under consideration, and summarize the associated conditions on the interface. Towards that end, suppose we are given a smooth simple curve $\Gamma$ separating the plane into a lower region $\Omega_{1}$ and an upper region $\Omega_{2}$. Let $\gamma: \mathbb{R} \rightarrow \mathbb{R}^{2}$ be an arclength parameterization. Moreover, with $\hat{n}(t)$ the normal
vector to $\gamma$ at $t \in \mathbb{R}$ pointing in the direction of $\Omega_{2}$ (using the same notation $\hat{n}$ for $\hat{n}(t)$ and $\hat{n}(\gamma(t))$ to simplify), we assume that $\left(\gamma^{\prime}(t), \hat{n}(t)\right)$ has positive orientation. For concreteness, we additionally assume that:

$$
\begin{align*}
\left|\gamma^{\prime}(t)\right|=1, \quad\left|\gamma^{\prime \prime}(t)\right|^{2}+\left|\gamma^{\prime \prime \prime}(t)\right| \leq C_{0} e^{-\beta|t|}  \tag{5.1.1}\\
\lim _{t \rightarrow \infty}|\gamma( \pm t)|=\infty, \quad \lim _{t \rightarrow \infty}|\gamma(t)-\gamma(-t)|=\infty \tag{5.1.2}
\end{align*}
$$

where $C_{0}$, and $\beta$ are positive real constants. The fact that $\gamma$ is one-to-one (along with assumption (5.1.2) above) implies the existence of some $c>0$ such that

$$
\begin{equation*}
\frac{|\gamma(t)-\gamma(s)|}{|t-s|} \geq c, \quad s, t \in \mathbb{R} \tag{5.1.3}
\end{equation*}
$$

Given a suitably-differentiable function $u$, for ease of exposition we set

$$
\begin{aligned}
{[[\hat{n} \cdot \nabla u]](t) } & :=\lim _{s \rightarrow 0^{+}} \hat{n}(t) \cdot \nabla[u(\gamma(t)+s \hat{n}(t))-u(\gamma(t)-s \hat{n}(t))] \\
{[[u]](t) } & :=\lim _{s \rightarrow 0^{+}}[u(\gamma(t)+s \hat{n}(t))-u(\gamma(t)-s \hat{n}(t))]
\end{aligned}
$$

Finally, we suppose that we are given positive real numbers $m_{1}, m_{2}$, as well as a real number $E$ such that $|E|<\min \left(m_{1}, m_{2}\right)$. In the following we set $\omega_{j}=\sqrt{m_{j}^{2}-E^{2}}$.


Figure 5.1: Geometry

In this paper, we consider the time-harmonic Klein Gordon equation with piecewise
discontinuous masses meeting at a single one-dimensional interface:

$$
\begin{array}{ll}
-\Delta u(x)+\omega_{2}^{2} u(x)=f_{2}(x), & x \in \Omega_{2},  \tag{5.1.4}\\
-\Delta u(x)+\omega_{1}^{2} u(x)=f_{1}(x), & x \in \Omega_{1} .
\end{array}
$$

Here the functions $f_{1}$ and $f_{2}$ correspond to compactly-supported sources in the lower and upper regions, respectively. For ease of exposition in the following we will denote by $f$ the function which is equal to $f_{2}$ in $\Omega_{2}$ and to $f_{1}$ in $\Omega_{1}$. Along the interface we enforce continuity of $u$ as well as a jump condition in the normal derivative, as below

$$
\begin{align*}
{[[\hat{n} \cdot \nabla u]](\gamma(t)) } & =-\left(m_{2}+m_{1}\right) u(\gamma(t)), & & t \in \mathbb{R}  \tag{5.1.5}\\
{[[u]](\gamma(t)) } & =0, & & t \in \mathbb{R}
\end{align*}
$$

Remark 5.1.1. The continuity condition is relatively standard. The jump condition in the normal derivative is possibly less so. Our principle motivation comes from the consideration of topological insulators, as is briefly discussed in Section 5.1.2. Boundary conditions such as this also arise in the study of "leaky" waveguides, particularly in the context of "leaky quantum waveguides" (see [43] and the references therein).

With the above assumptions, we will show that there exist solutions $u$ of the PDE which propagate along $\Gamma$. These solutions can be interpreted as consisting in part of a surface wave emanating from the sources $f_{1}$ and $f_{2}$, confined in an exponential neighborhood of $\Gamma$. To enforce the condition that the surface wave should travel outwards (i.e. no energy should come in from infinity) we impose additional radiation conditions,

$$
\begin{align*}
\lim _{t \rightarrow \pm \infty}\left( \pm \partial_{t}-i E\right) u(\gamma(t)+r \hat{n}(t)) & =0, \quad r \neq 0  \tag{5.1.6}\\
\lim _{2} u(x) & =0, \tag{5.1.7}
\end{align*}
$$

where $d(x, \Gamma):=\min \{|x-y|: y \in \Gamma\}$ is the distance between $x$ and $\Gamma$. The requirement
(5.1.6) is known as an outgoing radiation condition, and intuitively means that $u(\gamma(t)+r \hat{n}(t))$ ought to look like $C e^{i E|t|}$ when $|t|$ is large (where $C \in \mathbb{R}$ is some constant). The outward propagation of $u$ can be seen from the corresponding solution of the Schrödinger equation, which is of the form $e^{i E(|t|-s)}$, where $s$ represents the time variable. Indeed, this solution is a function of $t-s$ (resp. $t+s$ ) when $t>0$ (resp. $t<0$ ). We refer to [39] for more details on this topic in similar settings.

Our main objective in this paper is the analysis of the problem (5.1.4) with boundary conditions (5.1.5, 5.1.6, 5.1.7).

### 5.1.2 Relation to Dirac equations and topological insulators

The above equations are closely related to certain Dirac equations arising in the study of topological insulators. In this section we elaborate on this connection in more detail. We begin by recalling that in two dimensions, the time-harmonic Dirac equation is given by

$$
\begin{equation*}
-i \sigma_{3} \partial_{x} \psi-i \sigma_{1} \partial_{y} \psi+m \sigma_{2} \psi=E \psi \tag{5.1.8}
\end{equation*}
$$

where $\psi: \mathbb{R}^{2} \rightarrow \mathbb{C}^{2}, m: \mathbb{R}^{2} \rightarrow \mathbb{R}$, and $\sigma_{1}, \sigma_{2}$, and $\sigma_{3}$ are the Pauli spin matrices defined by

$$
\sigma_{1}=\left(\begin{array}{cc}
0 & 1 \\
1 & 0
\end{array}\right), \quad \sigma_{2}=\left(\begin{array}{cc}
0 & -i \\
i & 0
\end{array}\right), \quad \text { and } \quad \sigma_{3}=\left(\begin{array}{cc}
1 & 0 \\
0 & -1
\end{array}\right)
$$

Dirac equations are natural models in the description of topological insulators and the transport observed at an interface separating two insulators in different phases [17, 10, 11, 14]. In particular, if we assume that two insulators are brought together along an interface $\{y=0\}$, we set $m(x, y)=m_{2}$ for $y>0$ and $m(x, y)=-m_{1}$ for $y<0$. While each region on its own acts as an insulator provided $|E|<\min \left(\left|m_{1}\right|,\left|m_{2}\right|\right)=: m_{*}$, this is no longer the case for the combination, which admits absolutely continuous spectra in $\left(-m_{*}, m_{*}\right)$ when the signs of
$m_{1}$ and $m_{2}$ are the same.
Squaring the Dirac equation (5.1.8) one obtains

$$
-\partial_{x}^{2} \psi-\partial_{y}^{2} \psi+\left[m_{2}^{2} \theta(y)+m_{1}^{2} \theta(-y)\right] \psi-E^{2} \psi+\left(m_{2}+m_{1}\right) \delta_{0}(y) \sigma_{3} \psi=0
$$

Here $\theta(y)$ is the Heaviside function, equal to 1 for $y>0$ and to 0 for $y \leq 0$. The above equation is diagonal, i.e. the components of $\psi$ are not coupled. The equation for the first component is

$$
-\Delta \psi_{1}+\left[m_{2}^{2} \theta(y)+m_{1}^{2} \theta(-y)\right] \psi_{1}-E^{2} \psi_{1}+\left(m_{2}+m_{1}\right) \delta_{0}(y) \psi_{1}=0
$$

which admits only the trivial solution $\psi_{1}=0$ (provided $\psi_{1}$ is not exponentially increasing in $|y|)$. The more interesting second component satisfies

$$
-\Delta \psi_{2}+\left[m_{2}^{2} \theta(y)+m_{1}^{2} \theta(-y)\right] \psi_{2}-E^{2} \psi_{2}-\left(m_{2}+m_{1}\right) \delta_{0}(y) \psi_{2}=0
$$

which is exactly $(5.1 .4,5.1 .5)$, at least when sources are neglected.
Our main objective in this paper is to analyze such a scalar equation and focus on the resulting propagation of signals along the interface $\Gamma$ separating the insulators. In particular, we are interested in the setting where $\Gamma$ is curved. The corresponding analysis to the vectorvalued Dirac equation is postponed to a later study.

For an analysis of the temporal propagation of wavepackets along a curved interface in the semiclassical regime (i.e, for wavepackets asymptotically localized in the near vicinity of the interface) in both topologically trivial (Klein-Gordon) and non-trivial (Dirac) settings, see $[9,12]$.

### 5.1.3 Boundary integral operators and their properties

In this section we introduce several frequently-encountered boundary integral operators which will be useful in defining the boundary integral equations for the solution of PDE (5.1.4). We begin by recalling that for any $\omega$ in the right-half of the complex plane, the Green's function $G_{\omega}(x, y)$ for the PDE

$$
\begin{align*}
-\Delta u(x)+\omega^{2} u(x) & =\delta(x-y) \\
\lim _{|x| \rightarrow \infty} u(x) & =0 \tag{5.1.9}
\end{align*}
$$

is given by

$$
\begin{equation*}
G_{\omega}(x, y)=\frac{1}{2 \pi} K_{0}(\omega|x-y|) \tag{5.1.10}
\end{equation*}
$$

where $K_{0}$ is the modified Bessel function of the second kind.
Given a function $\rho \in L^{2}(\mathbb{R})$ we define its single-layer potential $S_{\omega}[\rho]$ by

$$
S_{\omega}[\rho](x)=\int_{\mathbb{R}} G_{\omega}(x, \gamma(t)) \rho(t) \mathrm{d} t
$$

and its double-layer potential $D_{\omega}[\rho]$ by

$$
D_{\omega}[\rho](x)=-\int_{\mathbb{R}} \hat{n}(t) \cdot \nabla_{x} G_{\omega}(x, \gamma(t)) \rho(t) \mathrm{d} t
$$

for $x \notin \Gamma$. Note that standard definitions of the above operators usually contain a factor of $\left|\gamma^{\prime}(t)\right|$ in the integrand, while we assume that $\left|\gamma^{\prime}\right| \equiv 1$. It is well-known (see [54, Lemmas 3.3 and 3.5 ] for example) that $S_{\omega}$ is continuous across $\Gamma$ while $D_{\omega}$ and $\hat{n} \cdot \nabla S_{\omega}$ satisfy the
following jump relations

$$
\begin{align*}
\lim _{s \rightarrow 0^{+}} D_{\omega}[\rho](t \pm s \hat{n}(t)) & = \pm \frac{1}{2} \rho(t)-\int_{\mathbb{R}} \hat{n}\left(t^{\prime}\right) \cdot \nabla G_{\omega}\left(\gamma(t), \gamma\left(t^{\prime}\right)\right) \rho\left(t^{\prime}\right) \mathrm{d} t^{\prime}  \tag{5.1.11}\\
\lim _{s \rightarrow 0^{+}} \hat{n}(t) \cdot \nabla S_{\omega}[\rho](t \pm s \hat{n}(t)) & =\mp \frac{1}{2} \rho(t)+\int_{\mathbb{R}} \hat{n}(t) \cdot \nabla G_{\omega}\left(\gamma(t), \gamma\left(t^{\prime}\right)\right) \rho\left(t^{\prime}\right) \mathrm{d} t^{\prime} \tag{5.1.12}
\end{align*}
$$

We note that the two integral operators appearing on the right-hand sides of the previous equations are compact (in $L^{2}(\mathbb{R})$ ), as their kernels are continuous and rapidly decaying in $t$ and $t^{\prime}$. With some abuse of notation, in the following we denote these operators by $\mathcal{D}_{\omega}$ and $\mathcal{S}_{\omega}^{\prime}$, respectively. We define $\mathcal{S}_{\omega}$ and $\mathcal{D}_{\omega}^{\prime}$ by

$$
\begin{aligned}
\mathcal{S}_{\omega}[\rho](t) & =\int_{\mathbb{R}} G_{\omega}\left(\gamma(t), \gamma\left(t^{\prime}\right)\right) \rho\left(t^{\prime}\right) \mathrm{d} t^{\prime} \\
\mathcal{D}_{\omega}^{\prime}[\rho](t) & =-\int_{\mathbb{R}} \hat{n}\left(t^{\prime}\right) \cdot \nabla^{2} G_{\omega}\left(\gamma(t), \gamma\left(t^{\prime}\right)\right) \hat{n}(t) \rho\left(t^{\prime}\right) \mathrm{d} t^{\prime}
\end{aligned}
$$

with $\nabla^{2} G_{\omega}$ the Hessian of $G_{\omega}$. We note that in this case, $\mathcal{S}_{\omega}, \mathcal{D}_{\omega}, \mathcal{S}_{\omega}^{\prime}$ and $\mathcal{D}_{\omega}^{\prime}$ can be viewed as operators from $L^{2}(\mathbb{R}) \rightarrow L^{2}(\mathbb{R})$. Finally, we remark that both $\mathcal{D}_{\omega}$ and $\mathcal{S}_{\omega}^{\prime}$ are zero when restricted to any portion of the boundary which is flat. Moreover, for flat interfaces the kernels of $\mathcal{S}_{\omega}$ and $\mathcal{D}_{\omega_{2}}^{\prime}-\mathcal{D}_{\omega_{1}}^{\prime}$ have the following Sommerfeld integral representations

$$
\begin{align*}
& \mathcal{S}_{\omega}[\rho](t)=\int_{\mathbb{R}} K_{\omega}\left(t-t^{\prime}\right) \rho\left(t^{\prime}\right) \mathrm{d} t^{\prime}  \tag{5.1.13}\\
& \left(\mathcal{D}_{\omega_{2}}^{\prime}-\mathcal{D}_{\omega_{1}}^{\prime}\right)[\rho](t)=\int_{\mathbb{R}} H_{\omega_{2}, \omega_{1}}\left(t-t^{\prime}\right) \rho\left(t^{\prime}\right) \mathrm{d} t^{\prime} \tag{5.1.14}
\end{align*}
$$

where

$$
\begin{align*}
& K_{\omega}(t)=\frac{1}{4 \pi} \int_{\mathbb{R}} \frac{e^{i \xi t}}{\sqrt{\xi^{2}+\omega^{2}}} \mathrm{~d} \xi  \tag{5.1.15}\\
& H_{\omega_{2}, \omega_{1}}(t)=\frac{1}{4 \pi} \int_{\mathbb{R}}\left(\sqrt{\xi^{2}+\omega_{2}^{2}}-\sqrt{\xi^{2}+\omega_{1}^{2}}\right) e^{i \xi t} \mathrm{~d} \xi \tag{5.1.16}
\end{align*}
$$

### 5.2 Analytical results

### 5.2.1 The BIEs

Let us now state the boundary integral formulation of the $\operatorname{PDE}(5.1 .4,5.1 .5,5.1 .6,5.1 .7)$.
We begin with the case $m_{1}=m_{2}=m$.
First, we define the operator $Q$ via the following formula

$$
\begin{equation*}
Q \rho(t)=\frac{m^{2}}{E} \int_{\mathbb{R}} e^{i E\left|t-t^{\prime}\right|} \rho\left(t^{\prime}\right) \mathrm{d} t^{\prime}, \quad \rho \in L^{2}(\mathbb{R}) \tag{5.2.1}
\end{equation*}
$$

and set $\mathcal{L}:=1-2 m \mathcal{S}_{\omega}$ and $\mathcal{P}:=1+Q$ so that

$$
\begin{align*}
& \mathcal{L} \rho(t)=\rho(t)-2 m \int_{\mathbb{R}} G_{\omega}\left(\gamma(t), \gamma\left(t^{\prime}\right)\right) \rho\left(t^{\prime}\right) \mathrm{d} t^{\prime} \\
& \mathcal{P} \rho(t)=\rho(t)+\frac{m^{2}}{E} \int_{\mathbb{R}} e^{i E\left|t-t^{\prime}\right|} \rho\left(t^{\prime}\right) \mathrm{d} t^{\prime} \tag{5.2.2}
\end{align*}
$$

for $\rho \in L^{2}(\mathbb{R})$.
Then, we seek a density $\rho: \mathbb{R} \rightarrow \mathbb{C}$, which satisfies the following boundary integral equation

$$
\begin{equation*}
\mathcal{L P} \rho=2 m u_{i}, \tag{5.2.3}
\end{equation*}
$$

where $u_{i}$ is given by

$$
\begin{equation*}
u_{i}(x):=\int_{\mathbb{R}^{2}} G_{\omega}(x, y) f(y) \mathrm{d} y \tag{5.2.4}
\end{equation*}
$$

Here, as above, $f$ denotes the function which is equal to $f_{2}$ in $\Omega_{2}$ and $f_{1}$ in $\Omega_{1}$.
The following theorem relates the solutions of (5.2.3) to the solutions of (5.1.4, 5.1.5, 5.1.6, 5.1.7).

Theorem 5.2.1. Suppose $m_{1}=m_{2}=$ : $m$. Let $\rho \in L^{1}$ be a solution of (5.2.3), and set

$$
\begin{equation*}
\mu:=\mathcal{P} \rho, \quad u_{s}:=S_{\omega}[\mu], \quad u:=u_{i}+u_{s} \tag{5.2.5}
\end{equation*}
$$

Then (5.1.4), (5.1.5), (5.1.6) and (5.1.7) hold. In particular, $u$ constructed in this way is a solution of the PDE.

Remark 5.2.2. The operator $Q$ can naturally be interpreted as the fundamental solution operator of the one-dimensional Helmholtz equation

$$
\Delta_{\Gamma} v+E^{2} v=m^{2} \rho,
$$

where $\Delta_{\Gamma}$ is the Laplace-Beltrami operator of the interface curve $\Gamma$.

Remark 5.2.3. Note that the operator $Q$ could have equally been defined by replacing $e^{i E\left|t-t^{\prime}\right|}$ by $e^{-i E\left|t-t^{\prime}\right|}$. The choice of sign in (5.2.1) amounts to a choice of outgoing radiation condition, namely that we filter out incoming radiation. Indeed, if $\rho \in L^{1}$, then $\lim _{t \rightarrow \pm \infty}\left(Q \rho(t)-\frac{m^{2}}{E} e^{i E|t|} \hat{\rho}( \pm E)\right)=0$, meaning that solutions of a time-dependent Klein Gordon problem would propagate to the right for $t>0$ and to the left for $t<0$ whenever $|t|$ is sufficiently large ( $t$ is the space, not time, variable here). See [39] for details on the notion of incoming and outgoing radiation.

Suppose now that $m_{1} \neq m_{2}$, and define $\bar{m}:=\frac{1}{2}\left(m_{1}+m_{2}\right)$. Analogous to (5.2.1), let $Q_{2}: L^{2}(\mathbb{R}) \rightarrow L^{2}(\mathbb{R})$ be given by

$$
\begin{equation*}
Q_{2} \rho(t)=\frac{\bar{m}^{2}}{E} \int_{\mathbb{R}} e^{i E\left|t-t^{\prime}\right|} \rho\left(t^{\prime}\right) \mathrm{d} t^{\prime}, \quad \rho \in L^{2}(\mathbb{R}) \tag{5.2.6}
\end{equation*}
$$

We define the operator $\mathcal{L}_{2}:\left(L^{2}(\mathbb{R})\right)^{2} \rightarrow\left(L^{2}(\mathbb{R})\right)^{2}$ by

$$
\mathcal{L}_{2}:=1-\left(\begin{array}{cc}
\mathcal{S}_{\omega_{2}}^{\prime}-\mathcal{S}_{\omega_{1}}^{\prime}+\bar{m}\left[\mathcal{S}_{\omega_{2}}+\mathcal{S}_{\omega_{1}}\right] & \bar{m}\left[\mathcal{D}_{\omega_{2}}+\mathcal{D}_{\omega_{1}}\right]+\mathcal{D}_{\omega_{2}}^{\prime}-\mathcal{D}_{\omega_{1}}^{\prime}  \tag{5.2.7}\\
-\left(\mathcal{S}_{\omega_{2}}-\mathcal{S}_{\omega_{1}}\right) & -\left(\mathcal{D}_{\omega_{2}}-\mathcal{D}_{\omega_{1}}\right)
\end{array}\right)
$$

and $\mathcal{P}_{2}:\left(L^{2}(\mathbb{R})\right)^{2} \rightarrow\left(L^{2}(\mathbb{R})\right)^{2}$ by

$$
\mathcal{P}_{2}:=\left(I+V\left(\begin{array}{cc}
Q_{2} & 0  \tag{5.2.8}\\
0 & 0
\end{array}\right) V^{-1}\right), \quad V:=\left(\begin{array}{cc}
-1 & \frac{1}{2 m_{2}}-\frac{1}{2 m_{1}} \\
\frac{1}{2 m_{2}}-\frac{1}{2 m_{1}} & 1
\end{array}\right)
$$

Our integral equation is then to find a density $\sigma \in\left(L^{2}(\mathbb{R})\right)^{2}$ such that

$$
\begin{equation*}
\mathcal{L}_{2} \mathcal{P}_{2} \sigma=r \tag{5.2.9}
\end{equation*}
$$

where

$$
\begin{align*}
r: & =\binom{\left.\left[\left[\hat{n} \cdot \nabla u_{i}\right]\right]+2 \bar{m} u_{i}\right)}{-\left[\left[u_{i}\right]\right]}, \\
u_{i}(x) & :=\left\{\begin{array}{cc}
\int_{\Omega_{2}} G_{\omega_{2}}(x, y) f_{2}(y) \mathrm{d} y, & x \in \Omega_{2} \\
\int_{\Omega_{1}} G_{\omega_{1}}(x, y) f_{1}(y) \mathrm{d} y, & x \in \Omega_{1} .
\end{array}\right. \tag{5.2.10}
\end{align*}
$$

We can then relate the solutions of (5.2.9) to solutions of (5.1.4, 5.1.5, 5.1.6, 5.1.7).

Theorem 5.2.4. Suppose $\sigma \in\left(L^{1}(\mathbb{R})\right)^{2}$ is a solution of (5.2.9), and set

$$
\binom{\mu}{\rho}:=\mathcal{P}_{2} \sigma, \quad u_{s}(x)= \begin{cases}D_{\omega_{2}}[\rho](x)+S_{\omega_{2}}[\mu](x), & x \in \Omega_{2}  \tag{5.2.11}\\ D_{\omega_{1}}[\rho](x)+S_{\omega_{1}}[\mu](x), & x \in \Omega_{1}\end{cases}
$$

Then $u:=u_{i}+u_{s}$ satisfies (5.1.4), (5.1.5), (5.1.6) and (5.1.7).

For proofs of Theorems 5.2.1 and 5.2.4, see Section 5.5.

As we will show in subsection 5.2.2 below, the integral equations (5.2.3) and (5.2.9) admit a unique solution when the interface is flat. The rest of this section concerns well-posedness of the boundary formulation for arbitrary interfaces satisfying assumptions (5.1.1) and (5.1.2). We will assume that $m_{1}=m_{2}=: m$ throughout. In principle, similar arguments should extend to the $m_{1} \neq m_{2}$ case, though the presentation would be more complicated due to the presence of terms involving double layer potentials and their derivatives.

Our main results are Theorems 5.2.5 and 5.2.7 below, which state that the integral formulation (5.2.3) is well posed for "almost all" choices of $m$ and $E$. The rapid decay of the curvature of $\gamma$ at infinity allows for tools from perturbation theory to be applied. To prove Theorem 5.2.5, we will compare the generic (curved-interface) integral operators with their flat-interface simplifications, and show that the operator norm of the difference is small in some asymptotic regime. The difference operator also plays a central role in the proof of Theorem 5.2.7, though a smallness condition is no longer required. To distinguish between the arbitrary and flat cases, let $\mathcal{L}_{0}: L^{2}(\mathbb{R}) \rightarrow L^{2}(\mathbb{R})$ be defined by

$$
\begin{equation*}
\mathcal{L}_{0} \mu(t):=\mu(t)-\frac{m}{2 \pi} \int_{\mathbb{R}} \int_{\mathbb{R}} \frac{e^{i \xi\left(t-t^{\prime}\right)}}{\sqrt{\xi^{2}+\omega^{2}}} \mathrm{~d} \xi \mu\left(t^{\prime}\right) \mathrm{d} t, \quad \mu \in L^{2}(\mathbb{R}) \tag{5.2.12}
\end{equation*}
$$

so that $\mathcal{L}_{0}$ is the operator $\mathcal{L}$ in (5.2.2) when the boundary is flat. Then (5.2.3) can be rewritten as $\left(\mathcal{L}_{0} \mathcal{P}+\left(\mathcal{L}-\mathcal{L}_{0}\right) \mathcal{P}\right) \rho=2 m u_{i}$, or equivalently

$$
\begin{equation*}
(1+\mathcal{M}) \rho=2 m\left(\mathcal{L}_{0} \mathcal{P}\right)^{-1} u_{i}, \quad \mathcal{M}:=\left(\mathcal{L}_{0} \mathcal{P}\right)^{-1}\left(\mathcal{L}-\mathcal{L}_{0}\right) \mathcal{P} \tag{5.2.13}
\end{equation*}
$$

Thus to prove well-posedness of the integral formulation, it suffices to show that the operator $1+\mathcal{M}$ has bounded inverse. As suggested above, one condition that would guarantee bounded invertibility is that $\mathcal{M}$ be bounded with operator norm strictly less than 1 , in which case $1+\mathcal{M}$ could be inverted by the Neumann series. But although $\left(\mathcal{L}{ }_{0} \mathcal{P}\right)^{-1}$ is bounded on $L^{2}$ and $\mathcal{L}-\mathcal{L}_{0}$ has a rapidly decaying kernel, $\mathcal{M}$ is not bounded on $L^{2}$ due to its right-most
factor of $\mathcal{P}$. We thus introduce the following weighted $L^{2}$ spaces on which $\mathcal{M}$ is not only bounded but compact (to be proved in Section 5.5; see Lemma 5.5.3). For $\alpha \in \mathbb{R}$, define $w_{\alpha}(t):=e^{\alpha|t|}$ and $L_{\alpha}^{2}:=\left\{\rho \in L^{2}(\mathbb{R}): w_{\alpha} \rho \in L^{2}(\mathbb{R})\right\}$. Define $\|\cdot\|_{L_{\alpha}^{2}}$ by $\|\rho\|_{L_{\alpha}^{2}}:=\left\|w_{\alpha} \rho\right\|_{L^{2}}$.

For the following, fix $m_{0}>0$ and $E_{0} \in\left(-m_{0}, m_{0}\right) \backslash\{0\}$, define $\omega_{0}:=\sqrt{m_{0}^{2}-E_{0}^{2}}$, and set $m=\lambda m_{0}$ and $E=\lambda E_{0}$ for $\lambda \in \mathbb{R}$. Let $\alpha_{*}>0$ such that $w_{\alpha_{*}}(\cdot) K_{0}\left(\omega_{0}|\cdot|\right) \in L^{1}$,

$$
\int_{\mathbb{R}^{2}}\left(e^{\alpha_{*} t} e^{\alpha_{*} s}|t-s|^{3} e^{-\beta \chi(s, t)} K_{1}\left(c \omega_{0}|t-s|\right)\right)^{2} \mathrm{~d} t \mathrm{~d} s<\infty, \quad \alpha_{*}<\sqrt{\frac{\sqrt{4 m_{0}^{4}+E_{0}^{4}}-E_{0}^{2}}{2}}
$$

$|\cdot|{ }^{3} K_{1}(c|\cdot|) \leq C e^{-\alpha_{*}|\cdot|}$ for some $C>0$, and $\alpha_{*}<\beta$, where $\beta$ and $c$ are defined by (5.1.1) and (5.1.3), and

$$
\chi(s, t):= \begin{cases}\min \{|s|,|t|\}, & s t>0 \\ 0, & \text { else }\end{cases}
$$

Here, the $K_{j}:(0, \infty) \rightarrow(0, \infty)$ are modified Bessel functions of the second kind. They are continuous, exponentially decaying at infinity, and satisfy $K_{0}(r) \sim-\log r$ and $K_{1}(r) \sim 1 / r$ as $r \downarrow 0$. Thus $\alpha_{*}$ is indeed well defined.

We now state our first well-posedness result. Its proof uses that $\mathcal{M}$ is small in operator norm for all $\lambda$ sufficiently large; see Section 5.5. Most of the theory from the proof would hold also on spaces with algebraic weights. However, the final step requires holomorphic continuation of $\lambda$ into the complex plane. When $\lambda$ (and hence $E$ ) has negative imaginary part, the kernel of $Q$ has exponential growth away from the diagonal. Thus to ensure that $Q$ is well defined, it is necessary to consider functions that decay exponentially at infinity.

Theorem 5.2.5. For any $\alpha \in\left(0, \alpha_{*}\right]$, the integral equation (5.2.3) admits a unique solution $\rho \in L_{\alpha}^{2}$ for all but a finite number of $\lambda \in[1, \infty)$.

Remark 5.2.6. We note that the above dependence of $m$ and $E$ on $\lambda$ is equivalent to setting
$m=m_{0}$ and $E=E_{0}$ while changing $\gamma$ by $\gamma_{\lambda}(t):=\lambda \gamma(t / \lambda)$. Thus increasing the value of $\lambda$ can be thought of as stretching out the interface.

We conclude the section with a second well-posedness result, which (drops the above $\lambda$-dependence and) states that for any fixed $m>0$, our integral formulation (5.2.3) has a unique solution for all but a countable number of values of $E \in(-m, m) \backslash\{0\}$, with 0 and $\pm m$ the only possible accumulation points of these undesirable $E$-values. The proof, which is postponed to Section 5.5 , uses the compactness of $\mathcal{M}$ but does not require its operator norm to be small.

Theorem 5.2.7. Fix $m>0$. For any $\varepsilon>0$ and $\alpha>0$ sufficiently small (depending on $\varepsilon$ ), the integral equation (5.2.3) admits a unique solution $\rho \in L_{\alpha}^{2}$ for all but a finite number of $E \in[-m+\varepsilon,-\varepsilon] \cup[\varepsilon, m-\varepsilon]$.

### 5.2.2 Intuitive derivation of the boundary integral equations

In this section we outline an intuitive derivation of the boundary integral equations, and sketch a proof of invertibility for the case of a flat interface. As before, we begin with the case $m_{1}=m_{2}$.

For equal masses, the equations (5.1.4) are

$$
\begin{equation*}
\Delta u(x)-\omega^{2} u(x)=f(x), \quad x \in \mathbb{R}^{2} \backslash \Gamma \tag{5.2.14}
\end{equation*}
$$

where $\omega:=\omega_{1}=\omega_{2}$.
Then if we define the functions $u_{i}$ and $u_{s}$ by

$$
\begin{equation*}
u_{i}(x):=\int_{\mathbb{R}^{2}} G_{\omega}(x, y) f(y) d y, \quad u_{s}:=u-u_{i} \tag{5.2.15}
\end{equation*}
$$

and assume the existence of a density $\mu \in L^{2}(\mathbb{R})$ such that $u_{s}(x)=S_{\omega}[\mu](x)$ for all $x \in \mathbb{R}^{2} \backslash \Gamma$,
it follows that

$$
\begin{equation*}
-\mu(t)+2 m \mathcal{S}_{\omega}[\mu](t)=-2 m u_{i}(\gamma(t)), \quad t \in \mathbb{R} \tag{5.2.16}
\end{equation*}
$$

Indeed, $u_{i}$ is smooth across $\Gamma$ and hence the definition of $u_{s}$ implies

$$
\left\{\begin{array}{l}
\Delta u_{s}(x)-\omega^{2} u_{s}(x)=0, \quad x \in \mathbb{R}^{2} \backslash \Gamma  \tag{5.2.17}\\
{\left[\left[\hat{n} \cdot \nabla u_{s}\right]\right]+2 m u_{s}=-2 m u_{i}, \quad x \in \Gamma}
\end{array}\right.
$$

with the added condition that $u_{s}+u_{i}$ is continuous across $\Gamma$. It is clear that the first line of (5.2.17) holds for any choice of $\mu$, while the boundary equation (5.2.16) is a consequence of the second line of $(5.2 .17)$ together with (5.1.12) and the definition of $\mu$.

We recognize the left-hand side of (5.2.16) as $-\mathcal{L} \mu$, where we recall the definition of $\mathcal{L}$ in (5.2.2). We will see below that for the flat-interface case, $\mathcal{L}$ has a continuous spectrum which passes through zero. As such, in general the solution will not be unique without imposing additional conditions. Moreover, if the domain is truncated, the resulting truncated operator will at best be poorly-conditioned in the limit as the length of the boundary tends to infinity and in general will not converge as the size of the truncated domain grows. For a flat interface, the problem is easily analyzed in the Fourier domain. Indeed, observing that

$$
\begin{equation*}
\mathcal{F}_{t \rightarrow \xi}\left\{\frac{1}{E} e^{i E|t|}\right\}=\lim _{\varepsilon \downarrow 0} \mathcal{F}_{t \rightarrow \xi}\left\{\frac{1}{E+i \varepsilon} e^{i(E+i \varepsilon)|t|}\right\}=-\frac{2 i}{\xi^{2}-E^{2}} \tag{5.2.18}
\end{equation*}
$$

we see that

$$
\begin{equation*}
\mathcal{F}\{\mathcal{L} \mathcal{P} \rho\}(\xi)=\left(1-\frac{m}{\sqrt{\xi^{2}+\omega^{2}}}\right)\left(1-\frac{2 i m^{2}}{\xi^{2}-E^{2}}\right) \tilde{\rho}(\xi)=: a(\xi) \tilde{\rho}(\xi) \tag{5.2.19}
\end{equation*}
$$

where $\mathcal{P}$ is defined in (5.2.2) and $\tilde{\rho}$ denotes the Fourier transform of $\rho$. Here, we used (5.1.15) to derive the Fourier representation of $\mathcal{L}$. The fact that $\mathcal{L P}$ is a point-wise multiplication
in the Fourier domain follows immediately from translation invariance of $\mathcal{L}$ and $\mathcal{P}$ (both kernels $k\left(t, t^{\prime}\right)$ are functions of only $\left.t-t^{\prime}\right)$. The singularities of the second factor of $a$ at $\xi= \pm E$ are canceled by the zeros of the first factor, making $a$ an analytic function. Since $a$ is nonzero for all $\xi$ and converges to 1 as $|\xi| \rightarrow \infty$, there exist constants $0<c<C$ such that $c \leq|a| \leq C$ uniformly in $\xi$.

We conclude that the operator $(\mathcal{L P})^{-1}$ is bounded on $L^{2}(\mathbb{R})$ with (bounded and analytic) Fourier symbol

$$
\begin{equation*}
a^{-1}=\left[\left(1-\frac{m}{\sqrt{\xi^{2}+\omega^{2}}}\right)\left(1-\frac{2 i m^{2}}{\xi^{2}-E^{2}}\right)\right]^{-1}=\left(1+\frac{m}{\sqrt{\xi^{2}+\omega^{2}}}\right)\left(1+\frac{(2 i+1) m^{2}}{\xi^{2}-E^{2}-2 i m^{2}}\right) . \tag{5.2.20}
\end{equation*}
$$

This means there exists a unique function $\rho \in L^{2}(\mathbb{R})$ such that $\mathcal{L P} \rho=2 m u_{i}$, and hence the boundary integral equation is invertible (with a bounded solution).

Suppose now $m_{1} \neq m_{2}$. This case is slightly more complicated, though the reasoning is similar to the previous case. Define the functions $u_{i}$ and $u_{s}$ by

$$
u_{i}(x)=\left\{\begin{array}{ll}
\int_{\Omega_{2}} G_{\omega_{2}}(x, y) f_{2}(y) \mathrm{d} y, & x \in \Omega_{2}  \tag{5.2.21}\\
\int_{\Omega_{1}} G_{\omega_{1}}(x, y) f_{1}(y) \mathrm{d} y, & x \in \Omega_{1}
\end{array}, \quad u_{s}:=u-u_{i}\right.
$$

Note that $u_{i}$ is no longer continuous across $\Gamma$. For notational convenience, on the boundary between $\Omega_{1}$ and $\Omega_{2}$ we set $u_{i}$ to be the average of the limits from above and below. If there are densities $\mu, \rho \in L^{2}(\mathbb{R})$ such that

$$
u_{s}(x)= \begin{cases}D_{\omega_{2}}[\rho](x)+S_{\omega_{2}}[\mu](x), & x \in \Omega_{2}  \tag{5.2.22}\\ D_{\omega_{1}}[\rho](x)+S_{\omega_{1}}[\mu](x), & x \in \Omega_{1}\end{cases}
$$

then we eventually arrive at the following linear system of equations for $\mu$ and $\rho$,

$$
\begin{align*}
& \rho(t)+\left[\mathcal{D}_{\omega_{2}}[\rho](t)-\mathcal{D}_{\omega_{1}}[\rho](t)\right]+\left[\mathcal{S}_{\omega_{2}}[\mu](t)-\mathcal{S}_{\omega_{1}}[\mu](t)\right]=-\left[\left[u_{i}\right]\right](\gamma(t))  \tag{5.2.23}\\
&-\mu(t)+\left[\mathcal{S}_{\omega_{2}}^{\prime}[\mu](t)-\mathcal{S}_{\omega_{1}}^{\prime}[\mu](t)\right]+\left.+\mathcal{D}_{\omega_{2}}^{\prime}[\rho](t)-\mathcal{D}_{\omega_{1}}^{\prime}[\rho](t)\right] \\
&+\bar{m}\left[\mathcal{S}_{\omega_{2}}[\mu](t)+\mathcal{S}_{\omega_{1}}[\mu](t)\right]+\bar{m}\left[\mathcal{D}_{\omega_{2}}[\rho](t)+\mathcal{D}_{\omega_{1}}[\rho](t)\right]  \tag{5.2.24}\\
&+2 \bar{m} u_{i}(\gamma(t))=-\left[\left[\hat{n} \cdot \nabla u_{i}\right]\right](\gamma(t))
\end{align*}
$$

which are understood to hold for all $t \in \mathbb{R}$. Recall the definition $\bar{m}:=\frac{1}{2}\left(m_{1}+m_{2}\right)$. To derive the above, observe that the definition of the scattered field $u_{s}$ implies that

$$
\left\{\begin{array}{l}
\Delta u_{s}(x)-\omega_{2}^{2} u_{s}(x)=0, \quad x \in \Omega_{2},  \tag{5.2.25}\\
\Delta u_{s}(x)-\omega_{1}^{2} u_{s}(x)=0, \quad x \in \Omega_{1} \\
{\left[\left[\hat{n} \cdot \nabla u_{s}\right]\right]+\left(m_{2}+m_{1}\right) u_{s}=-\left[\left[\hat{n} \cdot \nabla u_{i}\right]\right]-\left(m_{2}+m_{1}\right) u_{i}, \quad x \in \Gamma,}
\end{array}\right.
$$

with $u_{s}+u_{i}$ continuous across $\Gamma$ as before. The first two lines of (5.2.25) hold for any choice of $\mu$ and $\rho$. Enforcing continuity of $u$ at the interface, and using the jump relations (5.1.11, 5.1.12) for the layer potentials, we obtain (5.2.23). Since $u$ on $\Gamma$ takes the form

$$
\begin{equation*}
u(\gamma(t))=\frac{1}{2}\left[\mathcal{D}_{\omega_{2}}[\rho](t)+\mathcal{D}_{\omega_{1}}[\rho](t)\right]+\frac{1}{2}\left[\mathcal{S}_{\omega_{2}}[\mu](t)+\mathcal{S}_{\omega_{1}}[\mu](t)\right]+u_{i}(\gamma(t)), \tag{5.2.26}
\end{equation*}
$$

the derivative jump condition in (5.2.25) implies (5.2.24).
Observe that (5.2.23, 5.2.24) reads

$$
\begin{equation*}
\mathcal{L}_{2}\binom{\mu}{\rho}=r \tag{5.2.27}
\end{equation*}
$$

where $\mathcal{L}_{2}$ and $r$ are defined in (5.2.7) and (5.2.10), respectively. As above, $\mathcal{L}_{2}$ may admit a continuous spectrum passing through zero, hence solving for $(\mu, \rho)$ in (5.2.27) is in general an ill-posed problem.

To remedy this issue, we again turn our attention to the special case of a flat boundary, where Fourier representations of the relevant integral operators are given by (5.1.15) and (5.1.16). Taking the Fourier transform of (5.2.23) in the flat case, we find that

$$
\begin{equation*}
\tilde{\rho}(\xi)+\frac{1}{2}\left[\frac{1}{\sqrt{\xi^{2}+\omega_{2}^{2}}}-\frac{1}{\sqrt{\xi^{2}+\omega_{1}^{2}}}\right] \tilde{\mu}(\xi)=-\left[\left[\tilde{u}_{i}\right]\right](\xi) \tag{5.2.28}
\end{equation*}
$$

Similarly, upon taking the Fourier transform of (5.2.24), we obtain

$$
\begin{align*}
-\tilde{\mu}-\frac{1}{2} & {\left[\sqrt{\xi^{2}+\omega_{2}^{2}}-\sqrt{\xi^{2}+\omega_{1}^{2}}\right] \tilde{\rho}+\frac{m_{2}+m_{1}}{4}\left[\frac{1}{\sqrt{\xi^{2}+\omega_{2}^{2}}}+\frac{1}{\sqrt{\xi^{2}+\omega_{1}^{2}}}\right] \tilde{\mu} } \\
& =-\frac{m_{2}+m_{1}}{2}\left(\tilde{u}_{i, 1}+\tilde{u}_{i, 2}\right)-\left[\left[\hat{n} \cdot \widetilde{\nabla u_{i}}\right]\right] \tag{5.2.29}
\end{align*}
$$

After solving (5.2.28) for $\tilde{\rho}$ and substituting it into (5.2.29) we see that

$$
\begin{equation*}
\left[-1+\frac{1}{4}\left(\xi_{2}-\xi_{1}\right)\left(\xi_{2}^{-1}-\xi_{1}^{-1}\right)+\frac{m_{2}+m_{1}}{4}\left(\xi_{2}^{-1}+\xi_{1}^{-1}\right)\right] \tilde{\mu}=\tilde{\psi} \tag{5.2.30}
\end{equation*}
$$

where $\xi_{1,2}=\sqrt{\xi^{2}+\omega_{1,2}^{2}}=\sqrt{\xi^{2}+m_{1,2}^{2}-E^{2}}$, and

$$
\tilde{\psi}(\xi)=-\frac{m_{2}+m_{1}}{2}\left(\tilde{u}_{i, 1}+\tilde{u}_{i, 2}\right)-\left[\left[\hat{n} \cdot \widetilde{\nabla u_{i}}\right]\right]-\frac{1}{2}\left(\xi_{2}-\xi_{1}\right)\left[\left[\tilde{u}_{i}\right]\right](\xi)
$$

Let $R(\xi)$ denote the Fourier multiplier from the left-hand side of (5.2.30) defined by

$$
\begin{equation*}
R(\xi):=-1+\frac{1}{4}\left(\xi_{2}-\xi_{1}\right)\left(\xi_{2}^{-1}-\xi_{1}^{-1}\right)+\frac{m_{2}+m_{1}}{4}\left(\xi_{2}^{-1}+\xi_{1}^{-1}\right) \tag{5.2.31}
\end{equation*}
$$

Differentiating (5.2.31) with respect to $\xi^{2}$ we see that

$$
\frac{\mathrm{d}}{\mathrm{~d} \xi^{2}} R(\xi)=-\frac{m_{2}+m_{1}}{8} \frac{\xi_{2}^{3}+\xi_{1}^{3}}{\xi_{1}^{3} \xi_{2}^{3}}+\frac{1}{8} \frac{\left(\xi_{2}^{2}-\xi_{1}^{2}\right)^{2}}{\xi_{1}^{3} \xi_{2}^{3}}
$$

Now, $\xi_{2}^{2}-\xi_{1}^{2}=\omega_{2}^{2}-\omega_{1}^{2}, \xi_{2} \geq \omega_{2}, \xi_{1} \geq \omega_{1}$, and $m_{1}+m_{2}>\omega_{1}+\omega_{2}$, from which it follows that

$$
\left(m_{2}+m_{1}\right)\left(\xi_{2}^{3}+\xi_{1}^{3}\right)-\left(\xi_{2}^{2}-\xi_{1}^{2}\right)^{2} \geq \omega_{2}^{4}+\omega_{1}^{4}-\omega_{2}^{4}-\omega_{1}^{4}+2 \omega_{2}^{2} \omega_{1}^{2}>0
$$

In particular, $R(\xi)$ is decreasing for $\xi<0$ and increasing for $\xi>0$, from which it follows immediately that $\xi= \pm E$ are the only roots.

Finally, we observe that the nullvectors $(\tilde{\mu}(\xi), \tilde{\rho}(\xi))^{t}$ associated with $\xi= \pm E$ are both

$$
\begin{equation*}
v=\binom{-1}{\frac{1}{2 m_{2}}-\frac{1}{2 m_{1}}} \tag{5.2.32}
\end{equation*}
$$

Motivated by this, we change variables to remove the singularity captured by the nullvectors in (5.2.32). Namely, we introduce the new unknowns, $\sigma_{1}$ and $\sigma_{2}$ defined implicitly by

$$
\begin{equation*}
\binom{\mu}{\rho}=\mathcal{P}_{2}\binom{\sigma_{1}}{\sigma_{2}} \tag{5.2.33}
\end{equation*}
$$

with $\mathcal{P}_{2}$ given by (5.2.8). As with $Q$ for the one-mass case, the Fourier transform of $Q_{2}$ has singularities at $\xi= \pm E$. Thus the singularities of $\mathcal{P}_{2}$ exactly cancel the zeros of $\mathcal{L}_{2}$. Substituting (5.2.33) into the Fourier transformed boundary integral equations (5.2.28) and (5.2.29), we obtain an invertible system for $\sigma_{1}$ and $\sigma_{2}$. In particular, $\mathcal{L}_{2} \mathcal{P}_{2}$ is bounded with bounded inverse.

### 5.3 Numerical apparatus

In this section we describe an algorithm for solving equations (5.1.4, 5.1.6, 5.1.7) via the boundary integral equations (5.2.3) and (5.2.5). For notational convenience we present only the case in which $m_{1}=m_{2}=: m$. The more general case can be solved in a similar way. In subsection 5.3 .1 we briefly discuss the details of the discretization used. Following this, in subsection 5.3.2 we describe several accelerations that were made to improve the computational efficiency.

### 5.3.1 The discretization

In this section, we use the boundary integral equation package chunkie (see [27]) to discretize the interface and construct the entries of the discrete approximation to the operators $\mathcal{L}$ and $\mathcal{P}$ given by (5.2.2) and $\mathcal{L}_{2}$ and $\mathcal{P}_{2}$ defined in (5.2.7) and (5.2.8). To simplify the presentation of our method, we assume that $m_{1}=m_{2}=: m$ throughout this section. This means the relevant integral equation is $\mathcal{L P} \rho=2 m u_{i}$.

The interface is truncated to $[a, b]$ in parameter space, i.e. we restrict $\gamma(t):[a, b]$ to its image. This truncated curve is then adaptively split into "chunks" where each chunk is the image of a subinterval of $[a, b]$, and is discretized using 16 Gauss-Legendre points. Chunks are refined until the tails of the Legendre coefficients of the speed of parameterization, and of the $x, y$ coordinates of the curve are resolved. In particular, suppose that $x_{n}^{(j)}, y_{n}^{(j)}, s_{n}^{(j)}$ are the the Legendre coefficients computed using 32 nodes on a chunk $\left[a_{j}, b_{j}\right]$, i.e.

$$
\left[\begin{array}{l}
x(t)  \tag{5.3.1}\\
y(t) \\
s(t)
\end{array}\right]=\sum_{n=1}^{32}\left[\begin{array}{c}
x_{n}^{(j)} \\
y_{n}^{(j)} \\
s_{n}^{(j)}
\end{array}\right] P_{n}\left(a_{j}+\frac{(t+1)}{2}\left(b_{j}-a_{j}\right)\right)
$$

where $s(t)=\left|\gamma^{\prime}(t)\right|$, and $P_{n}(t)$ is the Legendre polynomial of degree $n$ on $[-1,1]$. Then the
chunk $\left[a_{j}, b_{j}\right]$ is resolved if

$$
\begin{equation*}
\max \left(\sqrt{\frac{\sum_{n=17}^{32}\left|x_{n}\right|^{2}}{16}}, \sqrt{\frac{\sum_{n=17}^{32}\left|y_{n}\right|^{2}}{16}}, \sqrt{\frac{\sum_{n=17}^{32}\left|s_{n}\right|^{2}}{16}}\right) \leq \varepsilon \tag{5.3.2}
\end{equation*}
$$

for a specified tolerance $\varepsilon$. If the chunk is not resolved, then it is split into two chunks of equal length in parameter space $\left[a_{j},\left(a_{j}+b_{j}\right) / 2\right]$, and $\left[\left(a_{j}+b_{j}\right) / 2, b_{j}\right]$. Once all of the chunks are resolved, they are subsequently balanced so that adjacent chunks satisfy a $2: 1$ length restriction: if $\gamma_{j}$ and $\gamma_{\ell}$ are adjacent to each other then they satisfy $\left|\gamma_{j}\right| /\left|\gamma_{\ell}\right| \in[0.5,2]$. At the end of the adaptive procedure, the restriction of $\gamma$ to $[a, b]$ is represented via a collection of $n_{c}$ chunks, $[a, b]=\cup_{j=1}^{n_{c}}\left[a_{j}, a_{j+1}\right]$ with the understanding that $a_{1}=a$, and $a_{n_{c}+1}=b$.

Remark 5.3.1. In the proofs, we have assumed $\gamma^{\prime} \equiv 1$ for convenience. The proofs can be suitably modified as long as $\left|\left|\gamma^{\prime}\right|-1\right| \leq C e^{-\alpha|t|}$ for some positive constants $C$ and $\alpha$. Relaxing this restriction provides greater flexibility for parametrizing complicated curves.

Recall that the solution $\rho \in L_{\alpha}^{2}(\mathbb{R})$, and hence there exists an $M$ such that

$$
\begin{equation*}
\sqrt{\frac{\int_{[-\infty,-M] \cup[M, \infty]}|\rho|^{2} d s}{\int_{-\infty}^{\infty}|\rho|^{2} d s}} \leq \varepsilon \tag{5.3.3}
\end{equation*}
$$

This estimate justifies the existence of a truncation $[a, b]$ such that the solution can be accurately represented via its restriction to some bounded interval of $\mathbb{R}$. In all the examples, the interval $[a, b]$ is taken to be the smallest interval satisfying the following two criteria: a) that the boundary is nearly flat outside of $[a, b]$, i.e. there exists a constant vector $c \in \mathbb{R}^{2}$, such that $\left|\gamma^{\prime}(t)-c\right| \leq \varepsilon$ for all $c \in \mathbb{R}^{2} \backslash[a, b]$, and b) the boundary data $u_{i}$ is numerically supported on $[a, b]$ to precision $\varepsilon$, i.e. $\left|u_{i}\right|_{L^{2}(\mathbb{R} \backslash[a, b])} /\left|u_{i}\right|_{L^{2}(\mathbb{R})} \leq \varepsilon$.

Given these restrictions, the composition $\mathcal{L P}$ is discretized as an intergral operator on $L^{2}$ functions defined on the interval $[a, b]$. Even though $\rho$ will be numerically supported on $[a, b]$, the operator $\mathcal{P}$ maps compactly supported functions to an oscillatory function that is
$O(1)$ on the whole real line. On the other hand, the kernel in the integral operator $\mathcal{L}$ decays exponentially as $\exp (-\omega d)$, where $d$ denotes the Euclidean distance between points on the interface. Thus, in order to compute the solution $\rho$ accurately, one needs to discretize the integral operator $\mathcal{P}$ from $L^{2}$ functions on $[a, b]$ to $L^{2}$ functions on $\left[a^{\prime}, b^{\prime}\right]$, and the integral operator $\mathcal{L}$ from $L^{2}$ functions on $\left[a^{\prime}, b^{\prime}\right]$ to $L^{2}$ functions $[a, b]$, where $a^{\prime}=a-\log (1 / \varepsilon) / \omega$, and $b^{\prime}=b+\log (1 / \varepsilon) / \omega$. The necessity and sufficiency of this choice of the buffer region is illustrated through the results in Section 5.4.1.

We now turn our attention to the discretization of the integral operators $\mathcal{L}$ and $\mathcal{P}$. Suppose that there are an additional $n_{\text {buffer }}$ points introduced in each of the buffer regions $\left[a^{\prime}, a\right]$ and $\left[b, b^{\prime}\right]$. Suppose that $t_{j} \in\left[a^{\prime}, b^{\prime}\right], j=1,2 \ldots n_{\text {over }}=16 n_{c}+2 n_{\text {buffer }}$ are the discretized values of $t$ in parameter space. Let $n_{0}=n_{\text {buffer }}+16 n_{c}$. Suppose that the points are ordered in increasing values of $t$. Then the points $j=1,2 \ldots n_{\text {buffer }}$ correspond to the left buffer region $\left[a^{\prime}, a\right]$, the points $j=n_{\text {buffer }}+1, \ldots n_{0}$ correspond to the interval $[a, b]$, and the points $j=n_{0}+1, \ldots n_{\text {over }}$ correspond to the right buffer region $\left[b, b^{\prime}\right]$. Note that the density $\rho$ is discretized through its values at $\rho\left(t_{j}\right)$, for $j=n_{\text {buffer }}+1, \ldots n_{0}$, and hence the discretized linear system corresponding to $\mathcal{L P}$ will be of size $n_{0}-n_{\text {buffer }}=16 n_{c}$.

Both the kernels $\mathcal{L}$ and $\mathcal{P}$ have non-smooth and at most weakly singular kernels for small distances and require specialized quadrature rules for integrating them. We use the generalized Gaussian quadrature rules of $[22,21]$ for the accurate computation of these integrals. In particular, the quadrature rule is a target dependent locally-corrected quadrature rule which accurately integrates the specific non-smooth behavior of the kernel in the vicinity of the origin. For every point $t_{j} \in\left[a_{i}, a_{i+1}\right]$, there exist weights $w_{j, \ell}$ for all $t_{\ell} \in\left[a_{i-1}, a_{i+1}\right]$ such that the discretized versions of $\mathcal{P}$ and $\mathcal{L}$, denoted by $\mathbf{P}$ and $\mathbf{L}$ respectively, are given by

$$
\begin{align*}
& \mathbf{P}[\rho]\left(t_{j}\right)=\frac{m^{2}}{E} \sum_{\substack{\ell=n_{\text {buffer }}+1 \\
n_{0}}}^{n_{0}}\left(\delta_{j, \ell}+e^{i E\left|t_{j}-t_{\ell}\right|} w_{\ell}\right) \rho\left(t_{\ell}\right) \\
& \\
& +\sum_{\substack{\ell=n_{\text {buffer }}+1 \\
t_{\ell} \in\left[a_{i-1}, a_{i+1}\right] \\
n_{\text {over }}}} \rho\left(t_{\ell}\right) w_{j, \ell}, \quad j=1,2, \ldots n_{\text {over }}  \tag{5.3.4}\\
& \mathbf{L}[\mu]\left(t_{j}\right)=\mu\left(t_{j}\right)-2 m \sum_{\substack{\ell=1 \\
\ell \neq j}} G_{\omega}\left(\gamma\left(t_{j}\right), \gamma\left(t_{\ell}\right)\right) \mu\left(t_{\ell}\right) w_{\ell} \\
& \quad+\sum_{\substack{\ell=1 \\
t_{\ell} \in\left[a_{i-1}, a_{i+1}\right]}}^{n_{\text {over }}} \mu\left(t_{\ell}\right) w_{j, \ell}, \quad j=n_{\text {buffer }}+1, n_{\text {buffer }}+2, \ldots n_{0} .
\end{align*}
$$

Here $\delta_{j, \ell}$ is the Kronecker delta, i.e. $\delta_{j, j}=1$ and $\delta_{j, \ell}=0$ otherwise, and $w_{j}, j=$ $1,2 \ldots n_{\text {over }}$, denote the quadrature weights for integrating smooth functions on the interface. The factorized linear systems $\mathbf{L}$ and $\mathbf{P}$ are illustrated in Figure 5.3.

Remark 5.3.2. Recall that $\mathcal{L}$ may have a continuous spectrum passing through zero while $\mathcal{P}$ is not bounded; see (5.2.19). Thus it is expected (and observed numerically) that in general, $\mathbf{L}$ and $\mathbf{P}$ both have very large condition numbers. This could in principle lead to catastrophic cancellation arising from the numerical implementation of the operator product LP. We postpone a thorough treatment of this potential issue to future study. Practically speaking, the accuracy of our numerical method does not seem to suffer from the poor condition numbers of $\mathbf{L}$ and $\mathbf{P}$. Indeed, as illustrated by Section 5.4.1, we observe low errors for a variety of interfaces and wide range of parameters.

### 5.3.2 Accelerations of the numerical method

In this section, we discuss a fast algorithm for the evaluation of the matrix vector product $\mathbf{L P}[\rho]$. The kernel of $\mathbf{P}$ is the Green's function of a one-dimensional translation-invariant el-


Figure 5.2: Schematic of the discretization approach used by chunkie. In the inlay, bounds between 'chunks' are shown with vertical lines, and discretization nodes are denoted by red triangles. For clarity, the 'panel' shown is $8^{\text {th }}$ rather than $16^{\text {th }}$.


Figure 5.3: The factorized linear system, after discretization.
liptic ordinary differential equation and hence can be accelerated using a sweeping algorithm. In order to apply $\mathbf{P}$ rapidly, we just need a fast algorithm for the evaluation of

$$
\begin{equation*}
\mathbf{P}[\rho]\left(t_{j}\right)=\frac{m^{2}}{E} \sum_{\ell=n_{\mathrm{buffer}}+1}^{n_{0}} e^{i E\left|t_{j}-t_{\ell}\right|^{\prime}} \rho\left(t_{\ell}\right) w_{\ell}, \quad j=1,2, \ldots n_{\mathrm{over}} \tag{5.3.5}
\end{equation*}
$$

since the rest of the interaction is sparse and can be computed in $O\left(n_{\text {over }}\right)$ CPU time. The main idea of the sweeping algorithm is to split the solution into two parts for any point $t_{j}$, $t \leq t_{j}$, and $t>t_{j}$, where both pieces can be updated in $O(1)$ operations as we move from $t_{j} \rightarrow t_{j+1}$ or $t_{j} \rightarrow t_{j-1}$. Let $v^{\uparrow}$ and $v^{\downarrow}$ denote the accumulation of the rightward moving
solution (corresponding to $t \leq t_{j}$ ) and the leftward moving solution(corresponding to $t>t_{j}$ ) respectively.

In particular, we split the solution as follows,

$$
\begin{align*}
P[\rho]\left(t_{j}\right) & =\frac{m^{2}}{E} \sum_{\substack{\ell=n_{\text {buffer }}+1 \\
\ell \leq j}}^{n_{0}} e^{i E\left(t_{j}-t_{\ell}\right)} \rho\left(t_{\ell}\right) w_{\ell}+\frac{m^{2}}{E} \sum_{\substack{\ell=n_{\text {buffer }}+1 \\
\ell>j}}^{n_{0}} e^{i E\left(t_{\ell}-t_{j}\right)} \rho\left(t_{\ell}\right) w_{\ell}  \tag{5.3.6}\\
& =v_{j}^{\uparrow}+v_{j}^{\downarrow}
\end{align*}
$$

A simple calculation shows that $v^{\uparrow}$ and $v^{\downarrow}$ satisfy the following recurrence relations

$$
\begin{align*}
v_{j}^{\uparrow} & =v_{j-1}^{\uparrow} e^{i E\left(t_{j}-t_{j-1}\right)}+\frac{m^{2}}{E} \rho\left(t_{j}\right) w_{j} I_{j \in\left[n_{\mathrm{buffer}}+1, n_{0}\right]}  \tag{5.3.7}\\
v_{j}^{\downarrow} & =e^{i E\left(t_{j+1}-t_{j}\right)}\left(v_{j+1}^{\downarrow}+\frac{m^{2}}{E} \rho\left(t_{j+1}\right) w_{j+1} I_{j+1 \in\left[n_{\mathrm{buffer}}+1, n_{0}\right]}\right),
\end{align*}
$$

where $I_{j \in A}$ is the indicator function of the set $A$, which is equal to 1 if $j \in A$, and 0 otherwise. Thus, $v^{\uparrow}$ satisfies an upward recurrence in $j$, while $v^{\downarrow}$ satisfies a downward recurrence, and both $v^{\uparrow}$ and $v^{\downarrow}$ can be computed for all $j$ in $O\left(n_{\text {over }}\right)$ work. The recurrences are initialized with $v_{1}^{\uparrow}=0$, and $v_{n_{\text {over }}}^{\downarrow}=0$.

On the other hand, the kernel of $\mathbf{L}$ is the Green's function of the two dimensional Helmholtz equation with imaginary wave number $\omega$ and the bulk of the computation is given by

$$
\begin{equation*}
-2 m \sum_{\substack{\ell=1 \\ \ell \neq j}}^{n_{\mathrm{over}}} G_{\omega}\left(\gamma\left(t_{j}\right), \gamma\left(t_{\ell}\right)\right) \mu\left(t_{\ell}\right) w_{\ell}, \quad j=n_{\text {buffer }}+1, \ldots n_{0} \tag{5.3.8}
\end{equation*}
$$

The above sum can be computed at all $t_{j}, j=n_{\text {buffer }}+1, \ldots n_{0}$ in $O\left(n_{\text {over }}\right)$ CPU time using the standard fast multipole method; see [73, 50]. We use the fast multipole implementation in $f m m 2 d$ for evaluating the sum in [45]. The rest of the computation in $\mathbf{L}$ is sparse whose number of nonzero elements is also $O$ ( $\left.n_{\text {over }}\right)$.

Combining both of these fast algorithms, the matrix vector product $\mathbf{L P}[\rho]$ can be applied
in $O\left(n_{\text {over }}\right)$ CPU time. Thus, the solution $\rho$ can be obtained in $O\left(n_{\text {over }} n_{\text {iter }}\right) \mathrm{CPU}$ time using iterative methods like the generalized minimum residual (GMRES) method, where $n_{\text {iter }}$ is the number of GMRES iterations required for the relative residual to drop below a prescribed tolerance. In practice, the integral equation tends to be well-conditioned which results in $n_{\text {iter }}=O(1)$, and thus the computational complexity of obtaining the solution $\rho$ is $O\left(n_{\text {over }}\right)$.

### 5.4 Numerical illustrations and examples

In this section, we provide several examples of the numerical method described in Section 5.3. We demonstrate accuracy or self-convergence of the algorithm and test its speed for a variety of interfaces; see subsection 5.4.1. In subsection 5.4.2, we plot corresponding solutions and compute reflection coefficients for a scattering theory. We refer to Figure 5.4 for an illustration of the interfaces used in our examples.

### 5.4.1 Illustration of the numerical method

This section presents the accuracy and speed of our numerical method for various interfaces. We begin with the flat-interface $\left(\Gamma_{0}\right)$ case, where there is an analytic expression for the Green's function. For simplicity, we assume that $m_{1}=m_{2}=: m$. If $u$ denotes our computed solution, we define the relative error of $u$ at the point $x_{T}$ by $\left|u\left(x_{T}\right)-u_{*}\left(x_{T}\right)\right| /\left|u_{*}\left(x_{T}\right)\right|$, where $u_{*}$ is the true solution. Figure 5.5 contains a plot of this relative error (computed at four arbitrary points) as a function of $n_{c}$, as well as illustrations of the computed Green's function and densities. Observe that our solution is highly accurate even for small values of $n_{c}$.

For arbitrary non-flat interfaces (such as $\Gamma_{1}, \Gamma_{2}, \Gamma_{3}$ ), analytic solutions are not known and hence we cannot compute the exact relative error. Instead, we perform a self-convergence test, which involves approximating the true solution by the numerical solution at some


Figure 5.4: The interfaces $\Gamma_{0}$ (top left), $\Gamma_{1}$ (top right), $\Gamma_{2}$ (bottom left) and $\Gamma_{3}$ (bottom right), with respective sources at $(0,2.5),(0,1),(0,-7)$ and $(0,3)$ as indicated by the red dot. Outside the plotted region, the interfaces extend linearly to infinity.


Figure 5.5: Densities $\mu$ and $\rho$ (top left panel), relative error of the computed solution at the indicated points as a function of $n_{c}$ (bottom left panel), and Green's function $u$ (center and right panels) corresponding to the flat interface $\Gamma_{0}$ with $m=2$ and $E=1$. The top left panel zooms in on the region $[10,10] \times\{0\} \subset \Gamma_{0}$, with $t=60$ corresponding to the point $(0,0) \in \Gamma_{0}$.
large value $N^{*}$ of $n_{c}$. If we now let $u_{N}$ denote the computed solution with $n_{c}=N$, our approximate relative error at the target $x_{T}$ is given by $\left|u_{N}\left(x_{T}\right)-u_{N^{*}}\left(x_{T}\right)\right| /\left|u_{N^{*}}\left(x_{T}\right)\right|$. For a corresponding plot with $N^{*}=512$, see Figure 5.6 (left panel). We again observe fast convergence in $n_{c}$, though understandably not as fast as the flat-interface case.

Another parameter of interest is the truncation length $n_{\text {buffer }}$ (introduced in Section 5.3.1). Its default value (unless otherwise specified) is

$$
n_{\text {buffer }}=2\left\lceil\frac{n_{c} \log \left(10^{16}\right)}{m_{0} \Delta_{t}}\right\rceil=: 2\left\lceil M_{b}\right\rceil,
$$

where $m_{0}:=\min \left(m_{1}, m_{2}\right)$ and $\Delta_{t}$ is the arclength of $\Gamma$ over the entire discretized region (that is, the region discretized by all $n_{\text {over }}=n+2 n_{\text {buffer }}$ grid points; see Section 5.3.1). Here, $M_{b} \in \mathbb{N}$ is sufficiently large so that interactions between points separated by at least $M_{b}$ grid points are bounded by $10^{-16}$ in absolute value. Some routines in the numerical experiments presented here were set to a tolerance of $10^{-12}$, so the error of $10^{-16}$ more than ensures that any error we observe would not be due to truncation.

To test the effect of $n_{\text {buffer }}$ on the convergence of our method, we introduce $\tau>0$ and set $n_{\text {buffer }}=2\left\lceil\tau M_{b}\right\rceil$. In the center panel of Figure 5.6, we plot

$$
\left|u_{n_{c}, \tau}\left(x_{T}\right)-u_{512, \tau}\left(x_{T}\right)\right| /\left|u_{512, \tau}\left(x_{T}\right)\right|
$$

as a function of $\tau$, where $u_{N, \tau}$ is our computed solution with $n_{c}=N$ and $n_{\text {buffer }}=2\left\lceil\tau M_{b}\right\rceil$. We set $n_{c}=128,64,256$ for $\Gamma_{1}, \Gamma_{2}, \Gamma_{3}$, respectively. Given the small truncation error tolerance at $\tau=1$, it makes sense that decreasing the value of $\tau$ from 1 would not immediately increase the relative error. Still, once $\tau$ gets small enough (say, less than 0.5 ), the truncation length is too small and the convergence of our method suffers. The relative error increases when $\tau$ increases from 1 , as we do not keep enough grid points in this case.

In summary, the left panel illustrates relative error at fixed $\tau=1$ and varying $n_{c}$, while


Figure 5.6: Self-convergence tests for varying $n_{c}$ (left panel) and truncation length (center panel). The relative error is computed at $x_{T}=(-1,1),(7,-7),(1,-3)$ for $\Gamma_{1}, \Gamma_{2}, \Gamma_{3}$ respectively. For each interface, the source location is given by Figure 5.4. The respective values of $m$ and $E$ are those from Figures 5.7 (top right panel), 5.8 and 5.9 below. The computational cost of obtaining $u\left(x_{T}\right)$ is illustrated by the right panel.
the center panel illustrates relative error at fixed $n_{c}$ (depending on the interface) and varying $\tau$. We refer to the right panel of Figure 5.6 for a plot of the speed of our method as a function of $n_{c}$. As predicted, the computation time grows only linearly in $n_{c}$. The respective slopes of the line of best fit for $\Gamma_{1}, \Gamma_{2}, \Gamma_{3}$ are $1.06 \times 10^{-2}, 1.03 \times 10^{-2}, 8.43 \times 10^{-3}$, with an average value of $9.78 \times 10^{-3}$. These slopes were computed using only the data for $n_{c}>50$ to eliminate the effect of the nonlinear behavior of the curves for small $n_{c}$. The lines of best fit are used to extrapolate the data to $n_{c}>256$ in the plot.

### 5.4.2 Examples of applications

We now present various examples corresponding to the non-flat interfaces from Figure 5.4. The Green's function, $u$, and densities, $\mu$ and $\rho$, for a source (whose location is given by Figure 5.4) near a non-flat section of the interface are plotted in Figures 5.7-5.10. As expected, $\omega=\sqrt{m^{2}-E^{2}}$ dictates the rate at which solutions decay away from the interface. In particular, larger values of $\omega$ result in faster decaying solutions. In the case of different masses, we observe that $u$ decays more rapidly in the domain with larger $\omega$; see Figures 5.7 and 5.10.

Our numerical experiments (Figures 5.5 and 5.9 ) verify that the density $\rho$ from (5.2.5) is


Figure 5.7: Green's function, $u$, for interface $\Gamma_{1}$ with $m_{1}=2, E=0.8$ and $m_{2}=3$ (top left) $m_{2}=2$ (top right), $m_{2}=1.5$ (bottom left) and $m_{2}=1$ (bottom right).


Figure 5.8: Densities, $\mu$ and $\rho$, and Green's function, $u$, for the interface $\Gamma_{2}$ with $m_{1}=m_{2}=$ $2 / 3$ and $E=1 / 3$. The left plot zooms in on the part of $\Gamma_{2}$ connecting the points $(-30.0,0.0)$ and $(14.8,-8.5)$, with $t=185$ corresponding to the point $(-13.5,-10.6)$.


Figure 5.9: Density, $\mu$, and Green's function, $u$, for the interface $\Gamma_{3}$ with $m_{1}=m_{2}=3$ and $E=2$. The left plot zooms in on the part of $\Gamma_{3}$ connecting the points $(-10.0,0.0)$ and (18.0, 0.0), with $t=70$ corresponding to the point (1.4, -0.2 ).


Figure 5.10: Green's function, $u$, for the interface $\Gamma_{3}$, where $\left(m_{1}, m_{2}, E\right)=(1.20,4.00,1.00)$ for the left panel, $\left(m_{1}, m_{2}, E\right)=(2.00,3.00,1.50)$ for the center panel, and $\left(m_{1}, m_{2}, E\right)=$ $(2.25,6.00,2.00)$ for the right panel.


Figure 5.11: Green's function, $u$, for the interface $\Gamma_{2}$ with source located at $(-40,1)$. Here, $(m, E)$ is $(0.75,0.25),(0.75,0.5),(1.5,1)$ and $(4,1)$ for the top-left, top-right, bottom-left and bottom-right panels, respectively.
rapidly decaying, while $\mu:=\mathcal{P} \rho$ obeys an outgoing radiation condition, oscillating without decay. For an illustration of $\rho$ and $\mu$ in the vicinity of a source (zooming in on the region where $\rho$ is not small), see Figure 5.8.

In Figure 5.11 we move the source to $(-40,1)$ and observe that the propagation of the resulting wave along the interface depends on the choice of $(m, E)$. As stated in Remark 5.2.6, increasing the value of $\omega$ is equivalent to smoothing out the interface, thus it makes sense that the solutions on the top row get reflected while those on the bottom get transmitted. Similarly, the small values of $m$ and $E$ in Figure 5.8 (combined with the corresponding source location; see Figure 5.4, bottom left panel) result in a solution that is concentrated near the oscillatory part of $\Gamma_{2}$.

We conclude this section with a scattering experiment. Let us consider the interface
parametrized by $\gamma(t)=2 e^{-0.05 t^{2}} \sin (b t+0.4)$, for some $b \geq 0$. Note that the value $b=2$ gives $\Gamma_{3}$. We place a source near the interface, to the left of and far away from its oscillations.

The density $\mu$ is proportional to the solution $u$ along the interface, as (5.2.16) implies that $\mu=2 m\left(u_{i}+u_{s}\right)=2 m u$ on $\Gamma$. The outgoing condition (5.1.6) implies that $\mu(t) \approx C e^{i E t}$ for $t>0$ sufficiently large. Between the source and oscillations, we have that $\mu(t) \approx$ $A e^{i E t}+B e^{-i E t}$, where $|A|$ and $|B|$ are the respective amplitudes of the incoming and reflected waves. The transmission and reflection coefficients are then defined by $T_{L}:=|C|^{2} /|A|^{2}$ and $R_{L}:=|B|^{2} /|A|^{2}$.

Since $\mu=(1+Q) \rho$ with $\rho \in L^{1}$, it follows that $\mu(t) \approx \frac{m^{2}}{E} e^{ \pm i E t} \hat{\rho}( \pm E)$ as $t \rightarrow \pm \infty$. It follows that $C=\frac{m^{2}}{2 E} \hat{\rho}(E)$. As $t \rightarrow-\infty$, we get contributions from both the reflection and the source. Hence $B=L-B_{0}$, where $L=\frac{m^{2}}{2 E} \hat{\rho}(-E)$ and $B_{0}=\frac{m^{2}}{2 E} \hat{\rho}_{0}(-E)$ with $\rho_{0}$ the solution corresponding to $k=0$. Thus, using the identity $T_{L}+R_{L}=1$, we compute the transmission and reflection coefficients using only $\hat{\rho}_{0}(-E)$ and $\hat{\rho}( \pm E)$.

For a plot of $R_{L}$ as a function of $b$, see Figure 5.12 (left panel). When $b$ is small, $\Gamma$ resembles a flat interface and thus $R_{L}$ is close to 0 . For larger values of $b$ (say $b>2$ ), the oscillations of the interface cause solutions to back-scatter, as $R_{L}$ is approximately 1 . The transition of $R_{L}$ from 0 to 1 contains a small interval in $b$ at which there is a sudden dip in $R_{L}$. See Figure 5.12 (center and right panels) for an illustration of the qualitatively different behavior of solutions corresponding to nearly identical interfaces. We suspect there may be other values of $b$ (out of the range of values in Figure 5.12) for which sharp transitions in $R_{L}$ occur, but postpone a thorough investigation of these critical values to future analyses.


Figure 5.12: Scattering experiment for an interface parametrized by $\gamma(t)=2 e^{-0.05 t^{2}} \sin (b t+$ $0.4)$, for $b \geq 0$. The source is located at $(-40,1)$, and $(m, E)=(4,1)$. The center and right panels illustrate the Green's function, $u$, corresponding to $b=1.87$ and $b=2.00$, respectively.

### 5.5 Proofs of main analytical results

This section is devoted to proving the statements from Section 5.2.1. First, we prove Theorems 5.2.1 and 5.2.4, verifying that $u$ obtained by our boundary integral formulation solves the PDE (5.1.4, 5.1.5, 5.1.6, 5.1.7). We then prove Theorem 5.2.5 using Lemma 5.5.3 and Proposition 5.5.6, which in turn require Lemma 5.5.2 below. For completeness, we also state and prove Proposition 5.5.5-an extension of Lemma 5.5.3 which could be used interchangeably with Proposition 5.5.6 in our proof of Theorem 5.2.5. These propositions and lemmas provide bounds on $\|\mathcal{M}\|$ that guarantee bounded invertibility of $1+\mathcal{M}$ in the $\lambda \rightarrow \infty$ limit. This section concludes with a proof of Theorem 5.2.7. The latter uses Lemma 5.5.7 below, which asserts that any solution of the homogeneous Klein-Gordon PDE (5.5.10) with complexified $E$ must be trivial.

Let us begin with the proofs of Theorems 5.2.1 and 5.2.4.

Proof of Theorem 5.2.1. That (5.1.7) holds is a consequence of the exponential decay of $K_{0}(\omega|x-\gamma(t)|)$ in $|x-\gamma(t)|$ and the Lebesgue dominated convergence theorem. The latter also implies that (5.1.4) holds, while the jump conditions (5.1.5) follow immediately from the derivation of (5.2.16).

Let us now prove the outgoing condition (5.1.6). Since $f \in \mathcal{C}_{c}^{\infty}\left(\mathbb{R}^{2}\right)$, we have that
$u_{i} \in \mathcal{C}^{\infty}$ with $u_{i}$ and all its derivatives decaying rapidly at infinity. Hence $u_{i}$ satisfies (5.1.6), so it remains to consider $u_{s}$. We will treat each term on the right-hand side of $\mu=\rho+Q \rho$ separately. Set $r \neq 0$ and take $t_{0}>0$ sufficiently large so that $\left\{\gamma(t)+r \hat{n}(t): t \geq t_{0}\right\} \cap \Gamma=\emptyset$. Note that our assumptions (5.1.1) and (5.1.2) on $\gamma$ ensure that $t_{0}$ is well defined. Let $t \geq t_{0}$. Then $g(s):=\gamma(t)+r \hat{n}(t)-\gamma(s)$ is a smooth non-vanishing function satisfying $|g(s)| \gtrsim|t-s|$ for all $|t-s|$ sufficiently large. Thus for $j \in\{0,1\}, \partial_{t}^{j} K_{0}(\omega|\gamma(t)+r \hat{n}(t)-\gamma(s)|)$ is bounded and exponentially decaying in $|t-s|$. Since $\rho \in L^{1}$, this means

$$
\begin{equation*}
\left(\partial_{t}-i E\right) S_{\omega}[\rho](\gamma(t)+r \hat{n}(t))=\frac{1}{2 \pi} \int_{\mathbb{R}}\left(\partial_{t}-i E\right) K_{0}(\omega|\gamma(t)+r \hat{n}(t)-\gamma(s)|) \rho(s) \mathrm{d} s \tag{5.5.1}
\end{equation*}
$$

goes to 0 as $t \rightarrow \infty$. We now consider $Q \rho$. Again using that $\rho \in L^{1}$, it follows that

$$
\begin{align*}
\left(\partial_{t}-i E\right) Q \rho(t) & =i m^{2} \int_{-\infty}^{\infty}\left(\operatorname{sgn}\left(t-t^{\prime}\right)-1\right) e^{i E\left|t-t^{\prime}\right|} \rho\left(t^{\prime}\right) \mathrm{d} t^{\prime} \\
& =-2 i m^{2} \int_{t}^{\infty} e^{-i E\left(t-t^{\prime}\right)} \rho\left(t^{\prime}\right) \mathrm{d} t^{\prime} \tag{5.5.2}
\end{align*}
$$

goes to 0 as $t \rightarrow \infty$. Writing

$$
\partial_{t} S_{\omega}[Q \rho](\gamma(t)+r \hat{n}(t))=\frac{1}{2 \pi} \int_{\mathbb{R}} \partial_{t} K_{0}(\omega|\gamma(t)+r \hat{n}(t)-\gamma(s)|) Q \rho(s) \mathrm{d} s
$$

the strategy now is to show that $\partial_{t} K_{0}$ above is well approximated by $-\partial_{s} K_{0}$, so that we can then integrate by parts in $s$ and apply (5.5.2). By the assumed exponential decay of $\gamma^{\prime \prime}$, it follows that

$$
\begin{equation*}
\left|\partial_{t}\right| \gamma(t)+r \hat{n}(t)-\gamma(s)\left|+\partial_{s}\right| \gamma(t)+r \hat{n}(t)-\gamma(s)| | \leq C_{1} e^{-\tilde{\beta} \min \{t, s\} 1_{\{s>0\}}} \tag{5.5.3}
\end{equation*}
$$

for some positive constants $C_{1}$ and $\tilde{\beta}$. Defining $R(t, s):=C_{1} \omega K_{1}(\omega|\gamma(t)+r \hat{n}(t)-\gamma(s)|)$,
this means

$$
\left|\left(\partial_{t}+\partial_{s}\right) K_{0}(\omega|\gamma(t)+r \hat{n}(t)-\gamma(s)|)\right| \leq R(t, s) e^{-\tilde{\beta} \min \{t, s\} 1_{\{s>0\}}}
$$

and thus

$$
\begin{aligned}
& \int_{\mathbb{R}}\left|\left(\partial_{t}+\partial_{s}\right) K_{0}(\omega|\gamma(t)+r \hat{n}(t)-\gamma(s)|) Q \rho(s)\right| \mathrm{d} s \\
& \leq \int_{-\infty}^{t / 2} R(t, s)|Q \rho(s)| \mathrm{d} s+\int_{t / 2}^{\infty} R(t, s) e^{-\tilde{\beta} \min \{t, s\}}|Q \rho(s)| \mathrm{d} s
\end{aligned}
$$

Note that $\rho \in L^{1}$ implies $Q \rho \in L^{\infty}$. Since there exist positive constants $C_{2}$ and $\eta$ such that $R(t, s) \leq C_{2} e^{-\eta|t-s|}$ for all $t \geq t_{0}$ and $s \in \mathbb{R}$, it follows that

$$
\begin{aligned}
\int_{-\infty}^{t / 2} R(t, s)|Q \rho(s)| \mathrm{d} s+\int_{t / 2}^{\infty} R(t, s) e^{-\tilde{\beta} \min \{t, s\}} \mid & Q \rho(s) \mid \mathrm{d} s \\
& \leq\left(\frac{C_{2}}{\eta} e^{-\eta t / 2}+\frac{2 C_{2}}{\eta} e^{-\tilde{\beta} t / 2}\right)\|Q \rho\|_{\infty}
\end{aligned}
$$

hence

$$
\lim _{t \rightarrow \infty} \int_{\mathbb{R}}\left|\left(\partial_{t}+\partial_{s}\right) K_{0}(\omega|\gamma(t)+r \hat{n}(t)-\gamma(s)|) Q \rho(s)\right| \mathrm{d} s=0
$$

Integrating by parts, we obtain

$$
\begin{aligned}
& \left|\int_{\mathbb{R}}\left(\partial_{s}+i E\right) K_{0}(\omega|\gamma(t)+r \hat{n}(t)-\gamma(s)|) Q \rho(s) \mathrm{d} s\right| \\
& =\left|\int_{\mathbb{R}} K_{0}(\omega|\gamma(t)+r \hat{n}(t)-\gamma(s)|)\left((Q \rho)^{\prime}(s)-i E Q \rho(s)\right) \mathrm{d} s\right| \\
& \leq \\
& \quad \int_{-\infty}^{t / 2} K_{0}(\omega|\gamma(t)+r \hat{n}(t)-\gamma(s)|)\left|(Q \rho)^{\prime}(s)-i E Q \rho(s)\right| \mathrm{d} s \\
& \quad \quad+\left\|(Q \rho)^{\prime}-i E Q \rho\right\|_{L^{\infty}[t / 2, \infty)} \int_{t / 2}^{\infty} K_{0}(\omega|\gamma(t)+r \hat{n}(t)-\gamma(s)|) \mathrm{d} s .
\end{aligned}
$$

The first term on the above right-hand side goes to 0 as $t \rightarrow \infty$ by the exponential decay of $K_{0}$. The second term goes to 0 by (5.5.2) and the fact that $\left\|K_{0}(\omega|\gamma(t)+r \hat{n}(t)-\gamma(\cdot)|)\right\|_{L^{1}} \leq$ $C$ uniformly in $t \geq t_{0}$. We conclude that

$$
\begin{aligned}
\left|\left(\partial_{t}-i E\right) S_{\omega}[Q \rho](\gamma(t)+r \hat{n}(t))\right| & \leq \frac{1}{2 \pi}\left(\int_{\mathbb{R}}\left|\left(\partial_{t}+\partial_{s}\right) K_{0}(\omega|\gamma(t)+r \hat{n}(t)-\gamma(s)|) Q \rho(s)\right| \mathrm{d} s\right. \\
& \left.+\left|\int_{\mathbb{R}}\left(\partial_{s}+i E\right) K_{0}(\omega|\gamma(t)+r \hat{n}(t)-\gamma(s)|) Q \rho(s) \mathrm{d} s\right|\right)
\end{aligned}
$$

with both terms on the right-hand side going to 0 as $t \rightarrow \infty$. We have thus shown that $\lim _{t \rightarrow \infty}\left(\partial_{t}-i E\right) u(\gamma(t)+s \hat{n}(t))=0$ The same argument applies when $t<0$ (with $\partial_{t}$ replaced by $-\partial_{t}$ ) and thus we have verified (5.1.6).

Proof of Theorem 5.2.4. As before, (5.1.7) and (5.1.4) follow immediately from dominated Lebesgue, while the derivation of $(5.2 .23,5.2 .24)$ implies (5.1.5). We now prove (5.1.6). Recalling the definition of $(\mu, \rho)$ in (5.2.11), the same arguments from Theorem 5.2.1 imply that $\mu=\mu_{0}+\mu_{1}$ and $\rho=\rho_{0}+\rho_{1}$, where $\mu_{0}, \rho_{0} \in L^{1}, \lim _{t \pm \infty}\left(\partial_{t} \mp i E\right) \mu_{1}(t)=0$ and $\lim _{t \pm \infty}\left(\partial_{t} \mp i E\right) \rho_{1}(t)=0$. Hence (following the proof of Theorem 5.2.1) $u_{i}+S_{\omega_{j}}[\mu]$ satisfies (5.1.6) for $j \in\{1,2\}$. It remains to analyze $D_{\omega_{j}}[\rho]$. Fix $r>0$, take $t_{0}>0$ sufficiently large so that $\left\{\gamma(t)+r \hat{n}(t): t \geq t_{0}\right\} \cap \Gamma=\emptyset$, and let $t \geq t_{0}$. Defining $g(t, s):=\gamma(t)+r \hat{n}(t)-\gamma(s)$, we have

$$
D_{\omega_{2}}[\rho](\gamma(t)+r \hat{n}(t))=\frac{1}{2 \pi} \int_{\mathbb{R}} \hat{n}(s) \cdot \frac{g(t, s)}{|g(t, s)|} K_{1}\left(\omega_{2}|g(t, s)|\right) \rho(s) \mathrm{d} s .
$$

The exponential decay (5.1.1) of $\gamma^{\prime \prime}$ implies that

$$
\left|\left(\partial_{t}+\partial_{s}\right) g(t, s)\right|+\left|\left(\partial_{t}+\partial_{s}\right) \hat{n}(s)\right| \leq C e^{-\tilde{\beta} \min \{s, t\} 1_{\{s>0\}}}
$$

for some positive constants $C$ and $\tilde{\beta}$. Since $K_{1}^{\prime}\left(\omega_{2}|g(t, s)|\right)$ decays exponentially in $|t-s|$ with $\left\|K_{1}\left(\omega_{2}|g(t, \cdot)|\right)\right\|_{L_{1}}+\left\|K_{1}^{\prime}\left(\omega_{2}|g(t, \cdot)|\right)\right\|_{L_{1}} \leq C$ uniformly in $t \geq t_{0}$, it follows (as in the
proof of Theorem 5.2.1) that

$$
\begin{align*}
&\left|\left(\partial_{t}-i E\right) D_{\omega_{2}}\left[\rho_{1}\right](\gamma(t)+r \hat{n}(t))\right| \\
& \leq \frac{1}{2 \pi}\left|\int_{\mathbb{R}}\left(\partial_{t}+\partial_{s}\right)\left(\hat{n}(s) \cdot \frac{g(t, s)}{|g(t, s)|} K_{1}\left(\omega_{2}|g(t, s)|\right)\right) \rho_{1}(s) \mathrm{d} s\right|  \tag{5.5.4}\\
& \quad+\frac{1}{2 \pi}\left|\int_{\mathbb{R}} \hat{n}(s) \cdot \frac{g(t, s)}{|g(t, s)|} K_{1}\left(\omega_{2}|g(t, s)|\right)\left(\partial_{s}-i E\right) \rho_{1}(s) \mathrm{d} s\right|,
\end{align*}
$$

with both terms on the right-hand side going to 0 as $t \rightarrow \infty$. Similarly, the fact that $\rho_{0} \in L^{1}$ directly implies that $\left(\partial_{t}-i E\right) D_{\omega_{2}}\left[\rho_{0}\right](\gamma(t)+r \hat{n}(t)) \rightarrow 0$ as $t \rightarrow \infty$. The limit $t \rightarrow-\infty$ is treated similarly, as is the case $r<0$. This completes the proof.

Remark 5.5.1. The solution $u$ can actually satisfy a stronger outgoing condition than (5.1.6). It is possible to show that

$$
\begin{equation*}
\lim _{t \rightarrow \pm \infty} \sup _{r \in[-1,1]}\left|\left( \pm \partial_{t}-i E\right) u(\gamma(t)+r \hat{n}(t))\right|=0 \tag{5.5.5}
\end{equation*}
$$

but this requires regularity of $\rho$ (when $m_{1}=m_{2}$ ) or $\sigma$ (when $m_{1} \neq m_{2}$ ). More specifically, the left-hand side of (5.5.3) and $R(t, s):=C_{1} \omega K_{1}(\omega|\gamma(t)+r \hat{n}(t)-\gamma(s)|)$ can instead be bounded above by $C_{1}|t-s| e^{-\tilde{\beta} \min \{t, s\} 1_{\{s>0\}}}$ and $\frac{C_{2}}{\sqrt{|t-s|^{2}+r^{2}}} e^{-\eta|t-s|}$, respectively, where the positive constants $C_{1}, C_{2}, \tilde{\beta}, \eta$ are independent of $r$. Thus all bounds after (5.5.3) can be shown to hold uniformly in $r$. The only obstruction to obtaining (5.5.5) is (5.5.1), due to the singularity of $K_{1}=-K_{0}^{\prime}$ at zero. But if $\rho$ were (weakly) differentiable with $\rho^{\prime}(t) \rightarrow 0$ as $|t| \rightarrow \infty$, then $\mu:=\mathcal{P} \rho$ would satisfy $\lim _{t \rightarrow \pm \infty}\left(\partial_{t} \mp i E\right) \mu(t)=0$, meaning that all the bounds starting from (5.5.3) would hold with $Q \rho$ replaced by $\mu$. Thus we would obtain (5.5.5) in this case. Similarly, if $\sigma$ is weakly differentiable with $\sigma^{\prime}(t) \rightarrow 0$ as $|t| \rightarrow \infty$, then (5.5.5) holds for the $m_{1} \neq m_{2}$ case.

To guarantee this extra regularity of $\rho$ and $\sigma$, we would need to show that $\mathcal{L P}$ and $\mathcal{L}_{2} \mathcal{P}_{2}$ are invertible on $H_{\alpha}^{1}:=\left\{\rho \in L_{\alpha}^{2}: \rho^{\prime} \in L_{\alpha}^{2}\right\}$. Such results are natural extensions of the ones
presented in Section 5.2.1 for $L_{\alpha}^{2}$, and their proofs would likely use similar techniques to the ones presented below. However, we do not pursue this issue further here.

The next part of this section is devoted to proving Theorem 5.2.5. As in the statement of the theorem, $m_{0}>0$ and $E_{0} \in\left(-m_{0}, m_{0}\right) \backslash\{0\}$ will be fixed, with $\omega_{0}:=\sqrt{m_{0}^{2}-E_{0}^{2}}$. The parameters $m$ and $E$ of our integral equation (5.2.3) will depend on the parameter $\lambda \in[1, \infty)$ via the relations $m=\lambda m_{0}$ and $E=\lambda E_{0}$. The constant $\alpha_{*}$ is defined above Theorem 5.2.5. Recall the definitions of $\mathcal{L}, \mathcal{P}, \mathcal{L}_{0}$ and $\mathcal{M}$ in (5.2.2), (5.2.12) and (5.2.13). We will use $\|\cdot\|$ to denote the operator norm. More specifically, if $\mathcal{B}_{1}$ and $\mathcal{B}_{2}$ are Banach spaces and $\mathcal{A}: \mathcal{B}_{1} \rightarrow \mathcal{B}_{2}$ is a linear operator, then $\|\mathcal{A}\|:=\sup \left\{\|\mathcal{A} \psi\|_{2}:\|\psi\|_{1}=1\right\}$, where $\|\cdot\|_{j}$ denotes the norm on $\mathcal{B}_{j}$.

We begin with the following important lemmas.
Lemma 5.5.2. For all $\alpha \in\left(0, \alpha_{*}\right]$, the operator $\left(\mathcal{L}_{0} \mathcal{P}\right)^{-1}$ is bounded on $L_{\alpha}^{2}$ with $\left\|\left(\mathcal{L}_{0} \mathcal{P}\right)^{-1}\right\| \leq$ $C$ uniformly in $\lambda \in[1, \infty)$.

Proof. Fix $\alpha \in\left(0, \alpha_{*}\right]$. By (5.2.20) we know that $\left(\mathcal{L}_{0} \mathcal{P}\right)^{-1}=\left(1+R_{0}\right)\left(1+R_{1}\right)$, where $R_{0}$ and $R_{1}$ are convolutions by $r_{0}:=m K_{0}(\omega|\cdot|)$ and $r_{1}:=\frac{(i-2) m^{2}}{2 \zeta} e^{i \zeta|\cdot|}$, respectively, with $\zeta:=a_{+}+i a_{-}$and

$$
a_{ \pm}:=\sqrt{\frac{\sqrt{4 m_{0}^{4}+E_{0}^{4}} \pm E_{0}^{2}}{2}}
$$

Thus there exists $r \in L^{1}$ such that $e^{\alpha|\cdot|} r(\cdot) \in L^{1}$ and $\left|r_{j}(t)\right| \leq \lambda r(\lambda(t))$ for all $\lambda \in[1, \infty)$ and $j \in\{1,2\}$. Using the identity $\left\|g_{1} * g_{2}\right\|_{L^{2}} \leq\left\|g_{1}\right\|_{L^{1}}\left\|g_{2}\right\|_{L^{2}}$ with $g_{1}(t)=\lambda e^{\alpha|t|} r(\lambda|t|)$ and $g_{2}(t)=e^{\alpha|t|}|\rho|(t)$, we obtain that for $\rho \in L_{\alpha}^{2}$ and $j \in\{1,2\}$,

$$
\begin{aligned}
\left\|R_{j} \rho\right\|_{L_{\alpha}^{2}}^{2} & \leq \int_{\mathbb{R}} e^{2 \alpha|t|}\left(\int_{\mathbb{R}} \lambda r\left(\lambda\left|t-t^{\prime}\right|\right) \rho\left(t^{\prime}\right) \mathrm{d} t^{\prime}\right)^{2} \mathrm{~d} t \\
& \leq \int_{\mathbb{R}}\left(\int_{\mathbb{R}} \lambda e^{\alpha\left|t-t^{\prime}\right|} r\left(\lambda\left|t-t^{\prime}\right|\right) e^{\alpha\left|t^{\prime}\right|}|\rho|\left(t^{\prime}\right) \mathrm{d} t^{\prime}\right)^{2} \mathrm{~d} t
\end{aligned}
$$

is bounded uniformly in $\lambda \in[1, \infty)$. This completes the proof.
Lemma 5.5.3. For all $\alpha \in\left(0, \alpha_{*}\right]$ and $\lambda \in[1, \infty)$, the operator $\mathcal{M}$ is Hilbert-Schmidt (hence compact) on $L_{\alpha}^{2}$.

Proof. Fix $\alpha \in\left(0, \alpha_{*}\right]$ and $\lambda \in[1, \infty)$, and note that $L_{\alpha}^{2} \subset L^{1} \cap L^{2}$. Define $A_{1}: L^{2} \rightarrow$ $L_{\alpha}^{2}, A_{2}: L^{2} \rightarrow L^{2}, A_{3}: L_{\alpha}^{2} \rightarrow L^{2}$ by

$$
A_{1}=W_{\alpha}^{-1}, \quad A_{2}=W_{\alpha}\left(\mathcal{L}_{0}-\mathcal{L}\right) W_{\alpha}, \quad A_{3}=W_{\alpha}^{-1} \mathcal{P}
$$

where $W_{\alpha}$ is point-wise multiplication by $w_{\alpha}$. Our strategy is to bound each term on the right-hand side of $\mathcal{M}=\left(\mathcal{L}_{0} \mathcal{P}\right)^{-1} A_{1} A_{2} A_{3}$. We immediately have $\left\|A_{1} \rho\right\|_{L_{\alpha}^{2}}=\|\rho\|_{L^{2}}$, so $A_{1}$ is bounded. Using that

$$
\begin{equation*}
|\mathcal{P} \rho|(t) \leq|\rho|(t)+\frac{m^{2}}{|E|}\|\rho\|_{L^{1}} \leq|\rho|(t)+C\|\rho\|_{L_{\alpha}^{2}} \tag{5.5.6}
\end{equation*}
$$

we have $\left\|A_{3} \rho\right\|_{L^{2}} \leq C\|\rho\|_{L_{\alpha}^{2}}$, hence $A_{3}$ is also bounded. By Lemma 5.5.2, $\left(\mathcal{L}_{0} \mathcal{P}\right)^{-1}$ is bounded on $L_{\alpha}^{2}$.

It remains to show that $A_{2}$ is Hilbert-Schmidt. To do so, we will prove exponential decay of the kernel of $\mathcal{L}_{0}-\mathcal{L}$ using the decay properties of $K_{0}$ and the fact that our interface rapidly converges to a straight line at infinity (5.1.1). Let $\ell_{\Delta}(t, s)=K_{0}(\omega|\gamma(t)-\gamma(s)|)-K_{0}(\omega|t-s|)$ so that $-\frac{m}{\pi} \ell_{\Delta}(t, s)$ is the kernel of $\mathcal{L}-\mathcal{L}_{0}$. Since $K_{0}$ and $K_{1}=-K_{0}^{\prime}$ are both positive and monotonically decreasing, and $|\gamma(t)-\gamma(s)| \leq|t-s|$, it follows that

$$
0 \leq \ell_{\Delta}(t, s)=\int_{\omega|\gamma(t)-\gamma(s)|}^{\omega|t-s|} K_{1}(z) \mathrm{d} z \leq \omega(|t-s|-|\gamma(t)-\gamma(s)|) K_{1}(\omega|\gamma(t)-\gamma(s)|) .
$$

Hence

$$
\begin{equation*}
0 \leq \ell_{\Delta}(t, s) \leq \omega(|t-s|-|\gamma(t)-\gamma(s)|) K_{1}(c \omega|t-s|), \tag{5.5.7}
\end{equation*}
$$

where $c>0$ is defined by (5.1.3). The factor $K_{1}(c \omega|t-s|)$ decays exponentially in $|t-s|$ but has a singularity at $t=s$. To show that $\ell_{\Delta}$ is bounded and decays exponentially in $|t|+|s|$, we will control the difference $|t-s|-|\gamma(t)-\gamma(s)|$.

By smoothness of $\gamma$, for every $s, t \in \mathbb{R}$ there exist $r_{1}, r_{2} \in[\min \{s, t\}, \max \{s, t\}]$ such that

$$
\begin{equation*}
\gamma(t)=\gamma(s)+(t-s) \gamma^{\prime}(s)+\frac{1}{2}(t-s)^{2} \gamma^{\prime \prime}(s)+\frac{1}{3!}(t-s)^{3}\left(\gamma_{1}^{\prime \prime \prime}\left(r_{1}\right), \gamma_{2}^{\prime \prime \prime}\left(r_{2}\right)\right) \tag{5.5.8}
\end{equation*}
$$

Since $\left|\gamma^{\prime}\right| \equiv 1$ and thus $\gamma^{\prime} \cdot \gamma^{\prime \prime} \equiv 0$, we have

$$
\gamma^{\prime}(s) \cdot(\gamma(t)-\gamma(s))=t-s+\frac{1}{3!}(t-s)^{3} \gamma^{\prime}(s) \cdot\left(\gamma_{1}^{\prime \prime \prime}\left(r_{1}\right), \gamma_{2}^{\prime \prime \prime}\left(r_{2}\right)\right)
$$

Multiplying both sides by $t-s$ and using (5.5.8) to expand $(t-s) \gamma^{\prime}(s)$, it follows that

$$
\begin{aligned}
|\gamma(t)-\gamma(s)|^{2}=(t-s)^{2}+\frac{1}{3!}(t-s)^{4} \gamma^{\prime}(s) \cdot & \left(\gamma_{1}^{\prime \prime \prime}\left(r_{1}\right), \gamma_{2}^{\prime \prime \prime}\left(r_{2}\right)\right)+\frac{1}{2}(t-s)^{2} \gamma^{\prime \prime}(s)(\gamma(t)-\gamma(s)) \\
& +\frac{1}{3!}(t-s)^{3}\left(\gamma_{1}^{\prime \prime \prime}\left(r_{1}\right), \gamma_{2}^{\prime \prime \prime}\left(r_{2}\right)\right) \cdot(\gamma(t)-\gamma(s))
\end{aligned}
$$

Again using (5.5.8) to expand $\gamma(t)-\gamma(s)$ on the bottom line, we obtain that

$$
|\gamma(t)-\gamma(s)|^{2} \leq(t-s)^{2}+C J(s, t)(t-s)^{4}, \quad J(s, t):=\left|\gamma^{\prime \prime}\right|_{s, t}^{2}+\left|\gamma^{\prime \prime \prime}\right|_{s, t}
$$

for some $C>0$, where the semi-norm $|\cdot|_{s, t}$ is defined by

$$
|\eta|_{s, t}:=\|\eta\|_{L^{\infty}[\min \{s, t\}, \max \{s, t\}]}
$$

This implies $0 \leq|t-s|-|\gamma(t)-\gamma(s)| \leq C J(s, t)|t-s|^{3}$. From the rapid decay (5.1.1) of
$\left|\gamma^{\prime \prime}\right|^{2}+\left|\gamma^{\prime \prime \prime}\right|$ at infinity, it follows that

$$
0 \leq|t-s|-|\gamma(t)-\gamma(s)| \leq \frac{C_{0}}{\sqrt{2}}|t-s|^{3} e^{-\beta \chi(s, t)}, \quad \chi(s, t):= \begin{cases}\min \{|s|,|t|\}, & \text { st }>0 \\ 0, & \text { else. }\end{cases}
$$

Using (5.5.7), we have

$$
\begin{equation*}
0 \leq \ell_{\Delta}(t, s) \leq \frac{\omega}{\sqrt{2}} C_{0}|t-s|^{3} e^{-\beta \chi(s, t)} K_{1}(c \omega|t-s|) \tag{5.5.9}
\end{equation*}
$$

By definition of $\alpha_{*}$, we conclude that

$$
\int_{\mathbb{R}^{2}} w_{\alpha}^{2}(t) w_{\alpha}^{2}(s) \ell_{\Delta}^{2}(t, s) \mathrm{d} t \mathrm{~d} s<\infty
$$

This means $A_{2}$ is Hilbert-Schmidt and the proof is complete.

Henceforth, $\|\cdot\|_{2}$ denotes the Hilbert-Schmidt norm.

Remark 5.5.4. Since the constant $C$ in (5.5.6) is proportional to $1 /|E|$, we have shown that $\|\mathcal{M}\|_{2} \leq C /|E|$ as $E \rightarrow 0$. The singularity at $E=0$ should not be surprising, as outgoing and incoming conditions at infinity are the same in this case. It turns out that an appropriate linear combination of outgoing and incoming conditions produces an operator that behaves better in the $E \rightarrow 0$ limit. Indeed, we could have instead defined $Q: L_{\alpha}^{2} \rightarrow L^{\infty}$ by $Q \rho(t)=\frac{i}{2 E} \int_{-\infty}^{\infty}\left(e^{i E\left|t-t^{\prime}\right|}-e^{-i E\left|t-t^{\prime}\right|}\right) \rho\left(t^{\prime}\right) \mathrm{d} t^{\prime}$, which is bounded uniformly in $E$ (for any $\alpha>0$ ). The resulting solution would now look like $\sin (E|t|)$ as $|t| \rightarrow \infty$.

Lemma 5.5.3 implies that if (5.2.13) does not have a unique solution $\rho$, then the kernel of $1+\mathcal{M}$ is a nontrivial finite-dimensional subspace of $L_{\alpha}^{2}$. An extension of Lemma 5.5.3 is the following

Proposition 5.5.5. For any $\alpha \in\left(0, \alpha_{*}\right],\|\mathcal{M}\|_{2} \leq C / \sqrt{\lambda}$ as $\lambda \rightarrow \infty$.

Proof. Suppose $\lambda \geq 1$. As in the proof of Lemma 5.5.3, we will bound each factor on the right-hand side of $\mathcal{M}=\left(\mathcal{L}_{0} \mathcal{P}\right)^{-1} A_{1} A_{2} A_{3}$. By Lemma 5.5.2, the operator norm of $\left(\mathcal{L}_{0} \mathcal{P}\right)^{-1}$ on $L_{\alpha}^{2}$ is bounded uniformly in $\lambda$. From (5.5.6) it is clear that $\left\|A_{3}\right\| \leq C \lambda$. Since $A_{1}$ is independent of $\lambda$, it remains to bound $\left\|A_{2}\right\|_{2}$.

The kernel of $A_{2}$ is $k(t, s)=\frac{m}{\pi} w_{\alpha}(t) \ell_{\Delta}(t, s) w_{\alpha}(s)$, with $\ell_{\Delta}(t, s)=K_{0}(\omega|\gamma(t)-\gamma(s)|)-$ $K_{0}(\omega|t-s|)$ as before. From (5.5.9) it follows that

$$
k^{2}(t, s) \leq\left(\frac{m \omega C_{0}}{\sqrt{2} \pi}\right)^{2} w_{\alpha}^{2}(t) w_{\alpha}^{2}(s)(t-s)^{6} e^{-2 \beta \chi(s, t)} K_{1}^{2}(c \omega|t-s|)
$$

Performing the change of variables $(\xi, \zeta):=(t-s, t+s)$ and using the fact that $|\zeta+\xi|+|\zeta-\xi| \leq$ $|\zeta+\lambda \xi|+|\zeta-\lambda \xi|$ and $\chi\left(\frac{\zeta-\lambda \xi}{2}, \frac{\zeta+\lambda \xi}{2}\right) \leq \chi\left(\frac{\zeta-\xi}{2}, \frac{\zeta+\xi}{2}\right)$ for all $\lambda \geq 1$ and $\xi, \zeta \in \mathbb{R}$, it follows that

$$
\begin{aligned}
\|k\|_{L^{2}}^{2} & \leq \frac{1}{2}\left(\frac{m \omega C_{0}}{\sqrt{2} \pi}\right)^{2} \int_{\mathbb{R}^{2}} w_{\alpha}^{2}\left(\frac{\zeta+\lambda \xi}{2}\right) w_{\alpha}^{2}\left(\frac{\zeta-\lambda \xi}{2}\right) \xi^{6} e^{-2 \beta \chi\left(\frac{\zeta-\lambda \xi}{2}, \frac{\zeta+\lambda \xi}{2}\right)} K_{1}^{2}\left(c \omega_{0} \lambda|\xi|\right) \mathrm{d} \zeta \mathrm{~d} \xi \\
& =\frac{1}{2 \lambda^{7}}\left(\frac{m \omega C_{0}}{\sqrt{2} \pi}\right)^{2} \int_{\mathbb{R}^{2}} w_{\alpha}^{2}\left(\frac{\zeta+\xi}{2}\right) w_{\alpha}^{2}\left(\frac{\zeta-\xi}{2}\right) \xi^{6} e^{-2 \beta \chi\left(\frac{\zeta-\xi}{2}, \frac{\zeta+\xi}{2}\right)} K_{1}^{2}\left(c \omega_{0}|\xi|\right) \mathrm{d} \zeta \mathrm{~d} \xi
\end{aligned}
$$

Since the above integral is finite (by definition of $\alpha_{*}$ ) and independent of $\lambda$, we have shown that $\|k\|_{L^{2}} \leq C / \lambda^{3 / 2}$ for all $\lambda$ sufficiently large. Thus

$$
\|\mathcal{M}\|_{2} \leq\left\|\left(\mathcal{L}_{0} \mathcal{P}\right)^{-1}\right\|\left\|A_{1}\right\|\left\|A_{2}\right\|_{2}\left\|A_{3}\right\| \leq C / \sqrt{\lambda}
$$

and the proof is complete.

The above result implies that $\|\mathcal{M}\| \rightarrow 0$ as $\lambda \rightarrow \infty$. However, it is possible to get faster decay of $\|\mathcal{M}\|$ in $\lambda$.

Proposition 5.5.6. For any $\alpha \in\left(0, \alpha_{*}\right],\|\mathcal{M}\| \leq C / \lambda$ as $\lambda \rightarrow \infty$.

Proof. Recall Lemma 5.5.2, which states that $\left\|\left(\mathcal{L}_{0} \mathcal{P}\right)^{-1}\right\| \leq C$ uniformly in $\lambda$. Thus it
remains to bound the norm of $\left(\mathcal{L}-\mathcal{L}_{0}\right) \mathcal{P}$. We will bound the terms $\mathcal{L}-\mathcal{L}_{0}$ and $\left(\mathcal{L}-\mathcal{L}_{0}\right) Q$ separately.

We begin with the latter. We have that $\|Q \rho\|_{\infty} \leq \frac{C m^{2}}{E}\|\rho\|_{L_{\alpha}^{2}}$, thus for $\left(\mathcal{L}-\mathcal{L}_{0}\right) Q$ it suffices to bound $\mathcal{L}-\mathcal{L}_{0}$ from $L^{\infty}$ to $L_{\alpha}^{2}$. The kernel of $\mathcal{L}-\mathcal{L}_{0}$ is $-\frac{m}{\pi} \ell_{\Delta}(t, s)$, where $\ell_{\Delta}(t, s)=K_{0}(\omega|\gamma(t)-\gamma(s)|)-K_{0}(\omega|t-s|)$. Using (5.5.9), it follows that for $\rho \in L^{\infty}$ we have

$$
\begin{aligned}
\left|\left(\mathcal{L}-\mathcal{L}_{0}\right) \rho\right|(t) & \leq \frac{m}{\pi}\|\rho\| \int_{\mathbb{R}} \ell_{\Delta}(t, s) \mathrm{d} s \\
& \leq C\|\rho\| \lambda^{2} \int_{\mathbb{R}}|t-s|^{3} e^{-\beta \chi(s, t)} K_{1}(c \omega|t-s|) \mathrm{d} s \\
& =: C\|\rho\| \lambda^{2} I,
\end{aligned}
$$

where $\|\rho\|:=\|\rho\|_{\infty}$. For concreteness, suppose $t \geq 0$. This means

$$
\chi(s, t)= \begin{cases}0, & s<0 \\ s, & 0 \leq s \leq t \\ t, & s>t\end{cases}
$$

so that

$$
I=\int_{-\infty}^{0} \kappa_{\omega}(t-s) \mathrm{d} s+\int_{0}^{t} \kappa_{\omega}(t-s) e^{-\beta s} \mathrm{~d} s+\int_{t}^{\infty} \kappa_{\omega}(t-s) e^{-\beta t} \mathrm{~d} s=: I_{1}+I_{2}+I_{3}
$$

where we have defined $\kappa_{\omega}(x):=|x|^{3} K_{1}(c \omega|x|)$. Recall that by definition of $\alpha_{*}$, we have
$0<\alpha_{*}<\beta$ and $\kappa_{1}(x) \leq C e^{-\alpha_{*}|x|}$ for all $x \in \mathbb{R}$. Thus

$$
\begin{aligned}
I_{1} & \leq C \omega^{-4} \int_{t}^{\infty} e^{-\alpha_{*} s} \mathrm{~d} s=\frac{C}{\alpha_{*} \omega^{4}} e^{-\alpha_{*} t}=C_{1} \omega^{-4} e^{-\alpha_{*} t} \\
I_{2} & =\int_{0}^{t} \kappa_{\omega}(s) e^{-\beta(t-s)} \mathrm{d} s \leq C \omega^{-4} \int_{0}^{t} e^{-\alpha_{*} s} e^{-\beta(t-s)} \mathrm{d} s \\
& \leq \frac{C}{\left(\beta-\alpha_{*}\right) \omega^{4}} e^{-\beta t}\left(e^{\left(\beta-\alpha_{*}\right) t}-1\right) \leq C_{2} \omega^{-4} e^{-\alpha_{*} t} \\
I_{3} & =e^{-\beta t} \int_{-\infty}^{0} \kappa_{\omega}(s) \mathrm{d} s \leq \frac{C}{\alpha_{*} \omega^{4}} e^{-\beta t}=C_{3} \omega^{-4} e^{-\beta t} \leq C_{3} \omega^{-4} e^{-\alpha_{*} t}
\end{aligned}
$$

uniformly in $t \geq 0$, for some positive constants $C_{1}, C_{2}$ and $C_{3}$. Repeating the same argument for $t<0$ and recalling that $\alpha<\alpha_{*}$, it follows that $\left\|\left(\mathcal{L}-\mathcal{L}_{0}\right) \rho\right\|_{L_{\alpha}^{2}} \leq \frac{C\|\rho\|}{\lambda^{2}}$. Thus we have shown that $\left\|\left(\mathcal{L}-\mathcal{L}_{0}\right) Q\right\| \leq C / \lambda$.

It remains to bound $\mathcal{L}-\mathcal{L}_{0}$ from $L_{\alpha}^{2}$ to itself. We already showed in the proof of Proposition 5.5.5 that $A_{2}=W_{\alpha}\left(\mathcal{L}_{0}-\mathcal{L}\right) W_{\alpha}$ is Hilbert-Schmidt on $L^{2}$ with $\left\|A_{2}\right\|_{2} \leq C / \lambda^{3 / 2}$. This means $\left\|\mathcal{L}-\mathcal{L}_{0}\right\|_{L_{\alpha}^{2} \rightarrow L_{\alpha}^{2}} \leq C / \lambda^{3 / 2}$ and the proof is complete.

Finally, we prove the first well-posedness result of Section 5.2.1.

Proof of Theorem 5.2.5. Fix $\alpha \in\left(0, \alpha_{*}\right]$ and $\delta \in(0, \alpha)$.
Since $\mathcal{M}: L_{\alpha}^{2} \rightarrow L_{\alpha}^{2}$ is holomorphic in $\lambda \in[1, \infty) \times(-\delta, \delta)$, we apply Lemma 5.5.3, Proposition 5.5.6 and Kato perturbation theory [58, Theorem VII.1.9] to prove that $\mathcal{M}$ has an eigenvalue of -1 for only a finite number of $\lambda \in[1, \infty)$. The result then follows from (5.2.13).

As opposed to the above, Theorem 5.2.7 does not require $\mathcal{M}$ to be small in any limit. Instead, we show that -1 belongs to the resolvent set of $\mathcal{M}$ whenever $\Im E>0$. To this end, we state the following

Lemma 5.5.7. Fix $m>0$. For all $E \in \mathbb{C}$ with $0<|\Re E|<m$ and $\Im E \neq 0$, if $u \in H^{1}\left(\mathbb{R}^{2}\right)$ solves (5.5.10) below, then $u \equiv 0$.

Proof. Integrating by parts, we see that

$$
\int_{\Omega_{j}}|\nabla u|^{2} \mathrm{~d} x=-\omega^{2} \int_{\Omega_{j}}|u|^{2} \mathrm{~d} x+\varepsilon_{j} \int_{\Gamma} u \hat{n} \cdot \nabla u \mathrm{~d} t
$$

where $\varepsilon_{1}=1$ and $\varepsilon_{2}=-1$. Taking the sum over $j$ and using the second line of (5.5.10), we obtain

$$
\int_{\mathbb{R}^{2}}|\nabla u|^{2} \mathrm{~d} x+\omega^{2} \int_{\mathbb{R}^{2}}|u|^{2} \mathrm{~d} x=2 m \int_{\Gamma}|u|^{2} \mathrm{~d} t
$$

Taking the imaginary part of both sides, we conclude that $\int_{\mathbb{R}^{2}}|u|^{2} \mathrm{~d} x=0$ and the result is complete.

Using the above lemma, we now prove the second well-posedness result from Section 5.2.1.

Proof of Theorem 5.2.7. Fix $\varepsilon>0$ and for $\alpha>0$ define $Z_{\alpha, \varepsilon}:=([-m+\varepsilon,-\varepsilon] \cup[\varepsilon, m-\varepsilon]) \times$ $(-\alpha / 2, \alpha / 2) \subset \mathbb{C}$. The same arguments as above show that $\mathcal{M}: L_{\alpha}^{2} \rightarrow L_{\alpha}^{2}$ is Hilbert-Schmidt and holomorphic in $E \in Z_{\alpha, \varepsilon}$ for all $\alpha>0$ sufficiently small.

Fix $E \in Z_{\alpha, \varepsilon}$ such that $\Im E>0$. By contradiction suppose $\mathcal{M}$ has an eigenvalue of -1 . Then there exists $0 \neq \rho \in L_{\alpha}^{2}$ such that $(1+\mathcal{M}) \rho=0$. Letting $\mu:=\mathcal{P} \rho$, it follows that $u=S_{\omega}[\mu]$ solves the homogeneous problem

$$
\left\{\begin{array}{l}
\Delta u(x)-\omega^{2} u(x)=0, \quad x \in \mathbb{R}^{2} \backslash \Gamma  \tag{5.5.10}\\
{[[\hat{n} \cdot \nabla u]](\gamma(t))=-2 m u(\gamma(t)), \quad t \in \mathbb{R}} \\
{[[u]](\gamma(t))=0, \quad t \in \mathbb{R}}
\end{array}\right.
$$

where $\omega^{2}=m^{2}-E^{2}$. Since $E$ has positive imaginary part, $Q$ is bounded from $L_{\alpha}^{2}$ to itself, meaning that $\mu \in L_{\alpha}^{2}$. It follows that $u$ and $\nabla u$ decay rapidly at infinity; for example, $u \in H^{1}\left(\mathbb{R}^{2}\right)$. Lemma 5.5 .7 then implies that $S_{\omega}[\mu] \equiv 0$; hence $\mu \equiv 0$ by (5.1.12). It follows
from (5.2.18) that $0=\tilde{\mu}(\xi)=\left(1-\frac{2 i m^{2}}{\xi^{2}-E^{2}}\right) \tilde{\rho}(\xi)$, from which we conclude that $\rho \equiv 0$, a contradiction. Indeed, if $\alpha$ is chosen sufficiently small, then 1 is not in the image of the $\operatorname{map} \mathbb{R} \ni \xi \mapsto \frac{2 i m^{2}}{\xi^{2}-E^{2}}$. We have thus shown that -1 is not an eigenvalue of $\mathcal{M}$ whenever $E \in Z_{\alpha, \varepsilon}$ has positive imaginary part. This means we can again apply [58, Theorem VII.1.9] to complete the result.

Remark 5.5.8. It is crucial that $Z_{\alpha, \varepsilon}$ be bounded away from 0 as $Q$ is not holomorphic there. Thus we cannot guarantee that the number of "bad" E-values near 0 is finite. As suggested by Remark 5.5.4, one could redefine $Q$ by $Q \rho(t)=\frac{i}{2 E} \int_{-\infty}^{\infty}\left(e^{i E\left|t-t^{\prime}\right|}-e^{-i E\left|t-t^{\prime}\right|}\right) \rho\left(t^{\prime}\right) \mathrm{d} t^{\prime}$ to make it holomorphic, but then the outgoing conditions (5.1.6) would not be satisfied.

## APPENDIX A <br> PSEUDO-DIFFERENTIAL CALCULUS

## A. 1 Notation and functional setting

Given a bounded linear operator $A: \mathcal{H} \rightarrow \mathcal{H}$ for $\mathcal{H}$ a Hilbert space, we denote by $A^{*}$ its adjoint and $\|A\|$ its operator norm. If, in addition, $A^{*} A$ is compact, then by the spectral theorem, $A^{*} A$ admits a countable collection of eigenvalues $\left\{\lambda_{j}\right\} \subset[0, \infty)$ converging to 0. The operator $A$ is Hilbert-Schmidt if $\|A\|_{2}:=\sum_{j} \lambda_{j}<\infty$ and trace-class if $\|A\|_{1}:=$ $\sum_{j} \sqrt{\lambda_{j}}<\infty$. If $A$ is trace-class, we define the trace of $A$ by

$$
\operatorname{Tr} A:=\sum_{j \in \mathbb{N}}\left(\psi_{j}, A \psi_{j}\right),
$$

where $\left\{\psi_{j}\right\}_{j \in \mathbb{N}}$ is any Hilbert basis of $\mathcal{H}$ (the trace is independent of the chosen Hilbert basis).

Weyl Quantization. See [34, Chapter 7]. Let $\mathcal{S}\left(\mathbb{R}^{d}\right) \otimes \mathbb{C}^{n}$ be the Schwartz space of vectorvalued functions and $\mathcal{S}^{\prime}\left(\mathbb{R}^{d}\right) \otimes \mathbb{C}^{n}$ its dual. Given a parameter $h \in(0,1]$ and a symbol $a(x, \xi ; h)=a \in \mathcal{S}^{\prime}\left(\mathbb{R}^{d} \times \mathbb{R}^{d}\right) \otimes \mathbb{C}^{n \times n}$, we define the Weyl quantization of $a$ as the operator

$$
\begin{equation*}
\mathrm{Op}_{h}(a) \psi(x):=\frac{1}{(2 \pi h)^{d}} \int_{\mathbb{R}^{2 d}} e^{i(x-y) \cdot \xi / h} a\left(\frac{x+y}{2}, \xi ; h\right) \psi(y) d y d \xi, \quad \psi \in \mathcal{S}\left(\mathbb{R}^{d}\right) \otimes \mathbb{C}^{n} \tag{A.1.1}
\end{equation*}
$$

When $a$ is polynomial in $\xi$, it follows that $\operatorname{Op}_{h}(a)$ is a differential operator. We denote by $\mathrm{Op} a=\mathrm{Op}_{1} a$ for $h=1$.

Any bounded matrix-valued operator $A$ from $\mathcal{S}\left(\mathbb{R}^{d}\right) \otimes \mathbb{C}^{n}$ to $\mathcal{S}^{\prime}\left(\mathbb{R}^{d}\right) \otimes \mathbb{C}^{n}$ admits a

Schwartz kernel $k_{A} \in \mathcal{S}^{\prime}\left(\mathbb{R}^{d} \times \mathbb{R}^{d}\right) \otimes \mathbb{C}^{n \times n}$ such that

$$
A \psi(x)=\int_{\mathbb{R}^{d}} k_{A}(x, y) \psi(y) d y, \quad \psi \in \mathcal{S}\left(\mathbb{R}^{d}\right) \otimes \mathbb{C}^{n}
$$

If $A=\operatorname{Op}(a)$, then

$$
a(x, \xi)=\int_{\mathbb{R}^{d}} e^{-i y \cdot \xi} k_{A}\left(x+\frac{y}{2}, x-\frac{y}{2}\right) d y
$$

Order functions and symbol classes. See [20] and [34, Chapter 7]. For $(x, \xi)=X \in \mathbb{R}^{2 d}$, we define $\langle X\rangle:=\sqrt{1+|X|^{2}}$. A function $\mathfrak{m}: \mathbb{R}^{2 d} \rightarrow[0, \infty)$ is called an order function if there exist constants $C_{0}>0, N_{0}>0$ such that $\mathfrak{m}(X) \leq C_{0}\langle X-Y\rangle^{N_{0}} \mathfrak{m}(Y)$ for all $X, Y \in \mathbb{R}^{2 d}$. Note that $\langle X\rangle^{p}$ and $\left\langle X_{ \pm}\right\rangle$are order functions for all $p \in \mathbb{R}$, where $X_{+}:=\max \{X, 0\}$ (with the max defined element-wise) and $X_{-}:=-(-X)_{+}$. Moreover, if $\mathfrak{m}_{1}$ and $\mathfrak{m}_{2}$ are order functions, then so is $\mathfrak{m}_{1} \mathfrak{m}_{2}$.

We say that $a \in S(\mathfrak{m})$ if for every $\alpha \in \mathbb{N}^{2 d}$, there exists $C_{\alpha}>0$ such that $\left|\partial^{\alpha} a(X ; h)\right| \leq$ $C_{\alpha} \mathfrak{m}(X)$ for all $X \in \mathbb{R}^{2 d}$ and $h \in(0,1]$. We write $S\left(\mathfrak{m}^{-\infty}\right)$ to denote the intersection over $s \in \mathbb{N}$ of $S\left(\mathfrak{m}^{-s}\right)$. For $\delta \in[0,1]$ and $k \in \mathbb{R}$, we say that $a(X ; h) \in S_{\delta}^{k}(\mathfrak{m})$ if for every $\alpha \in \mathbb{N}^{2 d}$, there exists $C_{\alpha}>0$ such that

$$
\begin{equation*}
\left|\partial^{\alpha} a(X ; h)\right| \leq C_{\alpha} \mathfrak{m}(X) h^{-\delta|\alpha|-k} \tag{A.1.2}
\end{equation*}
$$

uniformly in $X \in \mathbb{R}^{2 d}$ and $h \in(0,1]$. If either $k$ or $\delta$ are omitted, they are assumed to be zero. We will always write the order function $\mathfrak{m}$ when using these symbol classes.

By [34, Chapter 7], we know that if $a \in S\left(\mathfrak{m}_{1}\right)$ and $b \in S\left(\mathfrak{m}_{2}\right)$, then $\operatorname{Op}_{h}(c):=$ $\mathrm{Op}_{h}(a) \mathrm{Op}_{h}(b)$ is a pseudo-differential operator, with

$$
c(x, \xi ; h)=(a \sharp \nmid h)(x, \xi ; h):=\left.\left(e^{i \frac{h}{2}\left(\partial_{x} \cdot \partial_{\zeta}-\partial_{y} \cdot \partial_{\xi}\right)} a(x, \xi ; h) b(y, \zeta ; h)\right)\right|_{y=x, \zeta=\xi}
$$

and $c \in S\left(\mathfrak{m}_{1} \mathfrak{m}_{2}\right)$. See also Proposition A.2.1 for explicit bounds on $c$.
For $m \in \mathbb{Z}$, define the standard Hilbert spaces

$$
\begin{equation*}
\mathcal{H}^{m}:=\left\{\Psi \in \mathcal{S}^{\prime}\left(\mathbb{R}^{2}\right) \otimes \mathbb{C}^{n} \quad\left|\quad \partial_{\alpha} \Psi \in \mathcal{H} \quad \forall\right| \alpha \mid \leq m\right\} \tag{A.1.3}
\end{equation*}
$$

Following [20,55,61], we define the Hörmander class $S_{1,0}^{m}$ to be the space of symbols $a(x, \xi)$ that satisfy

$$
\begin{equation*}
\left|\left(\partial_{\xi}^{\alpha} \partial_{x}^{\beta} a\right)(x, \xi)\right| \leq C_{\alpha, \beta}\langle\xi\rangle^{m-|\alpha|} ; \quad \alpha, \beta \in \mathbb{N}^{d} \tag{A.1.4}
\end{equation*}
$$

We define $E S_{1,0}^{m}$ to be the space of Hermitian-valued symbols $a(x, \xi)$ that satisfy (A.1.4) and $\left|a_{\min }(x, \xi)\right| \geq c\langle\xi\rangle^{m}-1$ for some $c>0$, where $a_{\text {min }}$ is the smallest-magnitude eigenvalue of $a$.

If $\mathcal{A}$ is a symbol class (e.g. $S(\mathfrak{m}), S_{1,0}^{m}, E S_{1,0}^{m}$ ), we write $A \in \operatorname{Op}(\mathcal{A})$ to mean that $A=\operatorname{Op}(a)$ for some $a \in \mathcal{A}$. In the case $\mathcal{A}=S(\mathfrak{m})$, the notation $A \in \mathrm{Op}_{h}(S(\mathfrak{m}))$ means that $A=\mathrm{Op}_{h}(a)$ for some $a \in S(\mathfrak{m})$.

Helffer-Sjöstrand formula. See [34, Chapter 8]. Given $f \in \mathcal{C}_{c}^{\infty}(\mathbb{R})$, there exists an almost analytic extension $\tilde{f} \in \mathcal{C}_{c}^{\infty}(\mathbb{C})$ that satisfies

$$
\begin{equation*}
|\bar{\partial} \tilde{f}| \leq C_{N}|\omega|^{N}, \quad N \in\{0,1,2, \ldots\} ; \quad \tilde{f}(\lambda)=f(\lambda), \quad \lambda \in \mathbb{R} \tag{A.1.5}
\end{equation*}
$$

Here, $z=: \lambda+i \omega$ and $\bar{\partial}:=\frac{1}{2} \partial_{\lambda}+\frac{i}{2} \partial_{\omega}$. We now recall [34, Theorem 8.1]. If $H$ is a self-adjoint operator on a Hilbert space, then

$$
\begin{equation*}
f(H)=-\frac{1}{\pi} \int \bar{\partial} \tilde{f}(z)(z-H)^{-1} d^{2} z \tag{A.1.6}
\end{equation*}
$$

where $d^{2} z$ is the Lebesgue measure on $\mathbb{C}$. (A.1.6) is known as the Helffer-Sjöstrand formula, and we use it repeatedly in this thesis.

Trace-class operators. See [34, Chapter 9]. Suppose $\mathfrak{m} \in L^{1}\left(\mathbb{R}^{2 d}\right)$, and $\left|\partial^{\alpha} a(x, \xi ; h)\right| \leq$ $C_{\alpha} \mathfrak{m}(x, \xi)$ for all $\alpha \in \mathbb{N}^{2 d}$ and $h \in(0,1]$ (meaning that $a \in S(\mathfrak{m})$ ). Then $\operatorname{Op}_{h}(a)$ is trace-class with $\left\|\operatorname{Op}_{h}(a)\right\|_{1} \leq C \max _{|\alpha| \leq 2 d+1} C_{\alpha}\|\mathfrak{m}\|_{L^{1}}$ and

$$
\begin{equation*}
\operatorname{Tr} \mathrm{Op}_{h}(a)=\frac{1}{(2 \pi h)^{d}} \int_{\mathbb{R}^{2 d}} \operatorname{tr} a(x, \xi ; h) d x d \xi \tag{A.1.7}
\end{equation*}
$$

where $C$ depends only on $d$ and $t r$ is the standard matrix trace. To obtain the above equality, we use [34, Theorem 9.4] to write

$$
\begin{aligned}
\operatorname{Tr} \operatorname{Op}_{h}(a(x, \xi ; h)) & =\operatorname{Tr} \operatorname{Op}_{1}(a(x, h \xi ; h)) \\
& =\frac{1}{(2 \pi)^{d}} \int_{\mathbb{R}^{2 d}} \operatorname{tr} a(x, h \xi ; h) d x d \xi=\frac{1}{(2 \pi h)^{d}} \int_{\mathbb{R}^{2 d}} \operatorname{tr} a(x, \xi ; h) d x d \xi
\end{aligned}
$$

The results from [20,34] that were used in this section are stated for scalar symbols. They extend to the matrix-valued case; see [19].

## A. 2 Composition calculus

For $\mathfrak{m}: \mathbb{R}^{2 d} \rightarrow[0, \infty)$ an order function, $u \in S(\mathfrak{m})$ and $N \in \mathbb{N}$, define

$$
\tilde{C}_{N}(u, \mathfrak{m}):=\sum_{|\alpha| \leq N} \inf \left\{C>0:\left|\partial^{\alpha} u\right| \leq C \mathfrak{m}\right\}
$$

For $u_{1}, u_{2} \in \mathcal{C}^{\infty}\left(\mathbb{R}^{2 d} ; \mathbb{C}^{n \times n}\right)$ and $N \in \mathbb{N}$, define

$$
\begin{aligned}
T_{N}\left(u_{1}, u_{2}\right) & =\left(T_{N}\left(u_{1}, u_{2}\right)\right)(x, \xi) \\
& :=\left.\sum_{j=0}^{N-1}\left(\frac{\left(i h\left(D_{\xi} \cdot D_{y}-D_{x} \cdot D_{\eta}\right) / 2\right)^{j}}{j!} u_{1}(x, \xi) u_{2}(y, \eta)\right)\right|_{y=x, \eta=\xi}
\end{aligned}
$$

Proposition A.2.1. Let $\mathfrak{m}_{1}, \mathfrak{m}_{2}: \mathbb{R}^{2 d} \rightarrow[0, \infty)$ be order functions. Then there exists $s \in \mathbb{N}$
such that for every $N \in \mathbb{N}$ and $\alpha \in \mathbb{N}^{2 d}$,

$$
\left|\partial^{\alpha}\left(u_{1} \nVdash_{h} u_{2}-T_{N}\right)\right| \leq C_{\alpha, N} \sum_{j=0}^{|\alpha|} \tilde{C}_{2 N+s+j}\left(u_{1}, \mathfrak{m}_{1}\right) \tilde{C}_{2 N+s+|\alpha|-j}\left(u_{2}, \mathfrak{m}_{2}\right) h^{N_{\mathfrak{m}_{1}} \mathfrak{m}_{2}}
$$

uniformly in $u_{1} \in S\left(\mathfrak{m}_{1}\right)$, $u_{2} \in S\left(\mathfrak{m}_{2}\right)$ and $h \in(0,1]$.

Proof. We follow arguments from [34, Chapter 7] and [89, Chapters 3 \& 4]. Since

$$
u_{1} \sharp h u_{2}=\left.\left(e^{i \frac{h}{2}\left(D_{\xi} \cdot D_{y}-D_{x} \cdot D_{\eta}\right)} u_{1}(x, \xi) u_{2}(y, \eta)\right)\right|_{y=x, \eta=\xi}=:\left.\left(e^{i h A(D)} u_{1}(x, \xi) u_{2}(y, \eta)\right)\right|_{y=x, \eta=\xi}
$$

and $D^{\alpha}$ commutes with $e^{i h A(D)}$ (for any $\alpha \in \mathbb{N}^{d}$ ), it suffices to show that there exists $s \in \mathbb{N}$ such that for any $N \in \mathbb{N}$ and order function $\mathfrak{m}: \mathbb{R}^{4 d} \rightarrow[0, \infty)$,

$$
\left|e^{i h A(D)} u(X)-\sum_{j \leq N-1} \frac{(i h A(D))^{j}}{j!} u(X)\right| \leq C_{N} \tilde{C}_{2 N+s}(u, \mathfrak{m}) h^{N} \mathfrak{m}(X)
$$

uniformly in $u \in S(\mathfrak{m})$ and $h \in(0,1]$. Here, we use the shorthand $X:=(x, y, \xi, \zeta) \in \mathbb{R}^{4 d}$.
Since $A$ is non-degenerate, we can let $A^{-1}(X)=\frac{1}{2}\left\langle Q^{-1} X, X\right\rangle$ be the dual quadratic form on $\mathbb{R}^{4 d}$. Then, as stated in [34, bottom of page 80], $e^{i h A(D)} u=K_{h} * u$, where

$$
K_{h}(x)=C h^{-2 d} e^{-i A^{-1}(X) / h}
$$

We then write

$$
e^{i h A(D)} u=\left(\chi K_{h}\right) * u+\left((1-\chi) K_{h}\right) * u=: t_{1}+t_{2}
$$

where $\chi \in \mathcal{C}_{c}^{\infty}(B(0,2))$ is equal to 1 in $B(0,1)$. Set $p:=2 d+1$. By [89, Theorems $3.14 \&$
4.16 and their proofs], we obtain that

$$
\begin{aligned}
\left|t_{1}(X)-\sum_{j \leq N-1} \frac{(i h A(D))^{j}}{j!} u(X)\right| & \leq C_{N} h^{N} \sum_{|\alpha| \leq 2 N+p} \sup _{B(X, 2)}\left|\partial^{\alpha} u\right| \\
& \leq C_{N} \tilde{C}_{2 N+p}(u, \mathfrak{m}) h^{N} \mathfrak{m}(X)
\end{aligned}
$$

For the second term, we directly apply [34, equation (7.19)] to obtain that for every $k \in \mathbb{N}$,

$$
\left|t_{2}(X)\right| \leq C_{k} h^{k} \sum_{|\alpha| \leq k+p}\left\|\langle X-\cdot\rangle^{-k-2 d} \partial^{\alpha} u(\cdot)\right\|_{L^{1}} \leq C_{k} \tilde{C}_{k+p}(u, \mathfrak{m}) h^{k}\left\|\langle X-\cdot\rangle^{-k-2 d} \mathfrak{m}(\cdot)\right\|_{L^{1}}
$$

Using that $\mathfrak{m}(Y) \leq C\langle X-Y\rangle^{N_{0}} \mathfrak{m}(X)$ for some $N_{0}>0$, it follows that

$$
\left|t_{2}(X)\right| \leq C_{k} \tilde{C}_{k+p}(u, \mathfrak{m}) h^{k} \mathfrak{m}(X)\left\|\langle\cdot\rangle^{N_{0}-k-2 d}\right\|_{L^{1}}
$$

Thus for all $k>N_{0}+p$ (so that the above integral is finite),

$$
\left|t_{2}(X)\right| \leq C_{k} \tilde{C}_{k+p}(u, m) h^{k} \mathfrak{m}(X)
$$

This completes the result.
Proposition A.2.2. There exists $N \in \mathbb{N}$ such that $\left\|\operatorname{Op}_{h}(u)\right\| \leq \tilde{C}_{N}(u, 1)$ uniformly in $u \in S(1)$ and $h \in(0,1]$.

Proof. See [34, Theorem 7.11 and its proof].

## A. 3 Elliptic operators and resolvent estimates

Proposition A.3.1. Let $\sigma \in E S_{1,0}^{m}$. Then for all $h \in(0,1]$, the operator $H_{h}:=\operatorname{Op}_{h}(\sigma)$ is self-adjoint with domain of definition $\mathcal{H}^{m}$. This means we can define $\operatorname{Op}_{h}\left(r_{z, h}\right):=(z-$ $\left.H_{h}\right)^{-1}$ whenever $\Im z \neq 0$. Let $Z \subset \mathbb{C}$ be bounded such that $\Im z \neq 0$ for all $z \in Z$. Then there
exists $s \in \mathbb{N}$ such that for any $\alpha \in \mathbb{N}^{2 d}$,

$$
\left|\partial^{\alpha} r_{z, h}(x, \xi)\right| \leq C_{\alpha}|\Im z|^{-s-|\alpha|}\langle\xi\rangle^{-m}
$$

uniformly in $z \in Z$ and $h \in(0,1]$.

Proof. By [20, Corollary 2 and the paragraph following Theorem 3], we know that $H_{h}$ is selfadjoint with domain of definition $\mathcal{H}^{m}$. Applying also [34, the paragraphs between equation (8.11) and Proposition 8.5], it follows that $\left(i-H_{h}\right)^{-1} \in \mathrm{Op}_{h}\left(S\left(\langle\xi, \zeta\rangle^{-m}\right)\right)$ is a bijection of $L^{2}\left(\mathbb{R}^{d}\right) \otimes \mathbb{C}^{n}$ onto $\mathcal{H}^{m}$. For $\Im z \neq 0$, we have that $A_{z, h}:=1-(i-z)\left(i-H_{h}\right)^{-1}$ is a bijection of $L^{2}\left(\mathbb{R}^{d}\right) \otimes \mathbb{C}^{n}$ onto itself, with $\left\|A_{z, h}^{-1}\right\| \leq C|\Im z|^{-1}$ uniformly in $z \in Z$. Applying [34, Proposition 8.4] to $\mathrm{Op}_{h}\left(b_{z, h}\right):=A_{z, h}^{-1}$, we obtain that

$$
\left|\partial^{\alpha} b_{z, h}(x, \xi)\right| \leq C_{\alpha}|\Im z|^{-2 d-2-|\alpha|}
$$

uniformly in $z \in Z$ and $h \in(0,1]$. The result then follows from Proposition A.2.1 (with $N=0)$ and the fact that $\left(z-H_{h}\right)^{-1}=\left(i-H_{h}\right)^{-1} A_{z, h}^{-1}$.

Below, let $\mathbb{M}_{n}$ denote the space of Hermitian $n \times n$ matrices.
Proposition A.3.2. Let $\sigma \in E S_{1,0}^{m}$ and $\mathcal{W} \subset \mathcal{C}^{\infty}\left(\mathbb{R}^{2 d} ; \mathbb{M}_{n}\right)$. Suppose that for any $(\alpha, \beta) \in$ $\mathbb{N}^{2 d}$,

$$
\begin{equation*}
\left|\partial_{x}^{\alpha} \partial_{\xi}^{\beta} w(x, \xi)\right| \leq C_{\alpha}\langle\xi\rangle^{m-|\beta|} \tag{A.3.1}
\end{equation*}
$$

uniformly in $w \in \mathcal{W}$ (meaning that $\mathcal{W} \subset S_{1,0}^{m}$ ). For $\mu \in[0,1]$ and $w \in \mathcal{W}$, define $H_{\mu, w}:=$ $\operatorname{Op}(\sigma+\mu w)$. Then there exists $\mu_{0} \in(0,1]$ such that the following conditions hold:

1. If $w \in \mathcal{W}$ and $\mu \in\left[0, \mu_{0}\right]$, then $H_{\mu, w}$ is self-adjoint with domain of definition $\mathcal{H}^{m}$.
2. For $\Im z \neq 0$, define $\operatorname{Op}\left(r_{z, \mu, w}\right):=\left(z-H_{\mu, w}\right)^{-1}$. Let $Z \subset \mathbb{C}$ be bounded such that $\Im z \neq 0$ for all $z \in Z$. Then there exists $s \in \mathbb{N}$ such that for any $\alpha \in \mathbb{N}^{2 d}$,

$$
\left|\partial^{\alpha} r_{z, \mu, w}(x, \xi)\right| \leq C_{\alpha}|\Im z|^{-s-|\alpha|}\langle\xi\rangle^{-m}
$$

uniformly in $z \in Z, w \in \mathcal{W}$ and $\mu \in\left[0, \mu_{0}\right]$.
Proof. Define $\sigma^{(\mu, w)}:=\sigma+\mu w$.

1. The uniform bounds (A.3.1) imply that whenever $\mu>0$ is sufficiently small, $\sigma^{(\mu, w)} \in$ $E S_{1,0}^{m}$ for all $w \in \mathcal{W}$. Hence in this case, $H_{\mu, w}$ is self-adjoint with domain of definition $\mathcal{H}^{m}$.
2. We write $\left(i-H_{\mu, w}\right)^{-1}=\left(1+\left(i-H_{\mu, w}\right)^{-1} \mu W\right)(i-H)^{-1}$, which implies

$$
\left(i-H_{\mu, w}\right)^{-1}\left(1-\mu W(i-H)^{-1}\right)=(i-H)^{-1}
$$

where $W:=\mathrm{Op}_{h}(w)$ and $H:=\mathrm{Op}(\sigma)$. $\mathrm{By}(\mathrm{A} .3 .1)$ and Proposition A.2.2, we conclude that whenever $\mu>0$ is sufficiently small, $A_{\mu, w}:=1-\mu W(i-H)^{-1}$ is a bijection of $L^{2}\left(\mathbb{R}^{d}\right) \otimes \mathbb{C}^{n}$ onto itself, with $A_{\mu, w}^{-1} \in \operatorname{Op}(S(1))$ uniformly in $\mu$ and $w$. Since $(i-H)^{-1} \in \operatorname{Op}\left(S\left(\langle\xi\rangle^{-m}\right)\right)$ by Proposition A.3.1, it follows that $\left(i-H_{\mu, w}\right)^{-1} \in$ $\operatorname{Op}\left(S\left(\langle\xi\rangle^{-m}\right)\right)$ uniformly in $\mu$ and $w$. The bounds on $r_{z, \mu, w}$ then follow from the same argument that was used in the proof of Proposition A.3.1 (after it was shown that $\left(i-H_{h}\right)^{-1} \in \mathrm{Op}_{h}\left(S\left(\langle\xi, \zeta\rangle^{-m}\right)\right)$ there $)$.

Proposition A.3.3. Let $\sigma \in E S_{1,0}^{m}$ and define $H_{h}:=\operatorname{Op}_{h}(\sigma)$ for $h \in(0,1]$. Let $\phi \in$ $\mathcal{C}_{c}^{\infty}\left(E_{1}, E_{2}\right)$ and define $\operatorname{Op}_{h}\left(\nu_{h}\right):=\phi\left(H_{h}\right)$. Let $\mathfrak{m}: \mathbb{R}^{2 d} \rightarrow[0, \infty)$ be any order function such that $\nu_{h} \in S\left(\mathfrak{m}^{-\infty}\right)$ and all eigenvalues of $\sigma$ lie outside the interval $\left(E_{1}, E_{2}\right)$ whenever $\mathfrak{m}$ is
sufficiently large. For $z \in \mathbb{C}$, define $\sigma_{z}:=z-\sigma$. For $N \in \mathbb{N}_{+}$, define $q_{z, h, N}$ recursively by $q_{z, h, 1}:=\sigma_{z}^{-1}$ and

$$
q_{z, h, N}=\sigma_{z}^{-1}\left(1-\left.\sum_{j=1}^{N-1}\left(\frac{\left(i h\left(D_{\xi} \cdot D_{y}-D_{x} \cdot D_{\eta}\right) / 2\right)^{j}}{j!} \sigma_{z}(x, \xi) q_{z, h, N-j}(y, \eta)\right)\right|_{y=x, \eta=\xi}\right)
$$

for all $N \geq 2$. Then for all $N \in \mathbb{N}$,

$$
\begin{equation*}
\nu_{h}+\frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\phi}(z) q_{z, h, N} \mathrm{~d}^{2} z \in S^{-N+1 / 2}\left(\mathfrak{m}^{-\infty}\right) \tag{A.3.2}
\end{equation*}
$$

Proof. Note that $\operatorname{supp}(\tilde{\phi}) \subset\left(E_{1}, E_{2}\right) \times(-M, M)=: Z$ for some $0<M<\infty$, hence the integral over $\mathbb{C}$ in (A.3.2) can be restricted to an integral over $Z$. By assumption, for any $N \in \mathbb{N}, z \mapsto q_{z, h, N}$ is analytic in $z \in Z$ so long as $\mathfrak{m}(x, \xi)$ is sufficiently large (independent of $h$ ). Integrating by parts, we conclude that $\int_{Z} \bar{\partial} \tilde{\phi}(z) q_{z, h, N} \mathrm{~d}^{2} z$ vanishes whenever $\mathfrak{m}(x, \xi)$ is sufficiently large. This implies

$$
\nu_{h}+\frac{1}{\pi} \int_{Z} \tilde{\partial} \tilde{\phi}(z) q_{z, h, N} \mathrm{~d}^{2} z \in S\left(\mathfrak{m}^{-\infty}\right)
$$

The result will follow from an $h$-dependent bound of the above left-hand side and interpolation. Namely, it suffices to show that

$$
\begin{equation*}
\nu_{h}+\frac{1}{\pi} \int_{Z} \bar{\partial} \tilde{\phi}(z) q_{z, h, N} \mathrm{~d}^{2} z \in S^{-N+1 / 4}(1) \tag{A.3.3}
\end{equation*}
$$

By the Helffer-Sjöstrand formula, the above left-hand side equals

$$
-\frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\phi}(z) r_{z, h} \mathrm{~d}^{2} z+\frac{1}{\pi} \int_{Z} \bar{\partial} \tilde{\phi}(z) q_{z, h, N} \mathrm{~d}^{2} z=\frac{1}{\pi} \int_{Z} \bar{\partial} \tilde{\phi}(z)\left(q_{z, h, N}-r_{z, h}\right) \mathrm{d}^{2} z
$$

where $\mathrm{Op}_{h}\left(r_{z, h}\right):=\left(z-H_{h}\right)^{-1}$. Therefore, given the rapid decay of $\bar{\partial} \tilde{\phi}$ near the real axis,
(A.3.3) holds if there exists $s \in \mathbb{N}$ such that for all $\alpha \in \mathbb{N}^{d}$ and $N \in \mathbb{N}$,

$$
\begin{equation*}
\left|\partial^{\alpha}\left(q_{z, h, N}-r_{z, h}\right)\right| \leq C_{\alpha, N}|\Im z|^{-(2 N+2 s+|\alpha|+1)} h^{N} \tag{A.3.4}
\end{equation*}
$$

uniformly in $z \in Z$ and $h \in(0,1]$. We will now prove (A.3.4) by induction.
Using that $\sigma_{z} \sharp r_{z, h}=1$ for all $h \in(0,1]$, it follows from Proposition A.2.1 that

$$
\left|\partial^{\alpha}\left(1-\sigma_{z} r_{z, h}\right)\right| \leq C_{\alpha, 1} \sum_{j=0}^{|\alpha|} \tilde{C}_{2+s+j}\left(\sigma_{z},\langle\xi\rangle^{m}\right) \tilde{C}_{2+s+|\alpha|-j}\left(r_{z, h},\langle\xi\rangle^{-m}\right) h
$$

hence

$$
\begin{aligned}
&\left|\partial^{\alpha}\left(\sigma_{z}^{-1}-r_{z, h}\right)\right|=\left|\partial^{\alpha}\left(\sigma_{z}^{-1}\left(1-\sigma_{z} r_{z, h}\right)\right)\right| \\
& \leq C_{\alpha, 1} \sum_{j=0}^{|\alpha|} \sum_{k=0}^{|\alpha|-j} \tilde{C}_{j}\left(\sigma_{z}^{-1},\langle\xi\rangle^{-m}\right) \tilde{C}_{2+s+k}\left(\sigma_{z},\langle\xi\rangle^{m}\right) \tilde{C}_{2+s+|\alpha|-j-k}\left(r_{z, h},\langle\xi\rangle^{-m}\right) h
\end{aligned}
$$

uniformly in $z \in Z$ and $h \in(0,1]$. Observe that for $k \in \mathbb{N}$,

$$
\tilde{C}_{k}\left(\sigma_{z}^{-1},\langle\xi\rangle^{-m}\right) \leq C|\Im z|^{-1-k}, \quad \tilde{C}_{k}\left(\sigma_{z},\langle\xi\rangle^{m}\right) \leq C
$$

uniformly in $z \in Z$. Since $\tilde{C}_{k}\left(r_{z, h},\langle\xi\rangle^{-m}\right) \leq C|\Im z|^{-s-k}$ by Proposition A.3.1, we have verified (A.3.4) when $N=1$.

Now, fix $N \in \mathbb{N}$ and suppose that for all $k \in\{1, \ldots, N\}$,

$$
\left|\partial^{\alpha}\left(q_{z, h, k}-r_{z, h}\right)\right| \leq C_{\alpha, k}|\Im z|^{-(2 k+2 s+|\alpha|+1)} h^{k}
$$

uniformly in $z \in Z$ and $h \in(0,1]$. Then

$$
\begin{aligned}
& q_{z, h, N+1}-r_{z, h} \\
&= \sigma_{z}^{-1}\left(1-\sigma_{z} r_{z, h}-\left.\sum_{j=1}^{N}\left(\frac{\left(i h\left(D_{\xi} \cdot D_{y}-D_{x} \cdot D_{\eta}\right) / 2\right)^{j}}{j!} \sigma_{z}(x, \xi) q_{z, h, N+1-j}(y, \eta)\right)\right|_{y=x, \eta=\xi}\right) \\
&= \sigma_{z}^{-1}\left(1-\left.\sum_{j=0}^{N}\left(\frac{\left(i h\left(D_{\xi} \cdot D_{y}-D_{x} \cdot D_{\eta}\right) / 2\right)^{j}}{j!} \sigma_{z}(x, \xi) r_{z, h}(y, \eta)\right)\right|_{y=x, \eta=\xi}\right) \\
&+\left.\sigma_{z}^{-1} \sum_{j=1}^{N}\left(\frac{\left(i h\left(D_{\xi} \cdot D_{y}-D_{x} \cdot D_{\eta}\right) / 2\right)^{j}}{j!} \sigma_{z}(x, \xi)\left(r_{z, h}(y, \eta)-q_{z, h, N+1-j}(y, \eta)\right)\right)\right|_{y=x, \eta=\xi} \\
&= t_{1}+t_{2}
\end{aligned}
$$

Proposition A.2.1 implies that

$$
\begin{aligned}
&\left|\partial^{\alpha} t_{1}\right| \leq C_{\alpha, N+1} \sum_{j=0}^{|\alpha|} \sum_{k=0}^{|\alpha|-j} \tilde{C}_{j}\left(\sigma_{z}^{-1},\right. \\
&\left.,\langle\xi\rangle^{-m}\right) \tilde{C}_{2(N+1)+s+k}\left(\sigma_{z},\langle\xi\rangle^{m}\right) \\
& \times \tilde{C}_{2(N+1)+s+|\alpha|-j-k}\left(r_{z, h},\langle\xi\rangle^{-m}\right) h^{N+1} \\
& \leq C_{\alpha, N+1}|\Im z|^{-(2(N+1)+2 s+|\alpha|+1)} h^{N+1}
\end{aligned}
$$

Define $t_{2}=: \sigma_{z}^{-1} \sum_{j=1}^{N} t_{2, j}$. By our inductive hypothesis,

$$
\begin{aligned}
\left|\partial^{\alpha} t_{2, j}\right| & \leq C_{\alpha, j} h^{j} \sum_{i=0}^{|\alpha|+j} \tilde{C}_{i}\left(\sigma_{z},\langle\xi\rangle^{m}\right) \tilde{C}_{|\alpha|+j-i}\left(r_{z, h}-q_{z, h, N+1-j}, 1\right)\langle\xi\rangle^{m} \\
& \leq C_{\alpha, j}|\Im z|^{-(2(N+1-j)+2 s+|\alpha|+j+1)} h^{N+1}\langle\xi\rangle^{m} \\
& =C_{\alpha, j}|\Im z|^{-(2(N+1)+2 s+|\alpha|+1-j)} h^{N+1}\langle\xi\rangle^{m} \\
& \leq C_{\alpha, j}|\Im z|^{-(2(N+1)+2 s+|\alpha|)} h^{N+1}\langle\xi\rangle^{m}
\end{aligned}
$$

It follows that

$$
\left|\partial^{\alpha}\left(\sigma_{z}^{-1} t_{2, j}\right)\right| \leq C_{\alpha, j}|\Im z|^{-(2(N+1)+2 s+|\alpha|+1)} h^{N+1}
$$

for all $j \in\{1, \ldots, N\}$. We have thus verified (A.3.4) for all $N \in \mathbb{N}$, and the proof is complete.

For completeness, we now write an expansion for $\nu_{h}$ that does not involve an integral.

Proposition A.3.4. Take $\sigma, \nu_{h}, \varphi$ and $m$ as above. Let $\left\{\mu_{j}\right\}_{j=1}^{d}$ denote the eigenvalues of $\sigma$ and $\left\{\Pi_{j}\right\}_{j=1}^{d}$ the corresponding eigenprojections, so that $\sigma=\sum_{j=1}^{d} \mu_{j} \Pi_{j}$. Then

$$
\begin{array}{r}
\nu_{h}=\varphi^{\prime}(\sigma)+i h \sum_{\ell_{1}, \ell_{2}, \ell_{3}=1}^{d} \Psi\left(\mu_{\ell_{1}}, \mu_{\ell_{2}}, \mu_{\ell_{3}}\right) \Pi_{\ell_{1}} \sum_{k=1}^{n}\left(\partial_{x_{k}} \sigma \Pi_{\ell_{2}} \partial_{\xi_{k}} \sigma-\partial_{\xi_{k}} \sigma \Pi_{\ell_{2}} \partial_{x_{k}} \sigma\right) \Pi_{\ell_{3}} \\
+O\left(h^{3 / 2}\right)
\end{array}
$$

in $S\left(m^{-\infty}\right)$, where $\Psi \in \mathcal{C}^{\infty}\left(\mathbb{R}^{3}\right)$ is defined by

$$
\Psi(x, y, z)=\frac{1}{(x-y)(x-z)} \varphi^{\prime}(x)+\frac{1}{(y-x)(y-z)} \varphi^{\prime}(y)+\frac{1}{(z-x)(z-y)} \varphi^{\prime}(z) .
$$

Proof. This result is a direct application of Proposition A.3.3. Namely, $q_{z, h, 1}=\sigma_{z}^{-1}$, so the $O(1)$ in the expansion for $\nu_{h}$ is

$$
-\frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\varphi}^{\prime}(z) \sigma_{z}^{-1} \mathrm{~d}^{2} z=\varphi^{\prime}(\sigma)
$$

The $O(h)$ term is

$$
-\frac{i}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\varphi}^{\prime}(z) \sigma_{z}^{-1}\left\{\sigma_{z}, \sigma_{z}^{-1}\right\} \mathrm{d}^{2} z
$$

Using the eigendecomposition $\sigma=\sum_{j=1}^{d} \mu_{j} \Pi_{j}$, a simple calculation reveals that

$$
\begin{aligned}
& \sigma_{z}^{-1}\left\{\sigma_{z}, \sigma_{z}^{-1}\right\}= \\
& \quad \sum_{\ell_{1}, \ell_{2}, \ell_{3}=1}^{d}\left(z-\mu_{\ell_{1}}\right)^{-1}\left(z-\mu_{\ell_{2}}\right)^{-1}\left(z-\mu_{\ell_{3}}\right)^{-1} \Pi_{\ell_{1}} \sum_{k=1}^{n}\left(\partial_{x_{k}} \sigma \Pi_{\ell_{2}} \partial_{\xi_{k}} \sigma-\partial_{\xi_{k}} \sigma \Pi_{\ell_{2}} \partial_{x_{k}} \sigma\right) \Pi_{\ell_{3}} .
\end{aligned}
$$

The result then follows from the identity

$$
-\frac{1}{\pi} \int_{\mathbb{C}} \bar{\partial} \tilde{\varphi}^{\prime}(z)\left(z-\mu_{\ell_{1}}\right)^{-1}\left(z-\mu_{\ell_{2}}\right)^{-1}\left(z-\mu_{\ell_{3}}\right)^{-1} \mathrm{~d}^{2} z=\Psi\left(\mu_{\ell_{1}}, \mu_{\ell_{2}}, \mu_{\ell_{3}}\right)
$$

## REFERENCES

[1] M. Andelković, L. Covaci, and F. Peeters, Dc conductivity of twisted bilayer graphene: Angle-dependent transport properties and effects of disorder, Physical Review Materials, 2 (2018), p. 034004.
[2] X. Antoine, Advances in the on-surface radiation condition method: Theory, numerics and applications, Computational Methods for Acoustics Problems, (2008), pp. 169-194.
[3] X. Antoine and M. Darbas, Alternative integral equations for the iterative solution of acoustic scattering problems, The Quarterly Journal of Mechanics and Applied Mathematics, 58 (2005), pp. 107-128.
[4] X. Antoine, M. Darbas, and Y. Y. Lu, An improved surface radiation condition for high-frequency acoustic scattering problems, Computer Methods in Applied Mechanics and Engineering, 195 (2006), pp. 4060-4074.
[5] J. E. Avron, R. Seiler, and B. Simon, Quantum hall effect and the relative index for projections, Physical review letters, 65 (1990), p. 2185.
[6] __, Charge deficiency, charge transport and comparison of dimensions, Communications in mathematical physics, 159 (1994), pp. 399-422.
[7] G. BAL, Continuous bulk and interface description of topological insulators, Journal of Mathematical Physics, 60 (2019), p. 081506.
[8] _, Topological protection of perturbed edge states, Communications in Mathematical Sciences, 17 (2019), pp. 193-225.
[9] _—, Semiclassical propagation along curved domain walls, arXiv preprint arXiv:2206.09439, (2022).
[10] —_, Topological invariants for interface modes, Communications in Partial Differential Equations, 47(8) (2022), pp. 1636-1679.
[11] __, Topological charge conservation for continuous insulators, To appear in J. Math. Phys. and arXiv:2106.08480, (2023).
[12] G. Bal, S. Becker, A. Drouot, C. F. Kammerer, J. Lu, and A. Watson, Edge state dynamics along curved interfaces, arXiv preprint arXiv:2106.00729, (2021).
[13] G. Bal, P. Cazeaux, D. Massatt, and S. Quinn, Mathematical models of topologically protected transport in twisted bilayer graphene, arXiv preprint arXiv:2206.05580, (2022).
[14] G. Bal, J. G. Hoskins, and Z. Wang, Asymmetric transport computations in Dirac models of topological insulators, arXiv preprint arXiv:2206.08847, (2022).
[15] G. Bal and D. Massatt, Multiscale invariants of floquet topological insulators, arXiv preprint arXiv:2101.06330, (2021).
[16] J. Bellissard, A. van Elst, and H. Schulz-Baldes, The noncommutative geometry of the quantum hall effect, Journal of Mathematical Physics, 35 (1994), pp. 53735451.
[17] B. A. Bernevig and T. L. Hughes, Topological insulators and topological superconductors, Princeton university press, 2013.
[18] R. Bistritzer and A. H. MacDonald, Moiré bands in twisted double-layer graphene, Proceedings of the National Academy of Sciences - PNAS, 108 (2011), pp. 12233-12237.
[19] J. Bolte and R. Glaser, A semiclassical egorov theorem and quantum ergodicity for matrix valued operators, Communications in mathematical physics, 247 (2004), pp. 391419.
[20] J.-M. Bony, On the characterization of pseudodifferential operators (old and new), in Studies in Phase Space Analysis with Applications to PDEs, Springer, 2013, pp. 21-34.
[21] J. Bremer, Z. Gimbutas, and V. Rokhlin, A nonlinear optimization procedure for generalized Gaussian quadratures, SIAM Journal on Scientific Computing, 32 (2010), pp. 1761-1788.
[22] J. Bremer, V. Rokhlin, and I. Sammis, Universal quadratures for boundary integral equations on two-dimensional domains with corners, Journal of Computational Physics, 229 (2010), pp. 8259-8280.
[23] Y. Cao, V. Fatemi, S. Fang, K. Watanabe, T. Taniguchi, E. Kaxiras, and P. Jarillo-Herrero, Unconventional superconductivity in magic-angle graphene superlattices, Nature, 556 (2018), pp. 43-50.
[24] S. Carr, D. Massatt, S. B. Torrisi, P. Cazeaux, M. Luskin, and E. Kaxiras, Relaxation and domain formation in incommensurate two-dimensional heterostructures, Phys. Rev. B, 98 (2018), p. 224102.
[25] S. Chaillat, M. Darbas, and F. Le LouËr, A new analytic preconditioner for the iterative solution of Dirichlet exterior scattering problems in 3D elasticity, 2014.
[26] _—, Approximate local Dirichlet-to-Neumann map for three-dimensional timeharmonic elastic waves, Computer Methods in Applied Mechanics and Engineering, 297 (2015), pp. 62-83.
[27] ChUNKIE, https://github.com/fastalgorithms/chunkie, GitHub Repository, (2023).
[28] J.-M. Combes and F. Germinet, Edge and impurity effects on quantization of hall currents, Communications in mathematical physics, 256 (2005), pp. 159-180.
[29] H. D. Cornean, M. Moscolari, and K. S. Sørensen, Bulk-edge correspondence for unbounded dirac-landau operators, arXiv preprint arXiv:2208.02218, (2022).
[30] H. L. Cycon, R. G. Froese, W. Kirsch, and B. Simon, Schrödinger operators: With application to quantum mechanics and global geometry, Springer, 2009.
[31] S. Dai, Y. Xiang, and D. J. Srolovitz, Twisted bilayer graphene: Moiré with a twist, Nano Letters, 16 (2016), pp. 5923-5927.
[32] M. Darbas, E. Darrigrand, and Y. Lafranche, Combining analytic preconditioner and fast multipole method for the 3-D Helmholtz equation, Journal of Computational Physics, 236 (2013), pp. 289-316.
[33] P. Delplace, J. Marston, and A. Venaille, Topological origin of equatorial waves, Science, 358 (2017), pp. 1075-1077.
[34] M. Dimassi, J. Sjostrand, et al., Spectral asymptotics in the semi-classical limit, London Mathematical Society Lecture Note Series (268), Cambridge university press, 1999.
[35] N. Dombrowski, F. Germinet, and G. Raikov, Quantization of edge currents along magnetic barriers and magnetic guides, Annales Henri Poincaré, 12 (2011), pp. 1169-1197.
[36] A. Drouot, Characterization of edge states in perturbed honeycomb structures, Pure and Applied Analysis, 1 (2019), pp. 385-445.
[37] __, Microlocal analysis of the bulk-edge correspondence, Communications in Mathematical Physics, 383 (2021), pp. 2069-2112.
[38] B. A. Dubrovin, A. T. Fomenko, and S. P. Novikov, Modern geometry-methods and applications: Part II: The geometry and topology of manifolds, vol. 104, Springer Science \& Business Media, 2012.
[39] S. Dyatlov and M. Zworski, Mathematical theory of scattering resonances, vol. 200, American Mathematical Soc., 2019.
[40] P. Elbau and G.-M. Graf, Equality of bulk and edge hall conductance revisited, Communications in mathematical physics, 229 (2002), pp. 415-432.
[41] A. Elgart, G. Graf, and J. Schenker, Equality of the bulk and edge hall conductances in a mobility gap, Communications in mathematical physics, 259 (2005), pp. 185-221.
[42] A. M. Essin and V. Gurarie, Bulk-boundary correspondence of topological insulators from their respective green's functions, Physical Review B, 84 (2011), p. 125132.
[43] P. EXnER, Leaky quantum graphs: a review, arXiv preprint arXiv:0710.5903, (2007).
[44] C. L. Fefferman, J. P. Lee-Thorp, and M. I. Weinstein, Edge states in honeycomb structures, Annals of PDE, 2 (2016), p. 12.
[45] FMM2D, https://github.com/flatironinstitute/fmm2d, GitHub Repository, (2023).
[46] M. Fruchart and D. Carpentier, An introduction to topological insulators, Comptes Rendus Physique, 14 (2013), pp. 779-815.
[47] T. Fukui, K. Shiozaki, T. Fujiwara, and S. Fujimoto, Bulk-edge correspondence for chern topological phases: A viewpoint from a generalized index theorem, Journal of the Physical Society of Japan, 81 (2012), p. 114602.
[48] G. M. Graf, H. Jud, and C. Tauber, Topology in shallow-water waves: a violation of bulk-edge correspondence, Communications in Mathematical Physics, 383 (2021), pp. 731-761.
[49] G. M. Graf and M. Porta, Bulk-edge correspondence for two-dimensional topological insulators, Communications in Mathematical Physics, 324 (2013), pp. 851-895.
[50] L. Greengard and V. Rokhlin, A fast algorithm for particle simulations, Journal of computational physics, 73 (1987), pp. 325-348.
[51] J. Helsing and A. Karlsson, An extended charge-current formulation of the electromagnetic transmission problem, SIAM Journal on Applied Mathematics, 80 (2020), pp. 951-976.
[52] J. Helsing and K.-M. Perfekt, The spectra of harmonic layer potential operators on domains with rotationally symmetric conical points, Journal de Mathématiques Pures et Appliquées, 118 (2018), pp. 235-287.
[53] J. Helsing and A. Rosén, Dirac integral equations for dielectric and plasmonic scattering, Integral Equations and Operator Theory, 93 (2021), pp. 1-41.
[54] M. Holzmann and G. Unger, Boundary integral formulations of eigenvalue problems for elliptic differential operators with singular interactions and their numerical approximation by boundary element methods, arXiv preprint arXiv:1907.04282, (2019).
[55] L. HÖRmander, The weyl calculus of pseudo-differential operators, Communications on Pure and Applied Mathematics, 32 (1979), pp. 359-443.
[56] __, The analysis of linear partial differential operators III: Pseudo-differential operators, Springer Science \& Business Media, 2007.
[57] N. Kalton, Trace-class operators and commutators, Journal of Functional Analysis, 86 (1989), pp. 41-74.
[58] T. Kato, Perturbation theory for linear operators, vol. 132, Springer Science \& Business Media, 2013.
[59] G. A. Kriegsmann, A. Taflove, and K. R. Umashankar, A new formulation of electromagnetic wave scattering using an on-surface radiation boundary condition approach, IEEE Transactions on Antennas and Propagation, 35 (1987), pp. 153-161.
[60] A. Kurganov and S. V. Tsynkov, On spectral accuracy of quadrature formulae based on piecewise polynomial interpolation, tech. rep., North Carolina State University. Center for Research in Scientific Computation, 2007.
[61] N. Lerner, Metrics on the phase space and non-selfadjoint pseudo-differential operators, vol. 3, Springer Science \& Business Media, 2011.
[62] T. A. Loring, $k$-theory and pseudospectra for topological insulators, Annals of Physics, 356 (2015), pp. 383-416.
[63] T. A. Loring and H. Schulz-Baldes, Spectral flow argument localizing an odd index pairing, Canadian Mathematical Bulletin, 62 (2019), pp. 373-381.
[64] _-, The spectral localizer for even index pairings, Journal of Noncommutative Geometry, 14 (2020), pp. 1-23.
[65] J. Michala, A. Pierson, T. A. Loring, and A. B. Watson, Wave-packet propagation in a finite topological insulator and the spectral localizer index, Involve, a Journal of Mathematics, 14 (2021), pp. 209-239.
[66] J. Phillips, Self-adjoint fredholm operators and spectral flow, Canadian Mathematical Bulletin, 39 (1996), pp. 460-467.
[67] E. Prodan, A computational non-commutative geometry program for disordered topological insulators, vol. 23 of SpringerBriefs in Mathematical Physics, Springer, 2017.
[68] E. Prodan and H. Schulz-Baldes, Bulk and boundary invariants for complex topological insulators: from $k$-theory to physics, Mathematical Physics Studies, (2016).
[69] H. RaEther, Surface plasmons on smooth surfaces, Surface plasmons on smooth and rough surfaces and on gratings, (1988), pp. 4-39.
[70] M. Reed and B. Simon, Methods of modern mathematical physics, I: Functional analysis, Elsevier, 1972.
[71] __, Methods of modern mathematical physics, IV: Analysis of operators, Academic press, 1978.
[72] P. Rickhaus, J. Wallbank, S. Slizovskiy, R. Pisoni, H. Overweg, Y. Lee, M. Eich, M.-H. Liu, K. Watanabe, T. Taniguchi, T. Ihn, and K. Ensslin, Transport through a network of topological channels in twisted bilayer graphene, Nano Letters, 18 (2018), pp. 6725-6730.
[73] V. Rokhlin, Rapid solution of integral equations of scattering theory in two dimensions, Journal of Computational physics, 86 (1990), pp. 414-439.
[74] P. San-Jose and E. Prada, Helical networks in twisted bilayer graphene under interlayer bias, Phys. Rev. B, 88 (2013), p. 121408.
[75] A. SARD, The measure of the critical values of differentiable maps, Bulletin of the American Mathematical Society, 48 (1942), pp. 883-890.
[76] M. Sato and Y. Ando, Topological superconductors: a review, Reports on Progress in Physics, 80 (2017), p. 076501.
[77] H. Schulz-Baldes, J. Kellendonk, and T. Richter, Simultaneous quantization of edge and bulk hall conductivity, Journal of Physics A: Mathematical and General, 33 (2000), p. L27.
[78] H. Schulz-Baldes and T. Stoiber, The spectral localizer for semifinite spectral triples, Proceedings of the American Mathematical Society, 149 (2021), pp. 121-134.
[79] A. Souslov, K. Dasbiswas, M. Fruchart, S. Vaikuntanathan, and V. Vitelli, Topological waves in fluids with odd viscosity, Physical review letters, 122 (2019), p. 128001.
[80] H. Tang, S. Carr, and E. Kaxiras, Geometric origins of topological insulation in twisted layered semiconductors, Physical Review B, 104 (2021).
[81] G. Teschl, Mathematical methods in quantum mechanics, Graduate Studies in Mathematics, 99 (2009), p. 106.
[82] B. Thaller, The dirac equation, Springer Science \& Business Media, 2013.
[83] K. Thicke, A. B. Watson, and J. Lu, Computing edge states without hard truncation, SIAM Journal on Scientific Computing, 43 (2021), pp. B323-B353.
[84] L. N. Trefethen, Spectral methods in MATLAB, SIAM, 2000.
[85] D. Tzarouchis and A. Sihvola, Light scattering by a dielectric sphere: Perspectives on the mie resonances, Applied Sciences, 8 (2018), p. 184.
[86] G. E. Volovik, The universe in a helium droplet, vol. 117, Oxford University Press on Demand, 2003.
[87] E. Witten, Three lectures on topological phases of matter, La Rivista del Nuovo Cimento, 39 (2016), pp. 313-370.
[88] D. YafaEv, On spectral properties of translationally invariant magnetic schrödinger operators, Annales Henri Poincaré, 9 (2008), pp. 181-207.
[89] M. Zworski, Semiclassical analysis, vol. 138, American Mathematical Soc., 2012.


[^0]:    1. These figures were generated using a standard finite difference approximation of the Hamiltonian with
[^1]:    1. The sets $R=R_{\xi}$ and $R^{\prime}=R_{\xi}^{\prime}$ are defined between Propositions 4.1.6 and 4.1.8.
[^2]:    2. Not to be confused with the operators $H_{\lambda, \mu}$ from section 4.1.
