# A simple proof that the hp-FEM does not suffer from the pollution effect for the constant-coefficient full-space Helmholtz equation 

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#### Abstract

In $d$ dimensions, accurately approximating an arbitrary function oscillating with frequency $\lesssim k$ requires $\sim k^{d}$ degrees of freedom. A numerical method for solving the Helmholtz equation (with wavenumber $k$ ) suffers from the pollution effect if, as $k \rightarrow \infty$, the total number of degrees of freedom needed to maintain accuracy grows faster than this natural threshold. While the $h$-version of the finite element method (FEM) (where accuracy is increased by decreasing the meshwidth $h$ and keeping the polynomial degree $p$ fixed) suffers from the pollution effect, the $h p$-FEM (where accuracy is increased by decreasing the meshwidth $h$ and increasing the polynomial degree $p$ ) does not suffer from the pollution effect. The heart of the proof of this result is a PDE result splitting the solution of the Helmholtz equation into "high" and "low" frequency components. This result for the constant-coefficient Helmholtz equation in full space (i.e. in $\mathbb{R}^{d}$ ) was originally proved in Melenk and Sauter (Math. Comp 79(272), 1871-1914, 2010). In this paper, we prove this result using only integration by parts and elementary properties of the Fourier transform. The proof in this paper is motivated by the recent proof in Lafontaine et al. (Comp. Math. Appl. 113, 59-69, 2022) of this splitting for the variable-coefficient Helmholtz equation in full space use the more-sophisticated tools of semiclassical pseudodifferential operators.


Keywords Helmholtz equation • High frequency • Finite element method • Pollution effect

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## 1 Introduction and motivation

When computing approximations with the finite element method to the solution of the Helmholtz equation

$$
\begin{equation*}
\Delta u+k^{2} u=-f \tag{1.1}
\end{equation*}
$$

with wavenumber $k>0$, a fundamental question is:
How quickly must the meshwidth $h$ decrease with $k$ and/or the polynomial degree $p$ increase with $k$ to maintain accuracy as $k \rightarrow \infty$ ?

This question has been the subject of sustained interest since the late 1980s, with initially answers obtained for 1-d problems [1, 29, 30], and now answers obtained for 2- and 3-d problems in general geometries, both for "standard" FEMs [14, 16, 19, 24, $32-36,52,55]$ and for variations of these, e.g. discontinuous Galerkin methods [17, $18,37,54]$ and multiscale methods [ $3,7,9,10,20,45,46]$. Moreover, there is large current interest in this question when the Helmholtz equation (1.1) is replaced by its variable-coefficient generalisation $\nabla \cdot(A \nabla u)+k^{2} n u=0[4,6,11,21,22,25,27$, 32, 38] or even the time-harmonic Maxwell equations [39, 40, 44].

A highlight of this body of research is the result from [4, 16, 21, 22, 32, 34, 36, 37] that the $h p$-FEM does not suffer from the pollution effect; i.e. accuracy can be maintained with a choice of the number of degrees of freedom growing like $k^{d}$, where "accuracy" here means that the computed solution is quasi-optimal (see (3.2) below). This is contrast to the $h$-version of the FEM which, e.g. with $p=1$, needs the total number of degrees of freedom to grow like $k^{2 d}$ to maintain accuracy in this sense. Having the number of degrees of freedom growing like $k^{d}$ is the natural threshold for this problem since an oscillatory function in $d$ dimensions with frequency $\lesssim k$ requires $\sim k^{d}$ degrees of freedom to be well-approximated by piecewise polynomials; this is expected from the Nyquist-Shannon-Whittaker sampling theorem in 1-d [50,53], and from the recent results in general dimension in [23].

The proofs that the $h p$-FEM does not suffer from the pollution effect consist of the following three ingredients.

1. Sufficient conditions for FEM solutions to be quasi-optimal originating from the ideas of Schatz [49] (related to the classic "Aubin-Nitsche trick"), and then developed by Sauter [48].
2. Results from [34, Appendices B and C] about how well the hp-FEM spaces approximate analytic functions.
3. A PDE result splitting the solution of the Helmholtz equation into "high" and "low" frequency components.

The motivation for the present paper was the realisation that, for the constantcoefficient Helmholtz equation (1.1) posed in $\mathbb{R}^{d}$, given a bound on the solution in terms of the data (which can be proved using essentially only integration by parts), the splitting in point 3 above can be proved using only elementary properties of the Fourier transform, thus making the key ideas behind this body of work accessible to a wide audience.

The proof in this paper is motivated by the recent proof in [32] of this splitting for the variable-coefficient Helmholtz equation in $\mathbb{R}^{d}$; the proof in [32] uses the moresophisticated tools of semiclassical pseudodifferential operators that reduce to the elementary ones above in the constant-coefficient case - we discuss this further in Section 7.6.

We highlight that the splitting in point 3 above, and also the results of points 1 and 2, are all concerned with the analysis of the $h p-\mathrm{FEM}$; i.e. they influence the implementation of the $h p$-FEM only in that that together they give a prescription of how to tie $h$ and $p$ to $k$ to ensure that the FEM solution is quasi-optimal, uniformly in $k$, as $k \rightarrow \infty$.

Plan of the paper Section 2 recalls basic facts about the Helmholtz equation. Sections 3,4 , and 5 concern points 1,2 , and 3 above, respectively. Section 5 states the splitting for the constant-coefficient full-space Helmholtz equation, and then uses the results of points $1-3$ to prove that the $h p$-FEM does not suffer from the pollution effect when applied to this problem. Section 6 recaps the basic properties of the Fourier transform, and then Section 7 proves the splitting stated in Section 5 using the material in Section 6.

## 2 The Helmholtz equation

### 2.1 The model Helmholtz problem

As in [34], we consider the following model Helmholtz problem. Given $f \in L^{2}\left(\mathbb{R}^{d}\right)$ with compact support, let $u \in H_{\text {loc }}^{1}\left(\mathbb{R}^{d}\right)$ be the solution of

$$
\begin{equation*}
k^{-2} \Delta u+u=-f \text { in } \mathbb{R}^{d}, d=2,3 \tag{2.1}
\end{equation*}
$$

that is outgoing in the sense that it satisfies the Sommerfeld radiation condition

$$
\begin{equation*}
k^{-1} \partial_{r} u(x)-\mathrm{i} u(x)=o\left(r^{-(d-1) / 2}\right) \tag{2.2}
\end{equation*}
$$

as $r:=|x| \rightarrow \infty$, uniformly in $\widehat{x}:=x / r$. Rellich's uniqueness theorem (see, e.g. [13, Theorem 3.13]) implies that the solution of (2.1)-(2.2) is unique. For this particular model problem, the solution can be written down explicitly as an integral of $f$ against the fundamental solution of the Helmholtz equation (see (7.10) below), with then mapping properties of this integral operator (see, e.g. [41, Theorem 6.1]) showing existence of the solution to (2.1)-(2.2). Observe that we have multiplied the Helmholtz equation (1.1) by $k^{-2}$ and rescaled the right-hand side $f$; we see below how this rescaling by $k^{-2}$ allows us to keep better track of the $k$-dependence.

### 2.2 The variational formulation of the Helmholtz equation

Let $R>0$ be large enough so that $\operatorname{supp} f \subset B_{R}:=\left\{x \in \mathbb{R}^{d}:|x|<R\right\}$. The variational formulation of (2.1)-(2.2) is then

$$
\begin{equation*}
\text { find } \tilde{u} \in H^{1}\left(B_{R}\right) \text { such that } a(\tilde{u}, v)=F(v) \text { for all } v \in H^{1}\left(B_{R}\right) \tag{2.3}
\end{equation*}
$$

where

$$
\begin{equation*}
a(\tilde{u}, v):=\int_{B_{R}}\left(k^{-2} \nabla \tilde{u} \cdot \overline{\nabla v}-\tilde{u} \bar{v}\right)-k^{-1}\left\langle\operatorname{DtN}_{k} \tilde{u}, v\right\rangle_{\partial B_{R}} \tag{2.4}
\end{equation*}
$$

and

$$
F(v):=\int_{B_{R}} f \bar{v} .
$$

The operator $\mathrm{DtN}_{k}$ in (2.4) is the Dirichlet-to-Neumann map for the outgoing solution of the Helmholtz equation in the exterior of the ball $B_{R}$; i.e. given $g \in H^{1 / 2}\left(\partial B_{R}\right)$, let $v$ be the unique outgoing solution to

$$
\left(-k^{-2} \Delta-1\right) v=0 \quad \text { in } \mathbb{R}^{d} \backslash \overline{B_{R}} \quad \text { and } \quad v=g \text { on } \partial B_{R}
$$

then $\operatorname{DtN}_{k} g:=k^{-1} \partial_{r} v$, i.e.

$$
\operatorname{DtN}_{k} g(\theta)=\frac{1}{2 \pi} \sum_{n=-\infty}^{\infty} \frac{H_{n}^{(1)^{\prime}}(k R)}{H_{n}^{(1)}(k R)} \exp (\mathrm{i} n \theta) \int_{0}^{2 \pi} \mathrm{e}^{-\mathrm{i} n \theta} g(R, \theta) \mathrm{d} \theta
$$

for the analogous expression when $d=3$, see, e.g. [34, Equations 3.7 and 3.10].
Green's identity and the definitions of $\operatorname{DtN}_{k}$ and $a(\cdot, \cdot)$ imply that if $u$ is a solution of (2.1)-(2.2), then $\left.u\right|_{B_{R}}$ is a solution of the variational problem (2.3). Conversely, if $\tilde{u}$ is a solution of this variational problem, then there exists a solution $u$ of (2.1)(2.2) such that $\left.u\right|_{B_{R}}=\widetilde{u}$; thus, the solution of the variational problem (2.3) exists and is unique. Because of this equivalence result, and for simplicity, from now on, we denote by $u$ both the solution of (2.1)-(2.2) and the solution of the variational problem (2.3).

Remark 2.1 (Approximating $\mathrm{DtN}_{k}$ ) Implementing the operator $\mathrm{DtN}_{k}$ appearing in $a(\cdot, \cdot)(2.4)$ is computationally expensive, and so in practice one seeks to approximate this operator by, e.g. imposing an absorbing boundary condition on $\partial B_{R}$, or using a perfectly-matched layer (PML). For simplicity, in this paper (as in [34]), we analyse the FEM assuming that $\mathrm{DtN}_{k}$ is realised exactly.

### 2.3 The $\boldsymbol{k}$-dependence of the Helmholtz solution operator

We work with the weighted norm

$$
\begin{equation*}
\|v\|_{H_{k}^{m}\left(B_{R}\right)}^{2}:=\sum_{0 \leq|\alpha| \leq m}\left\|\left(k^{-1} \partial\right)^{\alpha} v\right\|_{L^{2}\left(B_{R}\right)}^{2} . \tag{2.5}
\end{equation*}
$$

with special case $\|v\|_{H_{k}^{1}\left(B_{R}\right)}^{2}:=\left\|k^{-1} \nabla v\right\|_{L^{2}\left(B_{R}\right)}^{2}+\|v\|_{L^{2}\left(B_{R}\right)}^{2}$. The rationale for using these norms is that if a function $v$ oscillates with frequency $k$, then we expect $\left|\left(k^{-1} \partial\right)^{\alpha} v\right| \sim|v|$ for all $\alpha$; this is true, e.g. if $v(x)=\exp (\mathrm{i} k x \cdot a)$.

Let $C_{\text {sol }}(k, R)$ be the operator norm of the map $L^{2}\left(B_{R}\right) \ni f \mapsto u \in H^{1}\left(B_{R}\right)$, where $u$ is the outgoing solution of the Helmholtz equation (2.1); i.e. given $k$ and $R$, $C_{\text {sol }}(k, R)$ is such that given $f \in L^{2}\left(B_{R}\right)$ if $u$ is the solution of (2.1)-(2.2), then

$$
\begin{equation*}
\|u\|_{H_{k}^{1}\left(B_{R}\right)} \leq C_{\text {sol }}\|f\|_{L^{2}\left(B_{R}\right)} . \tag{2.6}
\end{equation*}
$$

For the simple model problem (2.1)-(2.2), the following bound on $C_{\text {sol }}$ can be obtained by multiplying the Helmholtz equation (2.1) by a judiciously chosen test function and integrating by parts.

Theorem 2.2 (Morawetz bound on $C_{\text {sol }}$ ) For all $k>0$ and $R>0$,

$$
\begin{equation*}
C_{\mathrm{sol}} \leq 2 k R \sqrt{1+\left(\frac{d-1}{2 k R}\right)^{2}} \tag{2.7}
\end{equation*}
$$

This bound was essentially proved in [42, 43] (although the bound does not quite appear in this form in those papers) (see also [8, Lemma 3.5] and [28, Equations 1.9 and 1.10]). The details of the proof of this bound are not needed in the rest of the paper, but for completeness (and since they involve essentially only integrating by parts), we include them in Appendix A.

The bound (2.7) is sharp in its $k R$ dependence for $k R$ large; indeed, by considering $u(x)=\mathrm{e}^{\mathrm{i} k x_{1}} \chi(|x| / R)$ for $\chi \in C^{\infty}$ supported in [0, 1), one can show that given $k_{0}, R_{0}>0$, there exists $C>0$ such that $C_{\text {sol }} \geq C k R$ for all $k \geq k_{0}$ and $R \geq R_{0}$.

Corollary 2.3 (Bound on $H_{k}^{2}$ norm of the Helmholtz solution) Given $k_{0}, R_{0}>0$, there exists $C>0$ such that the solution $u$ of (2.1)-(2.2) with $\operatorname{supp} f \subset B_{R}$ with $R \geq R_{0}$ satisfies

$$
\begin{equation*}
\|u\|_{H_{k}^{2}\left(B_{R}\right)} \leq C k R\|f\|_{L^{2}\left(B_{R}\right)} \quad \text { for all } k \geq k_{0} . \tag{2.8}
\end{equation*}
$$

Comparing the bounds (2.7) and (2.8) shows the advantage of working in the particular weighted norms (2.5) - the $k$-dependence of the solution operator is the same regardless of the spaces it maps between.

Sketch proof of Corollary 2.3 Corollary 6.4 below uses the Fourier transform to prove $H^{2}$ regularity of the solution of $\left(-k^{-2} \Delta+1\right) v=g \in L^{2}\left(\mathbb{R}^{d}\right)$. To apply this bound to the solution of (2.1)-(2.2), let $\varphi \in C_{\text {comp }}^{\infty}\left(\mathbb{R}^{d},[0,1]\right)$ be equal to one on $B_{R}$ and vanish outside $B_{2 R}$, and then let $v:=\varphi u$ so that $g=\left(-k^{-2} \Delta+1\right)(\varphi u)=$ $\left(-k^{-2} \Delta-1\right)(\varphi u)+2 \varphi u$. The bound (2.8) then follows by bounding $\|g\|_{L^{2}\left(B_{2 R}\right)}$ in terms of $\|f\|_{L^{2}\left(B_{R}\right)}$ using the PDE (2.1) and the bounds (2.6) and (2.7).

Remark 2.4 Many papers on the numerical analysis of the Helmholtz equation use the weighted $H^{1}$ norm $\|u\|_{H_{k}^{1}\left(B_{R}\right)}^{2}:=\|\nabla u\|_{L^{2}\left(B_{R}\right)}^{2}+k^{2}\|u\|_{L^{2}\left(B_{R}\right)}^{2}$; we use (2.5) instead since weighting the $j$ th derivative by $k^{-j}$ is easier to keep track of than weighting it by $k^{-j+1}$ (especially for high derivatives).

## 3 The Galerkin method and sufficient conditions for quasioptimality

Let $\mathcal{H}_{N}$ be a finite-dimensional subspace of $H^{1}\left(B_{R}\right)$. The Galerkin method applied to the variational problem (2.3) is

$$
\begin{equation*}
\text { find } u_{N} \in \mathcal{H}_{N} \text { such that } a\left(u_{N}, v_{N}\right)=F\left(v_{N}\right) \text { for all } v_{N} \in \mathcal{H}_{N} \tag{3.1}
\end{equation*}
$$

The FEM is the Galerkin method (3.1) with $\mathcal{H}_{N}$ consisting of piecewise polynomials (we describe in Section 4 the specific assumptions we make on $\mathcal{H}_{N}$ ).

Given a sequence of finite-dimensional spaces $\left\{\mathcal{H}_{N}\right\}_{N=1}^{\infty}$, a standard error bound one seeks to prove on the sequence of Galerkin solutions $\left\{u_{N}\right\}_{N=1}^{\infty}$ is the following quasioptimal error bound: there exists a $C_{\text {qo }}>0$ and $N_{0} \in \mathbb{N}$ such that, for $N \geq N_{0}$,

$$
\begin{equation*}
\left\|u-u_{N}\right\|_{H_{k}^{1}\left(B_{R}\right)} \leq C_{\mathrm{qo}} \min _{v_{N} \in \mathcal{H}_{N}}\left\|u-v_{N}\right\|_{H_{k}^{1}\left(B_{R}\right)} . \tag{3.2}
\end{equation*}
$$

The main result of this section is Lemma 3.4 below, giving sufficient conditions for quasioptimality. This result crucially relies on the following properties of the sesquilinear form $a(\cdot, \cdot)$ and the following properties of the adjoint solution operator.

Lemma 3.1 (Properties of $a(\cdot, \cdot)$ )
(i) (Continuity) Given $k_{0}, R_{0}>0$ there exists $C_{\text {cont }}>0$ such that for all $k \geq k_{0}$ and $R \geq R_{0}$,

$$
|a(u, v)| \leq C_{\text {cont }}\|u\|_{H_{k}^{1}\left(B_{R}\right)}\|v\|_{H_{k}^{1}\left(B_{R}\right)} \quad \text { for all } u, v \in H^{1}\left(B_{R}\right) .
$$

(ii) (Gårding inequality)

$$
\Re a(v, v) \geq\|v\|_{H_{k}^{1}\left(B_{R}\right)}^{2}-2\|v\|_{L^{2}\left(B_{R}\right)}^{2} \quad \text { for all } v \in H^{1}\left(B_{R}\right) .
$$

References for the proof Part (i) follows from the Cauchy-Schwarz inequality and boundedness of $\mathrm{DtN}_{k}$; see [34, Lemma 3.3, Part 1]. Part (ii) follows from the fact that $-\mathfrak{R}\left\langle\operatorname{DtN}_{k} \phi, \phi\right\rangle_{\partial B_{R}} \geq 0$ for all $\phi \in H^{1 / 2}\left(\partial B_{R}\right)$; see [34, Lemma 3.3, Part 2].

Definition 3.2 (Adjoint solution operator $\mathcal{S}^{*}$ ) Given $f \in L^{2}\left(B_{R}\right)$, let $\mathcal{S}^{*} f \in$ $H^{1}\left(B_{R}\right)$ be defined by

$$
\begin{equation*}
a\left(v, \mathcal{S}^{*} f\right)=(v, f)_{L^{2}\left(B_{R}\right)} \text { for all } v \in H^{1}\left(B_{R}\right) \tag{3.3}
\end{equation*}
$$

The following lemma shows that our knowledge about outgoing Helmholtz solutions immediately gives us knowledge about $\mathcal{S}^{*}$.

Lemma 3.3 If $\mathcal{S}^{*}$ is defined as in (3.3) then

$$
a\left(\overline{\mathcal{S}^{*} f}, v\right)=(\bar{f}, v)_{L^{2}\left(B_{R}\right)} \quad \text { for all } v \in H^{1}\left(B_{R}\right)
$$

i.e. $\mathcal{S}^{*} f$ is the complex-conjugate of the outgoing Helmholtz solution with data $\bar{f}$.

Sketch proof Green's identity and the radiation condition (2.2) show that
$\left\langle\operatorname{DtN}_{k} \psi, \bar{\phi}\right\rangle_{\partial B_{R}}=\left\langle\operatorname{DtN}_{k} \phi, \bar{\psi}\right\rangle_{\partial B_{R}}$ for all $\phi, \psi \in H^{1 / 2}\left(\partial B_{R}\right)$. This implies that $a(\bar{v}, u)=a(\bar{u}, v)$ for all $u, v$, which implies the result.

Lemma 3.4 (Sufficient conditions for quasi-optimality) If

$$
\begin{equation*}
\eta\left(\mathcal{H}_{N}\right):=\sup _{0 \neq f \in L^{2}\left(B_{R}\right)} \min _{v_{N} \in \mathcal{H}_{N}} \frac{\left\|\mathcal{S}^{*} f-v_{N}\right\|_{H_{k}^{1}\left(B_{R}\right)}}{\|f\|_{L^{2}\left(B_{R}\right)}} \leq \frac{1}{2 C_{\mathrm{cont}}} \tag{3.4}
\end{equation*}
$$

then the Galerkin solution $u_{N}$ to the variational problem (3.1) exists, is unique, and satisfies the quasi-optimal error bound

$$
\begin{equation*}
\left\|u-u_{N}\right\|_{H_{k}^{1}\left(B_{R}\right)} \leq 2 C_{\mathrm{cont}}\left(\min _{v_{N} \in V_{N}}\left\|u-v_{N}\right\|_{H_{k}^{1}\left(B_{R}\right)}\right) \tag{3.5}
\end{equation*}
$$

References for the proof See [48, Theorem 2.5] or [34, Theorem 4.3] (note that these references consider the Helmholtz equation in the form $\Delta u+k^{2} u=-f$ and use the weighted norm discussed in Remark 2.4, but it is straightforward to convert the results to our setting).

## 4 Recap of approximation results in hp-FEM spaces

The result that the $h p$-FEM does not suffer from the pollution effect is proved under the following two assumptions on the finite-dimensional subspaces; these assumptions describe how well (as a function of $h$ and $p$ ) the spaces approximate functions with a given regularity.

We highlight immediately that both these assumptions are satisfied by $h p$ -finite-element spaces with curved elements that fit $\partial B_{R}$ exactly, provided that the triangulations are quasi-uniform and are constructed by refining a fixed triangulation that has analytic element maps (see Theorem 4.4 below). Nevertheless, we formulate these properties as specific assumptions to make it clear the actual properties of the subspaces that are needed in the proof of the result that the $h p$-FEM does not suffer from the pollution effect. Since we are ultimately thinking of the subspaces in these assumptions as $h p$-finite-element spaces, we denote the sequence of these subspaces as $\left\{\mathcal{H}_{h, p}\right\}_{h>0, p \in \mathbb{Z}^{+}}$.

Assumption 4.1 (Approximation in the finite-dimensional subspace of functions with finite regularity) Let $\left\{\mathcal{H}_{h, p}\right\}_{h>0, p \in \mathbb{Z}^{+}}$be a sequence of finite-dimensional subspaces of $H^{1}\left(B_{R}\right)$. Given $s, d$ with $d$ the spatial dimension and $s>d / 2$, there exists $C_{\text {approx }_{1}}>0$ such that if $v \in H^{s}\left(B_{R}\right)$ and $p \geq s-1$, then

$$
\begin{equation*}
\min _{w_{h, p} \in \mathcal{H}_{h, p}}\left\|v-w_{h, p}\right\|_{H_{k}^{1}\left(B_{R}\right)} \leq C_{\text {approx }_{1}}\left(\frac{h k}{p}\right)^{s-1}\left(1+\frac{h k}{p}\right)\|v\|_{H_{k}^{s}\left(B_{R}\right)} . \tag{4.1}
\end{equation*}
$$

Discussion of Assumption 4.1 A standard polynomial approximation result is the following. For $d=2,3$, given $s \geq 2$ and $p \geq s-1$, there exists $C>0$ such that if $v \in H^{s}(D)$ and $m=0$ or 1 , then

$$
\begin{equation*}
\left\|v-\mathcal{I}^{h} v\right\|_{H^{m}(D)} \leq C h^{s-m}|v|_{H^{s}(D)} \tag{4.2}
\end{equation*}
$$

where $\mathcal{I}^{h}$ is a global interpolation operator (see, e.g. [5, Equation 4.4.28], [12, Theorem 17.1]). The approximation result (4.1) is a generalisation of (4.2) which (i) makes explicit the dependence on $p$ of the constant $C$ in (4.2), and (ii) works in norms weighted with $k$.

Motivation for Assumption 4.3. Assumption 4.1 is about approximating in $k$ weighted norms an arbitrary function in $H^{s}$ for some $s>0$; note that the constant $C_{\text {approx }}{ }_{1}$ depends on $s$ in an unspecified way, and so we cannot use the bound (4.1) for arbitrarily large $s$, and hence arbitrarily large $p$, since $p$ is tied to $s$ via $p \geq s-1$. The next assumption, Assumption 4.3, allows us to take arbitrarily large $p$; the price one pays is that the function being approximated must be analytic.

Before stating Assumption 4.3, we recall the relationship between derivative bounds and analyticity for families of functions depending on $k$.

Lemma 4.2 ( $k$-explicit analyticity) Let $D$ be a bounded open subset of $\mathbb{R}^{d}$ and let $u \in C^{\infty}(D)$ be a family of functions depending on $k$.
(i) If there exist $C, C_{u}>0$, independent of $\alpha$, such that

$$
\begin{equation*}
\left\|\partial^{\alpha} u\right\|_{L^{2}(D)} \leq C_{u}(C k)^{|\alpha|} \quad \text { for all multiindices } \alpha, \tag{4.3}
\end{equation*}
$$

then $u$ is real analytic in $D$ and its power series has infinite radius of convergence, i.e. $u$ can be extended to an entire function on $\mathbb{R}^{d}$.
(ii) If there exist $C, C_{u}>0$, independent of $\alpha$, such that

$$
\left\|\partial^{\alpha} u\right\|_{L^{2}(D)} \leq C_{u}(C k)^{|\alpha|}|\alpha|!\quad \text { for all multiindices } \alpha,
$$

then $u$ is real analytic in $D$ with radius of convergence of its power series proportional to $(C k)^{-1}$.
(iii) If there exist $C, C_{u}>0$, independent of $\alpha$, such that

$$
\left\|\partial^{\alpha} u\right\|_{L^{2}(D)} \leq C_{u} C^{|\alpha|} \max \{|\alpha|, k\}^{|\alpha|} \quad \text { for all multiindices } \alpha,
$$

then $u$ is real analytic in $D$ with radius of convergence of its power series proportional to $C^{-1}$ and independent of $k$.

Sketch proof In each case, use the Sobolev embedding theorem (see, e.g. [41, Theorem 3.26]) to obtain a bound on $\left\|\partial^{\alpha} u\right\|_{L^{\infty}(D)}$, and then use this to bound the Lagrange form of the remainder in the Taylor series (see, e.g. [34, Proof of Lemma C.2]).

In the rest of the paper, we only use the class of functions in part (i) of Lemma 4.2 , but the classes in parts (ii) and (iii) are included for context.

Assumption 4.3 (Approximation in the finite-dimensional subspace of the class of functions in part (i) of Lemma 4.2) Let $\left\{\mathcal{H}_{h, p}\right\}_{h>0, p \in \mathbb{Z}^{+}}$be a sequence of finitedimensional subspaces of $H^{1}\left(B_{R}\right)$, and suppose $v \in C^{\infty}\left(B_{R}\right)$ is such that, given $k_{0}>0$ there exists $C_{1}, C_{2}>0$ such that

$$
\begin{equation*}
\left\|\left(k^{-1} \partial\right)^{\alpha} v\right\|_{L^{2}\left(B_{R}\right)} \leq C_{1}\left(C_{2}\right)^{|\alpha|} \quad \text { for all } \alpha \text { and for all } k \geq k_{0} . \tag{4.4}
\end{equation*}
$$

Given $\widetilde{C}$, there exist $\sigma, C_{\text {approx }_{2}}>0$, depending on $C_{2}$ and $\widetilde{C}$ (but not $C_{1}$ ), such that, if $k \geq k_{0}$ and $k, h$, and $p$ satisfy

$$
\begin{equation*}
\frac{h}{R}+\frac{h k}{p} \leq \widetilde{C} \tag{4.5}
\end{equation*}
$$

then

$$
\begin{equation*}
\min _{w_{h, p} \in \mathcal{H}_{h, p}}\left\|v-w_{h, p}\right\|_{H_{k}^{1}\left(B_{R}\right)} \leq C_{1} C_{\text {approx }_{2}}\left[\frac{1}{k R}\left(\frac{h / R}{h / R+\sigma}\right)^{p}+\left(\frac{h k}{\sigma p}\right)^{p}\right] . \tag{4.6}
\end{equation*}
$$

Discussion of Assumption 4.3 The key points about the bound (4.6) are the following.
(i) For fixed $p$ and $k$, as $h \rightarrow 0$, the right hand side of (4.6) is $O\left(h^{p}\right)$ (just like the right-hand side of (4.1) with $s=p+1$ ).
(ii) If $h k /(\sigma p)<1$ then the right-hand side of (4.6) decreases exponentially as $p \rightarrow \infty$.
(iii) The quantities $h / R, h k / p$, and $k R$ are dimensionless, and thus the right-hand sides of (4.5) and (4.6) involve only dimensionless quantities.

Approximation spaces satisfying Assumptions 4.1 and 4.3 Let $\left(\mathcal{T}_{h}\right)_{0<h \leq h_{0}}$ (with $h$ the maximum element diameter) be a sequence of triangulations of $B_{R}$, with each element $K \in \mathcal{T}_{h}$ the image of a reference element $\widehat{K}$ (a reference triangle in 2-d and a reference tetrahedron in 3-d) under the map $F_{K}: \widehat{K} \rightarrow K$. As is standard, we assume there are no hanging nodes and that the element maps of elements sharing an edge or face induce the same parametrisation on that edge or face. We consider the $h p$-finite-element spaces

$$
\begin{equation*}
\mathcal{H}^{p}\left(\mathcal{T}_{h}\right):=\left\{v \in H^{1}\left(B_{R}\right): \text { for each } K \in \mathcal{T}_{h},\left.v\right|_{K} \circ F_{K} \text { is a polynomial of degree } \leq p\right\} \tag{4.7}
\end{equation*}
$$

Since $B_{R}$ is curved, we consider triangulations with curved elements that fit $\partial B_{R}$ exactly (thus avoiding the issue of analysing the non-conforming error coming from using simplicial triangulations (see, e.g. [5, Chapter 10])). Recall that the family $\left(\mathcal{T}_{h}\right)_{0<h \leq h_{0}}$ is quasi-uniform if there exists $C>0$ such that

$$
h:=\max _{K \in \mathcal{T}_{h}} \operatorname{diam}(K) \leq C \min _{K \in \mathcal{T}_{h}} \operatorname{diam}(K) \quad \text { for all } 0<h \leq h_{0} ;
$$

for such triangulations, the dimension of $\mathcal{H}^{p}\left(\mathcal{T}_{h}\right)$ is proportional to $(p /(h / R))^{d}$.
Theorem 4.4 (Conditions under which Assumptions 4.1 and 4.3 hold) If $\left(\mathcal{T}_{h}\right)_{0<h \leq h_{0}}$ satisfies [34, Assumption 5.2], then $\mathcal{H}^{p}\left(\mathcal{T}_{h}\right)$ defined by (4.7) satisfies Assumptions 4.1 and 4.3.

Informally, [34, Assumption 5.2] is that $\left(\mathcal{T}_{h}\right)_{0<h \leq h_{0}}$ is quasi-uniform with each element map $F_{K}$ the composition of a affine map and an analytic map; [36, Remark 5.2] notes that $\left(\mathcal{T}_{h}\right)_{0<h \leq h_{0}}$ satisfying this assumption can be constructed by refining a fixed triangulation that has analytic element maps.

References for the proof of Theorem 4.4 That Assumption 4.1 holds follows from [34, Theorem B.4] (a result about approximation on the reference element) and a scaling argument (see [34, Bottom of Page 1895]). For Assumption 4.3, the bound (4.6) is proved in the course of [34, Proof of Theorem 5.5], see the last equation on [34, Page 1896]; note that (i) we have simplified this equation using the assumption (4.5), and (ii) the weighted $H^{1}$ norm in [34] is $k$ times $\|\cdot\|_{H_{k}^{1}\left(B_{R}\right)}$ defined by (2.5).

## 5 The splitting of the Helmholtz solution and the proof that the hp-FEM does not suffer from the pollution effect

### 5.1 Statement of the splitting

The crucial result used to prove that the $h p$-FEM applied to the problem (2.1)-(2.2) does not suffer from the pollution effect is the following (with the proof contained in Section 7).

Theorem 5.1 (Splitting of the Helmholtz solution) Given $k_{0}, R_{0}>0$, there exists $C_{\text {split, } H^{2}}, C_{\text {split, } \mathcal{A}}>0$ such that the following holds. Given $f \in L^{2}\left(B_{R}\right)$ with $R \geq R_{0}$, let $u$ satisfy the Helmholtz equation (2.1) and the Sommerfeld radiation condition (2.2). Then

$$
\begin{equation*}
\left.u\right|_{B_{R}}=u_{H^{2}}+u_{\mathcal{A}} \tag{5.1}
\end{equation*}
$$

where $u_{H^{2}} \in H^{2}\left(B_{R}\right)$ with

$$
\begin{equation*}
\left\|u_{H^{2}}\right\|_{H_{k}^{2}\left(B_{R}\right)} \leq C_{\text {split }, H^{2}}\|f\|_{L^{2}\left(B_{R}\right)} \quad \text { for all } k \geq k_{0} \tag{5.2}
\end{equation*}
$$

and $u_{\mathcal{A}} \in C^{\infty}\left(B_{R}\right)$ with

$$
\begin{equation*}
\left\|\left(k^{-1} \partial\right)^{\alpha} u_{\mathcal{A}}\right\|_{L^{2}\left(B_{R}\right)} \leq C_{\text {sol }}(k, 2 R)\left(C_{\text {split, } \mathcal{A}}\right)^{|\alpha|}\|f\|_{L^{2}\left(B_{R}\right)} \tag{5.3}
\end{equation*}
$$

for all $\alpha$ and for all $k \geq k_{0}$.
Discussion of the properties of $u_{H^{2}}$ and $u_{\mathcal{A}}$ in Theorem 5.1 Recall that the solution $u$ itself satisfies the bound (2.8); i.e. $\|u\|_{H_{k}^{2}\left(B_{R}\right)} \leq C k R\|f\|_{L^{2}\left(B_{R}\right)}$. Therefore,
(i) the bound (5.2) on $u_{H^{2}}$ is one power of $k$ better than the corresponding bound (2.8) on $u$, and
(ii) the bound (5.3) on $u_{\mathcal{A}}$ has the same $k$ dependence as the bound (2.8) on $u$ - both are governed by $C_{\text {sol }}$ - although $u_{\mathcal{A}}$ is $C^{\infty}$ (with each derivative incurring a power of $k$ ), and indeed analytic by Lemma 4.2.

We discuss in Section 5.3 why both of these points are crucial in proving that the $h p$-FEM does not suffer from the pollution effect.

We see in the proof of Theorem 5.1 in Section 7 that $u_{H^{2}}$ corresponds to components of $u$ with frequencies $\geq \lambda k$ and $u_{\mathcal{A}}$ corresponds to components of $u$ with frequencies $\leq \lambda k$, where $\lambda>1$, and the notion of "frequencies" is understood via the Fourier transform. We see in Section 7.4 below that $u_{H^{2}}$ satisfies the property
(i) above because the Helmholtz operator is "well behaved" (in a sense made precise below) on frequencies $\geq \lambda k$ with $\lambda>1$. We see in Section 7.3 that $u_{\mathcal{A}}$ satisfies the property (ii) above because a function with a compactly-supported Fourier transform is analytic.

### 5.2 The hp-FEM does not suffer from the pollution effect

Theorem 5.2 (Quasioptimality of the $h p$-FEM) Suppose that $\left\{\mathcal{H}_{h, p}\right\}_{h>0, p \in \mathbb{Z}^{+}}$satisfy Assumptions 4.1 and 4.3. Given $k_{0}, R_{0}>0$, there exist $C_{1}, C_{2}>0$ (independent of $k, R, h$, and p) such that the following holds. If $u$ is the solution of the variational problem (2.3), $k \geq k_{0}$,

$$
\begin{equation*}
\frac{h k}{p} \leq C_{1}, \quad \text { and } \quad p \geq C_{2} \log (k R) \tag{5.4}
\end{equation*}
$$

then the Galerkin solution exists, is unique, and satisfies the quasi-optimal error bound (3.5).

The pollution effect occurs when no choice of the number of degrees of freedom growing like $(k R)^{d}$ ensures that the quasi-optimal error bound (3.2) holds with $C_{\text {qo }}$ independent of $k$ (see [2, Definition 2.1] or [26, Equation 1.5] for more-precise statements of this). Since the number of degrees of freedom of $\mathcal{H}_{h, p}$ is proportional to $(p /(h / R))^{d}$, if $h$ and $p$ are chosen so that the inequalities (5.4) hold with equality, then the number of degrees of freedom of $\mathcal{H}_{h, p}$ is proportional to $(k R)^{d}$; i.e. Theorem 5.2 shows that the hp-FEM does not suffer from the pollution effect.

Proof of Theorem 5.2 The plan is to show that there exist $C_{1}, C_{2}>0$ such that if $h, k$, and $p$ satisfy (5.4), then the inequality (3.4) holds; the result then follows from Lemma 3.4. By Lemma 3.3, we can consider $\mathcal{S}^{*} f$ to be the solution of (2.1)-(2.2); we then use Theorem 5.1 to split $u$ into $u_{H^{2}}$ and $u_{\mathcal{A}}$, approximate $u_{H^{2}}$ using Assumption 4.1 (with $s=2$ ), and approximate $u_{\mathcal{A}}$ by Assumption 4.3. By the bounds (4.1) and (5.2), there exists $v_{h, p}^{(1)} \in \mathcal{H}_{h, p}$ such that

$$
\begin{align*}
\frac{\left\|u_{H^{2}}-v_{h, p}^{(1)}\right\|_{H_{k}^{1}\left(B_{R}\right)}}{\|f\|_{L^{2}\left(B_{R}\right)}} & \leq C_{\text {approx }_{1}}\left(1+\frac{h k}{p}\right)\left(\frac{h k}{p}\right)=\frac{\|u\|_{H_{k}^{2}\left(B_{R}\right)}}{\|f\|_{L^{2}\left(B_{R}\right)}} \\
& \leq C_{\text {approx }_{1}}\left(1+\frac{h k}{p}\right)\left(\frac{h k}{p}\right) C_{\text {split }, H^{2}} . \tag{5.5}
\end{align*}
$$

The bound (5.3) implies that $u_{\mathcal{A}}$ satisfies the conditions of Assumption 4.3 with $C_{2}:=C_{\text {split, } \mathcal{A}}$ and $C_{1}:=C_{\text {sol }}(k, 2 R)\|f\|_{L^{2}\left(B_{R}\right)}$. Therefore, by (4.6) and the bound (2.7) on $C_{\text {sol }}$, there exists $C>0$ and $v_{h, p}^{(2)} \in \mathcal{H}_{h, p}$ such that

$$
\begin{equation*}
\frac{\left\|u_{\mathcal{A}}-v_{h, p}^{(2)}\right\|_{H_{k}^{1}\left(B_{R}\right)}}{\|f\|_{L^{2}\left(B_{R}\right)}} \leq C_{\text {approx }_{2}} C\left[\left(\frac{h}{h+\sigma}\right)^{p}+k R\left(\frac{h k}{\sigma p}\right)^{p}\right] . \tag{5.6}
\end{equation*}
$$

Let $v_{h, p}:=v_{h, p}^{(1)}+v_{h, p}^{(2)}$. Using the triangle inequality and the decomposition $u=$ $u_{H^{2}}+u_{\mathcal{A}}$ on $B_{R}$, we obtain that

$$
\begin{equation*}
\eta\left(\mathcal{H}_{h, p}\right) \leq C_{\mathrm{approx}_{1}}\left(1+\frac{h k}{p}\right)\left(\frac{h k}{p}\right) C_{\text {split }, H^{2}}+C_{\text {approx }_{2}} C\left[\left(\frac{h}{h+\sigma}\right)^{p}+k R\left(\frac{h k}{\sigma p}\right)^{p}\right] . \tag{5.7}
\end{equation*}
$$

Therefore, to prove the bound (3.4) on $\eta\left(\mathcal{H}_{h, p}\right)$, it is sufficient to prove that the right-hand sides of (5.5) and (5.6) are each $\leq C_{\text {cont }} / 4$. To do this, first recall from Lemma 3.1 that $C_{\text {cont }}$ is independent of $k$. We then choose $C_{1}$ sufficiently small so that $C_{1} \leq \min \{\widetilde{C}, \sigma\}$ (where $\widetilde{C}$ and $\sigma$ are as in Assumption 4.3) and

$$
C_{\text {approx }}^{1} C_{1}\left(1+C_{1}\right) C_{\text {split }, H^{2}} \leq C_{\text {cont }} / 4 ;
$$

observe that, since $h k / p \leq C_{1}$, this last inequality is the desired bound on the righthand side of (5.5). Next let $\theta_{1}:=h /(\sigma+h)$ and $\theta_{2}:=C_{1} / \sigma$; observe that $\theta_{1}<1$ by definition, and $\theta_{2}<1$ by the definition of $C_{1}$. The right-hand side of (5.6) is then bounded by

$$
C_{\mathrm{approx}_{2}} C\left[\left(\theta_{1}\right)^{p}+k R\left(\theta_{2}\right)^{p}\right] .
$$

Since $\theta_{1}, \theta_{2}<1$, if $p \geq C_{2} \log (k R)$ for $C_{2}$ sufficiently large, then the decay of $\left(\theta_{2}\right)^{p}$ beats the growth of $k R$; thus, with $C_{2}$ sufficiently large, the right-hand side of (5.6) can be made $\leq C_{\text {cont }} / 4$ and the proof is complete.

### 5.3 Discussion of the insight the splitting gives into the pollution effect

This subsection discusses three natural questions:

1. Why is the splitting of Theorem 5.1 needed? That is, why does the no-pollution result not just follow from using the bound (2.8) on $u$ itself to bound $\eta\left(\mathcal{H}_{h, p}\right)$ ?
2. How are the properties of $u_{H^{2}}$ and $u_{\mathcal{A}}$ used to prove Theorem 5.2 (the no-pollution result)?
3. Why does one need $p \rightarrow \infty$ to remove the pollution effect?

Regarding 1: inputting the bound (2.8) on $u$ into the approximation result (4.1) in Assumption 4.1, we obtain that

$$
\begin{equation*}
\eta\left(\mathcal{H}_{h, p}\right) \leq C_{\text {approx }_{1}}\left(1+\frac{h k}{p}\right)\left(\frac{h k}{p}\right) C k R, \tag{5.8}
\end{equation*}
$$

which leads to the condition " $h k^{2} R / p$ sufficiently small" for quasioptimality. The condition " $h k^{2} R$ sufficiently small" is indeed the observed sharp condition for quasioptimality of the $h$-FEM when $p=1$ - see, e.g. [31, Fig. 8] - but in this case, the total number of degrees of freedom grows like $(k R)^{2 d}$; i.e. the $h$-FEM with $p=1$ suffers from the pollution effect.

Regarding 2 and 3: $u_{H^{2}}$ satisfying the bound (5.2), which is $k R$ better than the corresponding bound (2.8) on $u$, allows us to obtain the condition " $h k / p$ sufficiently small" for making the first term on the right-hand side of (5.7) small, as opposed to the condition " $h k^{2} R / p$ sufficiently small" for making the right-hand side of (5.8) small.

When $p$ is fixed, to make the second term on the right-hand side of (5.7) small, we need " $k R(h k)^{p}$ sufficiently small", and the same condition is obtained if we use the approximation result of Assumption 4.1 to approximate $u_{\mathcal{A}}$ instead of that of Assumption 4.3 (i.e. if we ignore the fact that $u_{\mathcal{A}}$ is analytic, and just use that $u_{\mathcal{A}} \in$ $H^{s}\left(B_{R}\right)$ for every $s>0$ ). The recent polynomial-approximation results of [23] show that, for fixed $p, \eta\left(\mathcal{H}_{h, p}\right) \geq C k R(h k)^{p}$ (or, more generally, $\eta\left(\mathcal{H}_{h, p}\right) \geq C C_{\text {sol }}(h k)^{p}$, where $C_{\text {sol }}$ is defined by (2.6)). That is, for fixed $p$, the condition " $k R(h k)^{p}$ sufficiently small" is the sharp condition for ensuring that $\eta\left(\mathcal{H}_{h, p}\right)$ is sufficiently small, and this condition is observed empirically to be the sharp condition required for the Galerkin method to be quasi-optimal with constant independent of $k$ (see, e.g. [11, Figs. 3, 5, and 8] for $p=1,2,3,4$ ); i.e. the $h$-FEM suffers from the pollution effect.

Since $u_{\mathcal{A}}$ is analytic, we can take $p \rightarrow \infty$ in the second term on the right-hand side of (5.7), with, importantly, the approximation result of Assumption 4.3 controlling the dependence of the constant on $p$ (as highlighted in the "Motivation for Assumption 4.3" paragraph in Section 4). The growth of $k R$ is then removed by the exponential decrease of $(h k / \sigma p)^{p}$ when $h k /(\sigma p)<1$. Note that if $k R$ were replaced by $(k R)^{M}$ for any fixed $M>0$, then this growth would also be removed by the exponential decrease of $(h k / \sigma p)^{p}$ when $h k /(\sigma p)<1$. Since $k R$ in (5.7) comes from $C_{\text {sol }}$, the $h p$-FEM results in $[4,16,21,22,32,36,37]$ all involve the assumption that $C_{\text {sol }}$ is polynomially bounded in $k R$.

## 6 Recap of results about the Fourier transform and Fourier multipliers

### 6.1 The Fourier transform $\mathcal{F}_{\boldsymbol{k}}$

Given $k>0$, let

$$
\begin{equation*}
\mathcal{F}_{k} \phi(\xi):=\int_{\mathbb{R}^{d}} \exp (-\mathrm{i} k x \cdot \xi) \phi(x) \mathrm{d} x \tag{6.1}
\end{equation*}
$$

i.e. $\mathcal{F}_{k}$ is the standard Fourier transform with frequency variable scaled by $k$. The reason for including this scaling is that $\mathcal{F}_{k}$ is then tailor-made to work in the weighted Sobolev spaces $H_{k}^{s}$ (see Section 6.2).

Let $\mathscr{S}\left(\mathbb{R}^{d}\right)$ be the Schwartz space of rapidly decreasing, $C^{\infty}$ functions; i.e.

$$
\mathscr{S}\left(\mathbb{R}^{d}\right):=\left\{\phi \in C^{\infty}\left(\mathbb{R}^{d}\right): \sup _{x \in \mathbb{R}^{d}}\left|x^{\alpha} \partial^{\beta} \phi(x)\right|<\infty \text { for all multiindices } \alpha \text { and } \beta\right\} .
$$

Let $\mathscr{S}^{*}\left(\mathbb{R}^{d}\right)$ be the space of continuous linear functionals on $\mathscr{S}\left(\mathbb{R}^{d}\right)$. Recall that $\mathcal{F}_{k}: \mathscr{S}\left(\mathbb{R}^{d}\right) \rightarrow \mathscr{S}\left(\mathbb{R}^{d}\right)$ (see, e.g. [41, Page 72], [47, Proposition 13.15]); then $\mathcal{F}_{k}:$ $\mathscr{S}^{*}\left(\mathbb{R}^{d}\right) \rightarrow \mathscr{S}^{*}\left(\mathbb{R}^{d}\right)$ via the definition that $\left\langle\mathcal{F}_{k} \phi, \psi\right\rangle:=\left\langle\phi, \mathcal{F}_{k} \psi\right\rangle$ for $\phi \in \mathscr{S}^{*}\left(\mathbb{R}^{d}\right)$ and $\psi \in \mathscr{S}\left(\mathbb{R}^{d}\right)$, where $\langle\cdot, \cdot\rangle$ is the duality pairing between $\mathscr{S}^{*}\left(\mathbb{R}^{d}\right)$ and $\mathscr{S}\left(\mathbb{R}^{d}\right)$. In the next subsection, we consider $\langle\xi\rangle^{s} \mathcal{F}_{k} \phi$ for $\phi \in \mathscr{S}^{*}\left(\mathbb{R}^{d}\right)$ and $\langle\xi\rangle:=\left(1+|\xi|^{2}\right)^{1 / 2}$; this is defined as an element of $\mathscr{S}^{*}\left(\mathbb{R}^{d}\right)$ by $\left\langle\langle\xi\rangle^{s} \mathcal{F}_{k} \phi, \psi\right\rangle:=\left\langle\mathcal{F}_{k} \phi,\langle\xi\rangle^{s} \psi\right\rangle$ for $\psi \in$ $\mathscr{S}\left(\mathbb{R}^{d}\right)$ (see, e.g. [47, Propositions 13.14 and 13.17]).

We recall the Fourier inversion theorem

$$
\mathcal{F}_{k}^{-1} \psi(x):=\left(\frac{k}{2 \pi}\right)^{d} \int_{\mathbb{R}^{d}} \exp (\mathrm{i} k x \cdot \xi) \psi(\xi) \mathrm{d} \xi
$$

the property

$$
\begin{equation*}
\mathcal{F}_{k}\left(\left(-\mathrm{i} k^{-1} \partial\right)^{\alpha} \phi\right)(\xi)=\xi^{\alpha} \mathcal{F}_{k} \phi(\xi), \tag{6.2}
\end{equation*}
$$

and Plancherel's theorem

$$
\begin{equation*}
\|\phi\|_{L^{2}\left(\mathbb{R}^{d}\right)}=\left(\frac{k}{2 \pi}\right)^{d / 2}\left\|\mathcal{F}_{k} \phi\right\|_{L^{2}\left(\mathbb{R}^{d}\right)} \tag{6.3}
\end{equation*}
$$

### 6.2 Sobolev spaces weighted with $k$

The natural spaces in which to study solutions of the Helmholtz equation are Sobolev spaces with derivatives weighted with $k$, as in (2.5). On $\mathbb{R}^{d}$, these are naturally defined using $\mathcal{F}_{k}$. For $s \in \mathbb{R}$, let

$$
H_{k}^{s}\left(\mathbb{R}^{d}\right):=\left\{u \in \mathscr{S}^{*}\left(\mathbb{R}^{d}\right),\langle\xi\rangle^{s} \mathcal{F}_{k} u \in L^{2}\left(\mathbb{R}^{d}\right)\right\}, \quad \text { where }\langle\xi\rangle:=\left(1+|\xi|^{2}\right)^{1 / 2}
$$

and let

$$
\begin{equation*}
\left\|\|u\|_{H_{k}^{s}\left(\mathbb{R}^{d}\right)}^{2}:=\left(\frac{k}{2 \pi}\right)^{d} \int_{\mathbb{R}^{d}}\langle\xi\rangle^{2 s}\left|\mathcal{F}_{k} u(\xi)\right|^{2} \mathrm{~d} \xi .\right. \tag{6.4}
\end{equation*}
$$

Because of (6.2), up to dimension-dependent constants, $\left\|\|u\|_{H_{k}^{s}\left(\mathbb{R}^{d}\right)}\right.$ defined by (6.4) is equivalent to $\|u\|_{H_{k}^{s}\left(\mathbb{R}^{d}\right)}$ defined by (2.5) (with $B_{R}$ replaced by $\left.\mathbb{R}^{d}\right)$.

### 6.3 Fourier multipliers

The Fourier multiplier given by a function $a$ is

$$
\begin{equation*}
\left(a\left(k^{-1} D\right) v\right)(x):=\mathcal{F}_{k}^{-1}\left(a(\cdot)\left(\mathcal{F}_{k} v\right)(\cdot)\right)(x) \tag{6.5}
\end{equation*}
$$

i.e. we multiply the Fourier transform of $v$ by $a$, and then apply the inverse Fourier transform. The rationale for the notation $a\left(k^{-1} D\right)$ is that, by (6.2), if $D:=-\mathrm{i} \partial$ then $\mathcal{F}_{k}$ maps $k^{-1} D$ to $\xi$. Our motivation for studying Fourier multipliers is that (i) the operator $-k^{-2} \Delta-1$ is one: by the derivative rule (6.2), $-k^{-2} \Delta-1=p\left(k^{-1} D\right)$ where $p(\xi):=|\xi|^{2}-1$, and (ii) the functions $u_{H^{2}}$ and $u_{\mathcal{A}}$ in our proof of Theorem 5.1 are defined by Fourier multipliers acting on $u$ (see (7.3) below).

We say that $a$ is a Fourier symbol of order $m$ if there exists $C>0$ such that

$$
\begin{equation*}
|a(\xi)| \leq C\langle\xi\rangle^{m} \quad \text { for all } \xi \in \mathbb{R}^{d}, \tag{6.6}
\end{equation*}
$$

where recall that $\langle\xi\rangle:=\left(1+|\xi|^{2}\right)^{1 / 2}$. We use the (non-standard) notation that $a \in$ $(F S)^{m}$.

Example 6.1 (Examples of Fourier symbols and multipliers)
(i) If $a(\xi)=1$, then $a \in(F S)^{0}$ with $a\left(k^{-1} D\right) v(x)=v(x)$ (since $\left.\mathcal{F}_{k}^{-1} \mathcal{F}_{k}=I\right)$.
(ii) If $p(\xi):=|\xi|^{2}-1$ then $p \in(F S)^{2}$ with $p\left(k^{-1} D\right) v=\left(-k^{-2} \Delta-1\right) v$.
(iii) If $\chi$ is bounded and has compact support, then $\chi \in(F S)^{-N}$ for all $N \geq 1$ and $1-\chi \in(F S)^{0}$.

Theorem 6.2 (Composition and mapping properties of Fourier multipliers) If $a \in$ $(F S)^{m_{a}}$ and $b \in(F S)^{m_{b}}$ then the following properties hold.
(i) $a b \in(F S)^{m_{a}+m_{b}}$.
(ii) $\quad a\left(k^{-1} D\right) b\left(k^{-1} D\right)=(a b)\left(k^{-1} D\right)=(b a)\left(k^{-1} D\right)=b\left(k^{-1} D\right) a\left(k^{-1} D\right)$.
(iii) $a\left(k^{-1} D\right): H_{k}^{s}\left(\mathbb{R}^{d}\right) \rightarrow H_{k}^{s-m_{a}}\left(\mathbb{R}^{d}\right)$ and, with $C$ the constant in (6.6), for all $s \in \mathbb{R}$ and $k>0$,

$$
\begin{equation*}
\left\|\mid a\left(k^{-1} D\right)\right\|_{H_{k}^{s}\left(\mathbb{R}^{d}\right) \rightarrow H_{k}^{s-m_{a}}\left(\mathbb{R}^{d}\right)} \leq C ; \tag{6.7}
\end{equation*}
$$

i.e. $a\left(k^{-1} D\right)$ is bounded uniformly in both $k$ and $s$ as an operator from $H_{k}^{s}$ to $H_{k}^{s-m_{a}}$.

Proof Part (i) follows directly from the definition (6.6), part (ii) follows directly from the definition (6.5) and the fact that multiplication in $\mathbb{C}$ is commutative, and part (iii) follows from the definitions of $\|\|\cdot\|\|_{H_{k}^{s}\left(\mathbb{R}^{d}\right)}(6.4)$ and $a\left(k^{-1} D\right) v(6.5)$.

The key result from this section used in the proof of the bound Eq. 5.2 on $u_{H^{2}}$ is the following.

Theorem 6.3 (Factoring an "elliptic" Fourier multiplier out of another) Suppose that $a \in(F S)^{m_{a}}, b \in(F S)^{m_{b}}$, and there exists $c>0$ such that

$$
\begin{equation*}
|b(\xi)| \geq c\langle\xi\rangle^{m_{b}} \quad \text { for } \xi \in \operatorname{supp} a \tag{6.8}
\end{equation*}
$$

Then

$$
\begin{equation*}
a\left(k^{-1} D\right)=q\left(k^{-1} D\right) b\left(k^{-1} D\right) \tag{6.9}
\end{equation*}
$$

where $q \in(F S)^{m_{a}-m_{b}}$ is defined by $q(\xi):=a(\xi) / b(\xi)$.

Proof The fact that $q$ is in $(F S)^{m_{a}-m_{b}}$ follows from the fact that $a \in(F S)^{m_{a}}$ and the bound (6.8). The result (6.9) then follows from part (ii) of Theorem 6.2.

We now combine Theorem 6.3 and the mapping property (6.7) to obtain the following result (which we use in the proof of Corollary 2.3); we highlight that a similar combination of these results in used in Section 7 in the proof of the bound (5.2) on $u_{H^{2}}$.

Corollary 6.4 (Elliptic regularity) If $\left(-k^{-2} \Delta+1\right) v \in L^{2}\left(\mathbb{R}^{d}\right)$ then $v \in H^{2}\left(\mathbb{R}^{d}\right)$ with

$$
\||v|\|_{H_{k}^{2}\left(\mathbb{R}^{d}\right)} \leq\left\|\left(-k^{-2} \Delta+1\right) v\right\|_{L^{2}\left(\mathbb{R}^{d}\right)} .
$$

Proof We apply Theorem 6.3 with $a(\xi)=1$ and $b(\xi)=|\xi|^{2}+1$, so that $a\left(k^{-1} D\right)=$ $I$ and $b\left(k^{-1} D\right)=-k^{-2} \Delta+1$. The theorem implies that $q(\xi):=\langle\xi\rangle^{-2} \in(F S)^{-2}$ (this is also clear from the definition (6.6)). Then, by the mapping property (6.7),

$$
\|\mid v\|_{H_{k}^{2}\left(\mathbb{R}^{d}\right)}=\| \| q\left(k^{-1} D\right)\left(-k^{-2} \Delta+1\right) v\left\|_{H_{k}^{2}\left(\mathbb{R}^{d}\right)} \leq\right\|\left(-k^{-2} \Delta+1\right) v \|_{L^{2}\left(\mathbb{R}^{d}\right)} .
$$

## 7 Proof of Theorem 5.1

### 7.1 Definition of high- and low-frequency cut-offs

Let

$$
\chi_{\lambda}(\xi)=1_{|\xi| \leq \lambda}(\xi)= \begin{cases}1 & \text { for }|\xi| \leq \lambda  \tag{7.1}\\ 0 & \text { for }|\xi|>\lambda\end{cases}
$$

We define the low-frequency cut-off $\Pi_{L}$ by

$$
\begin{equation*}
\Pi_{L}:=\chi_{\lambda}\left(k^{-1} D\right) ; \quad \text { i.e. } \quad \Pi_{L} v=\mathcal{F}_{k}^{-1}\left(\chi_{\lambda}(\cdot)\left(\mathcal{F}_{k} v\right)(\cdot)\right) \tag{7.2}
\end{equation*}
$$

by the definition (6.5) of a Fourier multiplier. We see that $\Pi_{L}$ acting on a function $v$ returns the frequencies of $v$ that are $\leq \lambda$; hence, why we call it a low-frequency cut-off. ${ }^{1}$ We define the high-frequency cut-off $\Pi_{H}$ by

$$
\Pi_{H}:=I-\Pi_{L}=\left(1-\chi_{\lambda}\right)\left(k^{-1} D\right)
$$

i.e. $\Pi_{H}$ acting on a function $v$ returns the frequencies of $v$ that are $\geq \lambda$.

### 7.2 Definition of $u_{H^{2}}$ and $u_{\mathcal{A}}$ via the frequency cut-offs

Let $u$ be as in Theorem 5.1; i.e. $u$ is the outgoing solution of the Helmholtz equation (2.1). Let $\varphi \in C_{\text {comp }}^{\infty}\left(\mathbb{R}^{d},[0,1]\right)$ be equal to one on $B_{R}$ and vanish outside $B_{2 R}$, and set

$$
\begin{equation*}
u_{\mathcal{A}}:=\left.\left(\Pi_{L}(\varphi u)\right)\right|_{B_{R}} \quad \text { and } \quad u_{H^{2}}:=\left.\left(\Pi_{H}(\varphi u)\right)\right|_{B_{R}} \tag{7.3}
\end{equation*}
$$

Since $\Pi_{L}+\Pi_{H}=I$, these definitions imply that, on $B_{R}, u_{\mathcal{A}}+u_{H^{2}}=\varphi u=u$; i.e. (5.1) holds. This splitting contains the arbitrary parameter $\lambda$; we fix this when proving the bound (5.2) on $u_{H^{2}}$.

### 7.3 Proof of the bound (5.3) on $u_{\mathcal{A}}$

The idea of the proof, in short, is that a function with a compactly supported Fourier transform is analytic. Plancherel's theorem (6.3), the derivative property (6.2), and the definition of $\Pi_{L}$ (7.2) imply that

$$
\begin{equation*}
\left\|\left(k^{-1} \partial\right)^{\alpha}\left(\Pi_{L} \varphi u\right)\right\|_{L^{2}\left(\mathbb{R}^{d}\right)}=\left(\frac{k}{2 \pi}\right)^{d / 2}\left\|(\cdot)^{\alpha} \chi_{\lambda}(\cdot) \mathcal{F}_{k}(\varphi u)(\cdot)\right\|_{L^{2}\left(\mathbb{R}^{d}\right)} \tag{7.4}
\end{equation*}
$$

[^1]The definition of $\chi_{\lambda}$ (7.1) implies that $\left|\xi^{\alpha} \chi_{\lambda}(\xi)\right| \leq \lambda^{|\alpha|}$ for all $\xi \in \mathbb{R}^{d}$. Using this fact in (7.4) and then (in this order) Plancherel's theorem (6.3), the fact that $\varphi=0$ outside $B_{2 R}$ and $\varphi \leq 1$ inside $B_{2 R}$, and the definition of $C_{\text {sol }}$ (2.6), we find that

$$
\begin{align*}
\left\|\left(k^{-1} \partial\right)^{\alpha}\left(\Pi_{L} \varphi u\right)\right\|_{L^{2}\left(\mathbb{R}^{d}\right)} & \leq\left(\frac{k}{2 \pi}\right)^{d / 2} \lambda^{|\alpha|}\left\|\mathcal{F}_{k}(\varphi u)\right\|_{L^{2}\left(\mathbb{R}^{d}\right)} \leq \lambda^{|\alpha|}\|\varphi u\|_{L^{2}\left(\mathbb{R}^{d}\right)} \\
& \leq \lambda^{|\alpha|}\|u\|_{L^{2}\left(B_{2 R}\right)} \leq \lambda^{|\alpha|} C_{\mathrm{sol}}(k, 2 R)\|f\|_{L^{2}\left(B_{R}\right)} . \tag{7.5}
\end{align*}
$$

By the definition (7.3) of $u_{\mathcal{A}}$,

$$
\left\|\left(k^{-1} \partial\right)^{\alpha} u_{\mathcal{A}}\right\|_{L^{2}\left(B_{R}\right)}=\left\|\left(k^{-1} \partial\right)^{\alpha}\left(\Pi_{L} \varphi u\right)\right\|_{L^{2}\left(B_{R}\right)} \leq\left\|\left(k^{-1} \partial\right)^{\alpha}\left(\Pi_{L} \varphi u\right)\right\|_{L^{2}\left(\mathbb{R}^{d}\right)} ;
$$

the bound (5.3) then follows from (7.5) with $C_{\text {split, } \mathcal{A}}:=\lambda$. We make two remarks.

- The fact that $\Pi_{L} \varphi u \in H_{k}^{s}\left(\mathbb{R}^{d}\right)$ for all $s$ follows from the fact that $\chi_{\lambda} \in(F S)^{-N}$ for every $N>0$ (from Part (iii) of Example 6.1) and the mapping property in Part (iii) of Theorem 6.2; we give the direct proof above, however, to have explicit control on the constants in the bound (to show that $\Pi_{L} \varphi u$ is actually analytic).
- The fact that $u_{\mathcal{A}}$ comes from a function with a compactly supported Fourier transform, and hence automatically is analytic, is one of the advantages of the current splitting compared to the original splitting in [34] (see the discussion in Section 7.5).


### 7.4 Proof of the bound (5.2) on $u_{H^{2}}$

The idea of the proof is to use Theorem 6.3, using the fact that the Helmholtz operator is an elliptic Fourier multiplier (in the sense of (6.8)) on high frequencies, and thus, if $\lambda$ is sufficiently large, on the support of the high-frequency cut-off $\Pi_{H}$. By the definition of $u_{H^{2}}$ (7.3) and the equivalence of $\|\cdot\|_{H_{k}^{2}}$ and $\mid\|\cdot\| \|_{H_{k}^{2}}$ described in Section 6.2,

$$
\left\|u_{H^{2}}\right\|_{H_{k}^{2}\left(B_{R}\right)}=\left\|\Pi_{H}(\varphi u)\right\|_{H_{k}^{2}\left(B_{R}\right)} \leq\left\|\Pi_{H}(\varphi u)\right\|_{H_{k}^{2}\left(\mathbb{R}^{d}\right)} \lesssim\| \| \Pi_{H}(\varphi u) \mid \|_{H_{k}^{2}\left(\mathbb{R}^{d}\right)},
$$

where we use the notation that $A \lesssim B$ if there exists $C>0$, independent of $k$ and $R$, such that $A \leq C B$. It is therefore sufficient to prove that

$$
\begin{equation*}
\left\|\left\|\Pi_{H}(\varphi u)\right\|\right\|_{H_{k}^{2}\left(\mathbb{R}^{d}\right)} \lesssim\|f\|_{L^{2}\left(B_{R}\right)} \quad \text { for all } k \geq k_{0} \tag{7.6}
\end{equation*}
$$

Lemma 7.1 ("Ellipticity" of $p(\xi):=|\xi|^{2}-1$ when $|\xi| \geq \lambda>1$ ) If $\lambda \geq \lambda_{0}>1$ then there exists $C>0$ such that

$$
\begin{equation*}
\left||\xi|^{2}-1\right| \geq C\langle\xi\rangle^{2} \quad \text { for }|\xi| \geq \lambda \tag{7.7}
\end{equation*}
$$

Proof It is straightforward to check that (7.7) holds with $C:=\left(1+2\left(\lambda_{0}^{2}-1\right)^{-1}\right)^{-1}$.

Corollary 7.2 (Factoring out $p\left(k^{-1} D\right)$ from $\left(1-\chi_{\lambda}\right)\left(k^{-1} D\right)$ ) Let $p(\xi):=|\xi|^{2}-1$. If $\lambda \geq \lambda_{0}>1$, then there exists $q \in(F S)^{-2}$ such that

$$
\left(1-\chi_{\lambda}\right)\left(k^{-1} D\right)=q\left(k^{-1} D\right) p\left(k^{-1} D\right)
$$

Proof By the definition (7.1) of $\chi_{\lambda}, \operatorname{supp}\left(1-\chi_{\lambda}\right)=\{\xi:|\xi| \geq \lambda\}$. The result follows by applying Theorem 6.3 with $a(\xi):=\left(1-\chi_{\lambda}\right)(\xi)$ and $b(\xi):=p(\xi)=|\xi|^{2}-1$, since (7.7) implies that the inequality (6.8) holds (i.e. $b$ is elliptic on $\operatorname{supp} a$ ).

We now use Corollary 7.2 and the mapping property (6.7) to prove the bound (7.6). By the definition of $u_{H^{2}}$ (7.3) and Corollary 7.2,

$$
\left\|\left\|\Pi_{H}(\varphi u)\right\|_{H_{k}^{2}}=\mid\right\|\left(1-\chi_{\mu}\right)\left(k^{-1} D\right)(\varphi u)\left\|_{H_{k}^{2}}=\right\|\left\|q\left(k^{-1} D\right) p\left(k^{-1} D\right)(\varphi u)\right\|_{H_{k}^{2}}
$$

with $q \in(F S)^{-2}$, and, by the mapping property (6.7),

$$
\left\|q\left(k^{-1} D\right) p\left(k^{-1} D\right)(\varphi u)\right\|_{H_{k}^{2}\left(\mathbb{R}^{d}\right)} \lesssim\left\|p\left(k^{-1} D\right)(\varphi u)\right\|_{L^{2}\left(\mathbb{R}^{d}\right)} ;
$$

the key point is that we now have the Helmholtz operator $p\left(k^{-1} D\right)$ on the righthand side, and we can start to relate this side to $f=p\left(k^{-1} D\right) u$. For brevity, let $P:=-k^{-2} \Delta-1=p\left(k^{-1} D\right)$. Then

$$
\begin{align*}
\left\|\Pi_{H}(\varphi u)\right\|_{H_{k}^{2}\left(\mathbb{R}^{d}\right)} \lesssim\|P(\varphi u)\|_{L^{2}\left(\mathbb{R}^{d}\right)} & =\|\varphi P u+[P, \varphi] u\|_{L^{2}\left(\mathbb{R}^{d}\right)} \\
& \leq\|f\|_{L^{2}\left(\mathbb{R}^{d}\right)}+\|[P, \varphi] u\|_{L^{2}\left(\mathbb{R}^{d}\right)} \tag{7.8}
\end{align*}
$$

where we have used the fact that $\varphi \equiv 1$ on supp $f$, and where the commutator $[A, B]$ is defined as $A B-B A$. By direct calculation,

$$
[P, \varphi] u=-k^{-2}(u \Delta \varphi+2 \nabla \varphi \cdot \nabla u)
$$

so that

$$
\begin{equation*}
\|[P, \varphi] u\|_{L^{2}\left(\mathbb{R}^{d}\right)} \lesssim(k R)^{-1}\|u\|_{H_{k}^{1}\left(B_{2 R}\right)} \tag{7.9}
\end{equation*}
$$

where we have used that the definition of $\varphi$ in Section 7.2 implies that $|\nabla \varphi| \sim R^{-1}$ and $|\Delta \varphi| \sim R^{-2}$. Therefore, by combining (7.8) and (7.9), and using (2.8) (with $R$ replaced by $2 R$ ), we have

$$
\left\|\Pi_{H}(\varphi u)\right\|_{H_{k}^{2}\left(\mathbb{R}^{d}\right)} \lesssim\|f\|_{L^{2}\left(\mathbb{R}^{d}\right)}+(k R)^{-1}\|u\|_{H_{k}^{1}\left(B_{2 R}\right)} \lesssim\|f\|_{L^{2}\left(\mathbb{R}^{d}\right)}
$$

which is the result (7.6).

### 7.5 Discussion of the original proof of Theorem 5.1 in [34]

The original proof of Theorem 5.1 in [34, Section 3.2] is also based on the idea of frequency cut-offs using the indicator function (7.1). However, in [34, Section 3.2], the frequency cut-offs are applied to the data $f$ (as opposed to $\varphi u$ in our case). Furthermore, the analysis in [34, Section 3.2] is based on writing the solution $u$ to the Helmholtz eqaution (2.1) satisfying the Sommerfeld radiation condition (2.2) as

$$
\begin{equation*}
u(x)=k^{2} \int_{B_{R}} \Phi_{k}(x, y) f(y) \mathrm{d} y \tag{7.10}
\end{equation*}
$$

where $\Phi_{k}(x, y)$ is the fundamental solution (satisfying $\left(\Delta_{y}+k^{2}\right) \Phi_{k}(x, y)=-\delta(x-$ $y)$ ) defined by

$$
\Phi_{k}(x, y):=\frac{\mathrm{i}}{4} H_{0}^{(1)}(k|x-y|), \quad d=2, \quad:=\frac{\mathrm{e}^{\mathrm{i} k|x-y|}}{4 \pi|x-y|}, \quad d=3 .
$$

The proof in [34, Section 3.2] then considers the function

$$
v_{\mu}(x):=k^{2} \int_{B_{R}} \Phi_{k}(x, y) \mu(|x-y|) f(y) \mathrm{d} y
$$

where $\mu \in C_{\text {comp }}^{\infty}(\mathbb{R})$ with $\left.\mu\right|_{[0,2 R]}=1$, supp $\mu \subset[0,4 R]$, and additionally the first and second derivatives of $\mu$ satisfying certain bounds (see [34, Equation 3.27]). This definition of $\mu$ implies that $\left.v_{\mu}\right|_{B_{R}}=\left.u\right|_{B_{R}}$. The advantage of studying $v_{\mu}$ instead of $u$ is that $v$ is the convolution of a compactly-supported kernel with $f$; thus, $v$ has compact support and its Fourier transform is well defined. (In contrast, in the present paper, we first multiply $u$ by $\varphi$ to ensure that $\mathcal{F}_{k}(\varphi u)$ makes sense.) The analysis in [34, Section 3.2 and Appendix A] then proceeds by studying the compactly supported integral kernel in $v_{\mu}$ and the Fourier transform of this kernel, with identities and bounds on Bessel and Hankel functions needed when both $d=2$ and $d=3$.

### 7.6 Generalising Theorem 5.1 to more-complicated Helmholtz problems

The proof of Theorem 5.1 above can be generalised to more-complicated Helmholtz problems, involving variable coefficients and obstacles, using pseudodifferential operators. Indeed, whereas the operator $-k^{-2} \Delta-1$ is a Fourier multiplier (by Part (ii) of Example 6.1), its variable coefficient analogue $-k^{-2} \nabla \cdot(A \nabla)-n$ is not; i.e. one cannot write the Fourier transform of $-k^{-2} \nabla \cdot(A \nabla v)-n v$ in terms of the Fourier transform of $v$. Nevertheless, [32] proves Theorem 5.1, using the same ideas in the proof above, with the Helmholtz equation (2.1) replaced by

$$
\begin{equation*}
k^{-2} \nabla \cdot(A \nabla u)+n u=-f \quad \text { in } \mathbb{R}^{d} \tag{7.11}
\end{equation*}
$$

This is achieved by replacing the Fourier multipliers in Section 6 by pseudodifferential operators (indeed, recall that one of the motivations for the development of the theory of pseudodifferential operators in the 1960s was to study variable-coefficient PDEs such as (7.11) using Fourier analysis). In particular, the pseudodifferential generalisation of the key result of Theorem 6.3 is the so-called elliptic parametrix (see, e.g. [15, Proposition E.32]).

These ideas can also be used to prove an analogue of Theorem 5.1 when the Helmholtz equation (with or without variable coefficients) is posed not in $\mathbb{R}^{d}$ but outside an impenetrable obstacle. In this case, the Fourier transform (6.1) is no longer available; nevertheless, the functional calculus (essentially the idea of expanding in eigenfunctions of the differential operator) can be used to define Fourier-type transforms, tailored to the particular problem. The ideas of the proof of Theorem 5.1 can then be implemented in this situation (albeit with more technicalities, and the requirement that the obstacle is analytic) (see [21, 22]).

Remark 7.3 (The smoothness of the frequency cut-offs) The theory of pseudodifferential operators is easiest when the symbols (i.e. the generalisation of the Fourier symbols in (6.6)) are $C^{\infty}$. To ensure this, [32] assumes that the coefficients $A$ and $n$ in (7.11) are $C^{\infty}$, and uses a smooth frequency cut-off; i.e. the indicator function in (7.1) is replaced by $\chi_{\lambda} \in C_{\text {comp }}^{\infty}\left(\mathbb{R}^{d} ;[0,1]\right)$ such that $\chi_{\lambda}(\xi)=1$ for $|\xi| \leq \lambda$ and $\chi_{\lambda}(\xi)=0$ for, say, $|\xi| \geq 2 \lambda$. The reader can check that the proof in Section 7 goes through as before with this smooth cut-off (with $C_{\text {split, } \mathcal{A}}$ in Section 7.3 changed from $\lambda$ to $2 \lambda$ ).

## Appendix A. Proof of Theorem 2.2

Lemma A. 1 (Morawetz identity for the Helmholtz operator [43, Section I.2]) If

$$
\begin{equation*}
\mathcal{L} v:=k^{-2} \Delta v+v \quad \text { and } \quad \mathcal{M}_{\beta, \alpha} v:=x \cdot \nabla v-\mathrm{i} k \beta v+\alpha v, \tag{A.1}
\end{equation*}
$$

with $\beta$ and $\alpha$ real-valued $C^{1}$ functions, then

$$
\begin{align*}
& 2 \mathfrak{R}\left(\overline{\mathcal{M}_{\beta, \alpha} v} \mathcal{L} v\right)=\nabla \cdot\left[2 k^{-1} \mathfrak{R}\left(\overline{\mathcal{M}_{\beta, \alpha} v} k^{-1} \nabla v\right)+\left(|v|^{2}-k^{-2}|\nabla v|^{2}\right) x\right] \\
& -2 \Re\left(\bar{v}\left(\mathrm{i} \nabla \beta+k^{-1} \nabla \alpha\right) \cdot k^{-1} \nabla v\right)-(d-2 \alpha)|v|^{2}-(2 \alpha-d+2) k^{-2}|\nabla v|^{2} . \tag{A.2}
\end{align*}
$$

Proof This follows in a straightforward (but slightly involved) way by expanding the divergence on the right-hand side; for this done step-by-step, see, e.g. [51, Proof of Lemma 2.1].

The idea of the proof of Theorem 2.2 is to integrate the identity (A.2) over $\mathbb{R}^{d}$ with $v=u, \alpha=(d-1) / 2$, and $\beta$ defined piecewise as $\beta=R$ for $r \leq R$ and $\beta=r$ for $r \geq R$. The choice $\beta=$ constant and $\alpha=(d-1) / 2$ means that the nondivergence terms on the right-hand side of (A.2) become $-|u|^{2}-k^{-2}|\nabla u|^{2}$; this is where we get $\|u\|_{H_{k}^{1}\left(B_{R}\right)}^{2}$ from. The choice $\beta=r$ deals with the contribution from infinity (although this is not immediately clear from (A.2)). We therefore first look at the special case of (A.2) with $\beta=r$.

Lemma A. 2 (Special case of (A.2) [42, Equation 1.2]) With $\mathcal{L} v$ and $\mathcal{M}_{\beta, \alpha} v$ as in Lemma A.l, show that if $\alpha \in \mathbb{R}$, then, with $v_{r}=x \cdot \nabla v / r$,

$$
\begin{align*}
2 \mathfrak{R}\left(\overline{\mathcal{M}_{r, \alpha} v} \mathcal{L} v\right)= & \nabla \cdot\left[2 k^{-1} \mathfrak{R}\left(\overline{\mathcal{M}_{r, \alpha} v} k^{-1} \nabla v\right)+\left(|v|^{2}-k^{-2}|\nabla v|^{2}\right) x\right]-\left|k^{-1} v_{r}-\mathrm{i} v\right|^{2} \\
& +(2 \alpha-(d-1))\left(|v|^{2}-k^{-2}|\nabla v|^{2}\right)-k^{-2}\left(|\nabla v|^{2}-\left|v_{r}\right|^{2}\right) . \tag{A.3}
\end{align*}
$$

Proof This follows from (A.2) by choosing $\beta=r$ and writing the term involving $\nabla \beta$ as

$$
-2 \Re\left(\mathrm{i} \bar{v} k^{-1} v_{r}\right)=-2 \mathfrak{R}\left(\overline{(-\mathrm{i} v)} k^{-1} v_{r}\right)=|v|^{2}+k^{-2}\left|v_{r}\right|^{2}-\left|k^{-1} v_{r}-\mathrm{i} v\right|^{2}
$$

using the identity $-2 \mathfrak{R}\left(z_{1} \overline{z_{2}}\right)=\left|z_{1}\right|^{2}+\left|z_{2}\right|^{2}-\left|z_{1}+z_{2}\right|^{2}$.

To integrate (A.3) over $\mathbb{R}^{d} \backslash B_{R}$, we integrate it over $B_{R_{1}} \backslash B_{R}$ and then send $R_{1} \rightarrow \infty$. In preparation for this, we look at the boundary term on $\partial B_{R_{1}}$.

Lemma A. 3 Let

$$
\begin{equation*}
Q_{r, \alpha}(v):=2 k^{-1} \mathfrak{R}\left(\overline{\mathcal{M}_{r, \alpha} v} k^{-1} \nabla v\right)+\left(|v|^{2}-k^{-2}|\nabla v|^{2}\right) x . \tag{A.4}
\end{equation*}
$$

If $u$ is an outgoing solution of $\mathcal{L} u=0$ in $\mathbb{R}^{d} \backslash \overline{B_{R_{0}}}$, then, for all $\alpha \in \mathbb{R}$,

$$
\begin{equation*}
\int_{\Gamma_{\partial B_{R_{1}}}} Q_{R_{1}, \alpha}(u) \cdot \widehat{x} \rightarrow 0 \quad \text { as } R_{1} \rightarrow \infty \tag{A.5}
\end{equation*}
$$

The proof of Lemma A. 3 requires the following classic result.
Theorem A. 4 (Atkinson-Wilcox expansion (see, e.g. [13, Theorem 3.7])) If $u \in$ $H_{l o c}^{1}\left(\mathbb{R}^{d} \backslash \overline{B_{R_{0}}}\right)$ is an outgoing solution of $k^{-2} \Delta u+u=0$ in $\mathbb{R}^{d} \backslash \overline{B_{R_{0}}}$ for some $R_{0}>0$, then there exist smooth functions $F_{n}$ such that, for any $R_{1}>R_{0}$,

$$
\begin{equation*}
u(x)=\frac{\mathrm{e}^{\mathrm{i} k r}}{r^{(d-1) / 2}} \sum_{n=0}^{\infty} \frac{F_{n}(x / r)}{r^{n}} \quad \text { for } r:=|x| \geq R_{1} \tag{A.6}
\end{equation*}
$$

where the sum in (A.6) (and all its derivatives) converges absolutely and uniformly.

Proof of Lemma A. 3 By the definitions of $Q_{r, \alpha}(v)(A .4)$ and $\mathcal{M}_{r, \alpha} v$ (A.1),

$$
\begin{align*}
Q_{r, \alpha}(u) \cdot \widehat{x} & =r k^{-2}\left(2 \mathfrak{R}\left(u_{r} \overline{\left(u_{r}-\mathrm{i} k u+\frac{\alpha}{r} u\right)}\right)+k^{2}|u|^{2}-|\nabla u|^{2}\right) \\
& =r k^{-2}\left(\left|u_{r}\right|^{2}+2 \mathfrak{R}\left(u_{r} \overline{\left(-\mathrm{i} k u+\frac{\alpha}{r} u\right)}\right)+k^{2}|u|^{2}-\left(|\nabla u|^{2}-\left|u_{r}\right|^{2}\right)\right) \\
& =r k^{-2}\left(\frac{\left|\mathcal{M}_{r, \alpha} u\right|^{2}}{r^{2}}+\alpha^{2} \frac{|u|^{2}}{r^{2}}-\left(|\nabla u|^{2}-\left|u_{r}\right|^{2}\right)\right), \tag{A.7}
\end{align*}
$$

where we have again used the identity $2 \Re\left(z_{1} \overline{z_{2}}\right)=\left|z_{1}+z_{2}\right|^{2}-\left|z_{1}\right|^{2}-\left|z_{2}\right|^{2}$. We now claim that the term in large brackets in (A.7) is $O\left(r^{-d-1}\right)$; if this is true, then

$$
\int_{\Gamma_{\partial B_{R_{1}}}} Q_{R_{1}, \alpha}(u) \cdot \widehat{x}=O\left(\frac{1}{R_{1}}\right) \text { as } R_{1} \rightarrow \infty,
$$

and thus (A.5) follows. By the Atkinson-Wilcox expansion (A.6), $|u|^{2}=O\left(r^{1-d}\right)$ and $r^{-2}\left|\mathcal{M}_{r, \alpha} u\right|^{2}=O\left(r^{-d-1}\right)$. To prove the result, therefore, we only need to show that $|\nabla u|^{2}-\left|u_{r}\right|^{2}=O\left(r^{-d-1}\right)$. The quantity $|\nabla u|^{2}-\left|u_{r}\right|^{2}$ equals $\left|\nabla_{S} u\right|^{2}$ where $\nabla_{S}$ is the surface gradient on $|x|=r$, which satisfies $\nabla_{S} u=\nabla u-\widehat{x} u_{r}$. This differential operator is equal to $1 / r$ multiplied by an operator acting only on $\widehat{x}$, i.e. the angular variables; thus, $\left|\nabla_{S} u\right|^{2}$ is $O\left(r^{-d-1}\right)$ and the proof is complete.

We now integrate (A.3) over $B_{R_{1}} \backslash B_{R}$, send $R_{1} \rightarrow \infty$, and obtain an inequality involving the boundary term on $\partial B_{R}$.

Lemma A. 5 If $u$ is an outgoing solution of $\mathcal{L} u=0$ in $\mathbb{R}^{d} \backslash \overline{B_{R_{0}}}$, for some $R_{0}>0$, then, for $R>R_{0}$,

$$
\begin{equation*}
\int_{\partial B_{R}} Q_{R,(d-1) / 2}(u) \cdot \widehat{x} \leq 0 . \tag{A.8}
\end{equation*}
$$

Proof We integrate the identity (A.3) over $B_{R_{1}} \backslash B_{R}$, where $R_{1}>R$, with $v=u$ and then use the divergence theorem $\int_{D} \nabla \cdot F=\int_{\partial D} F$. The divergence theorem is valid for $F \in C^{\infty}(\bar{D})$ and $D$ Lipschitz (see, e.g. [41, Theorem 3.34]); we can use it here since, by elliptic regularity, $u \in C^{\infty}\left(\overline{B_{R_{1}} \backslash B_{R}}\right)$ (see, e.g. [41, Theorem 4.16]). This results in

$$
\begin{aligned}
\int_{\partial B_{R_{1}}} Q_{R_{1}, \alpha}(u) \cdot \widehat{x}-\int_{\partial B_{R}} Q_{R, \alpha}(u) \cdot \widehat{x}= & \int_{B_{R_{1} \backslash B_{R}}}-(2 \alpha-(d-1))\left(|v|^{2}-k^{-2}|\nabla v|^{2}\right) \\
& +k^{-2}\left(|\nabla v|^{2}-\left|v_{r}\right|^{2}\right)+\left|k^{-1} v_{r}-\mathrm{i} v\right|^{2} .
\end{aligned}
$$

Setting $\alpha=(d-1) / 2$ eliminates the first term on the right-hand side of (A.9). Since $\left|v_{r}\right| \leq|\nabla v|$, the remaining terms on the right-hand side of (A.9) are non-negative, and thus

$$
\int_{\partial B_{R_{1}}} Q_{R_{1},(d-1) / 2}(u) \cdot \widehat{x}-\int_{\partial B_{R}} Q_{R,(d-1) / 2}(u) \cdot \widehat{x} \geq 0 .
$$

Sending $R_{1} \rightarrow \infty$ and using (A.5), we obtain the result (A.8).

Proof of Theorem 2.2 The plan is to integrate the identity (A.2) over $B_{R}$ with $v=u$, $\beta=R$, and $\alpha=(d-1) / 2$, and then use the divergence theorem. We justify using the divergence theorem just as we did at the beginning of the proof of Lemma A. 3 to find that if $v \in C^{\infty}(\bar{D})$ then

$$
\begin{align*}
\int_{B_{R}} 2 \mathfrak{R}\left(\overline{\mathcal{M}_{R,(d-1) / 2} v} \mathcal{L} v\right)= & \int_{\partial B_{R}} 2 k^{-1} \mathfrak{\Re}\left(\overline{\mathcal{M}_{R,(d-1) / 2} v} k^{-1} \nabla v\right)+\left(|v|^{2}-k^{-2}|\nabla v|^{2}\right) x \\
& -\int_{B_{R}}\left(|v|^{2}+k^{-2}|\nabla v|^{2}\right) . \tag{A.10}
\end{align*}
$$

We now claim that (A.10) holds for $v \in H^{2}\left(B_{R}\right)$; this follows since $C^{\infty}(\bar{D})$ is dense in $H^{2}(D)$ [41, Page 77], and, by the trace theorem (see, e.g. [41, Theorem 3.37]), (A.10) is continuous in $v$ with respect to the topology of $H^{2}\left(B_{R}\right)$. By Corollary 2.3, $u \in H^{2}\left(B_{R}\right)$, and thus (A.10) holds with $v=u$. Using in (A.10) the definition of $Q$ (A.4), the fact that $u$ is an outgoing solution of the Helmholtz equation (2.1), and Lemma A.5, we find that

$$
-\int_{B_{R}} 2 \mathfrak{R}\left(\overline{\mathcal{M}_{R,(d-1) / 2} u} f\right)+\|u\|_{H_{k}^{1}\left(B_{R}\right)}^{2}=\int_{\partial B_{R}} Q_{R,(d-1) / 2}(u) \cdot \widehat{x} \leq 0
$$

Thus

$$
\begin{equation*}
\|u\|_{H_{k}^{1}\left(B_{R}\right)}^{2} \leq 2\left\|\mathcal{M}_{R,(d-1) / 2} u\right\|_{L^{2}\left(B_{R}\right)}\|f\|_{L^{2}\left(B_{R}\right)} \tag{A.11}
\end{equation*}
$$

By the inequality $|a+b|^{2} \leq 2|a|^{2}+2|b|^{2}$, the fact that $|\mathrm{i} k r+\alpha|^{2}=k^{2} r^{2}+\alpha^{2}$, and the bound $r \leq R$ on $B_{R}$,

$$
\begin{aligned}
\left\|\mathcal{M}_{R,(d-1) / 2} u\right\|_{L^{2}\left(B_{R}\right)}^{2} & \leq 2 R^{2}\|\nabla u\|_{L^{2}\left(B_{R}\right)}^{2}+2\left((k R)^{2}+\alpha^{2}\right)\|u\|_{L^{2}\left(B_{R}\right)}^{2} \\
& \leq 2 k^{2} R^{2}\left(1+\frac{\alpha^{2}}{k^{2} R^{2}}\right)\|u\|_{H_{k}^{1}\left(B_{R}\right)}^{2}
\end{aligned}
$$

Using this in (A.11), and recalling that $\alpha=(d-1) / 2$, we find that $\|u\|_{H_{k}^{1}\left(B_{R}\right)} \leq$ $C\|f\|_{L^{2}\left(B_{R}\right)}$ with $C$ the right-hand side of (2.7).

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## Declarations

Conflict of interest The author declares no competing interests.

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[^1]:    ${ }^{1}$ The definition of $\mathcal{F}_{k}(6.1)$ implies that if $\mathcal{F}_{k} w$ is supported on $|\xi| \leq \lambda$, then the "standard" Fourier transform (i.e. with the transform variable not scaled by $k$ ) of $w$ is supported for $|\zeta| \leq \lambda k$.

