# Frames and Space Splittings in Hilbert Spaces 

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#### Abstract

This is the first part of lecture notes based on short courses held by the author at the Universities of Bonn and Lancaster in 1997. We review the abstract theory of frames and their generalization - the so-called stable space splittings - in a Hilbert space setting. While the frame concept was developed as part of (non-harmonic) Fourier analysis and mainly in connection with signal processing applications, the latter theory of stable subspace splitting has led to a better understanding of iterative solvers (multigrid/multilevel resp. domain decomposition methods) for large-scale discretizations of elliptic operator equations.


## 1 Hilbert space notation

For the basics of Hilbert space terminology, you may consult any of your favorite textbooks, additional sources $[1,2,3,4,5]$ can be found in the reference list. Throughout the paper, $V$ will denote a separable or finite-dimensional real Hilbert space (the occasional appearance of examples of complex Hilbert spaces or the use of results valid for complex Hilbert spaces in the real case via complexification will not be commented on). By default, $(f, g) \equiv(f, g)_{V}$ denotes the scalar product of two elements $f, g \in V$, and $\|f\| \equiv\|f\|_{V}=\sqrt{(f, f)_{V}}$ the norm of $f \in V$. Since often one and the same linear space $V$ will be equipped with several scalar products, we will occasionally use the notation $V=\{V ;(\cdot, \cdot)\}$ to indicate the specific scalar product. Recall that knowing the norm on a Hilbert space $V$ is equivalent to knowing the scalar product since

$$
4(f, g)=\|f+g\|^{2}-\|f-g\|^{2} .
$$

A subset $V_{1} \subset V$ is called subspace of $V$ if it is a linear manifold in $V$ (i.e., $f, g \in V_{1}$ and $\lambda \in \mathbb{R}$ always imply $f+g, \lambda f \in V_{1}$ ) and closed with respect to the Hilbert space
topology. Thus, any subspace equipped with a scalar product obtained by restriction from $V$ turns into a Hilbert space on its own. The orthogonal projection $P_{V_{1}}: V \rightarrow V_{1}$ is then well-defined by

$$
\left\|u-P_{V_{1}} u\right\|=\inf _{v_{1} \in V_{1}}\left\|u-v_{1}\right\|
$$

(i.e., $u_{1}=P_{V_{1}} u$ is the unique element in $V_{1}$ of minimal distance to $u$ ) or, equivalently, by the orthogonality requirement

$$
\left(u-P_{V_{1}} u, v_{1}\right)=0 \quad \forall v_{1} \in V_{1}
$$

The set

$$
V_{1}^{\perp}=\left\{u \in V:\left(u, v_{1}\right)=0 \quad \forall v_{1} \in V_{1}\right\}=\operatorname{ran}\left(\operatorname{Id}-P_{V_{1}}\right) .
$$

is a subspace in $V$, the orthogonal complement of $V_{1}$. Clearly, $P_{V_{1}^{\perp}}=\mathrm{Id}-P_{V_{1}}$, where Id denotes the identity operator.

For an arbitrary subset $F \subset V$ we define its span (denoted by $[F]$ ) as the smallest subspace in $V$ containing $F$. Equivalently,

$$
[F]=\operatorname{clos}_{V}\left\{\sum_{i=1}^{n} c_{i} f_{i}: f_{i} \in F, c_{i} \in \mathbb{R}, n \geq 1\right\}
$$

While the intersection $V_{1} \cap V_{2}$ of two subspaces $V_{1}, V_{2}$ of $V$ is automatically a subspace of $V$, their (closed) sum needs to be defined as

$$
V_{1}+V_{2}=\operatorname{clos}_{V}\left\{f_{1}+f_{2}: f_{1} \in V_{1}, f_{2} \in V_{2}\right\}=\left[V_{1} \cup V_{2}\right]
$$

the latter definition applies also to the case of arbitrarily many subspaces. The closure with respect to $V$ can be dropped only in special cases (e.g., if all but one of the subspaces $V_{j}$ in a sum are finite-dimensional). Counterexamples can be found in [1, p.110] or [2, p. 26/27], [3, p. 39/40]. Another case of interest is when the subspaces are mutually orthogonal, in this case we write $V_{1} \oplus V_{2} \oplus \ldots$ instead of $V_{1}+V_{2}+\ldots$. The closure operation can be dropped if the orthogonal sum has finitely many terms.

There are many more generic constructions of interest for Hilbert spaces. We note the Hilbert sum [4] (which is also called product space [1]) of a family of Hilbert spaces $\left\{V_{j} ;(\cdot, \cdot)_{j}\right\}, j \in I$. It is defined as the Hilbert space $\tilde{V}=\times_{I} V_{j}$ of all sequences $\tilde{v}=\left(v_{j}\right)$ where $v_{j} \in V_{j}, j \in I$, such that

$$
\|\tilde{v}\|_{\tilde{V}}^{2} \equiv \sum_{j \in I}\left\|v_{j}\right\|_{j}^{2}<\infty
$$

Applications will be considered in Section 3 and 4. Another possibility [5, Section 3.4] is the tensor product of Hilbert spaces. A formal way to produce $V_{1} \otimes V_{2}$ is to take complete orthonormal systems (CONS) $\left\{e_{i}^{1}\right\}$ and $\left\{e_{j}^{2}\right\}$ in $V_{1}$ and $V_{2}$, respectively, and to define

$$
V_{1} \otimes V_{2}=\left\{f=\sum_{i, j} c_{i j} e_{i}^{1} \otimes e_{j}^{2}:\|f\|_{V_{1} \otimes V_{2}}^{2}=\sum_{i, j} c_{i j}^{2}\right\}
$$

Partly, the motivation for this construction comes from considering the special case of Hilbert spaces $V_{k}$ of functions defined on domains $\Omega_{k}$. Then, by identifying $e_{i}^{1}(\cdot) \otimes e_{j}^{2}(\cdot)$ with the usual product $e_{i}^{1}\left(x_{1}\right) e_{j}^{2}\left(x_{2}\right)$, the resulting space $V_{1} \otimes V_{2}$ can be interpreted as space of functions on the product domain $\Omega_{1} \times \Omega_{2}$. E.g., $L_{2}\left(\Omega_{1} \times \Omega_{2}\right) \cong L_{2}\left(\Omega_{1}\right) \otimes L_{2}\left(\Omega_{2}\right)$ (see [5]).

Let us conclude with the variational setting of operator equations. Let $V^{\prime}$ denote the dual space to the Hilbert space $V$ (i.e., the set of all bounded linear functionals $\Phi: V \rightarrow \mathbb{R}$, equipped with the norm

$$
\left.\|\Phi\|_{V^{\prime}}=\sup _{\|v\|_{V}=1}|\Phi(v)| .\right)
$$

By the Riesz-Fischer theorem, $V^{\prime}$ is also a Hilbert space (which could be identified with $V)$. Set

$$
\langle\Phi, v\rangle_{V^{\prime} \times V}=\Phi(v) \quad \forall \Phi \in V^{\prime}, v \in V
$$

Then any bounded linear operator $A: V \rightarrow V^{\prime}$ generates a bounded bilinear form

$$
\begin{equation*}
a(u, v)=\langle A u, v\rangle_{V^{\prime} \times V}, \quad u, v \in V, \tag{1}
\end{equation*}
$$

vice versa. Boundedness of the bilinear form $a(\cdot, \cdot)$ means the existence of a constant $C_{1}<\infty$ such that

$$
\begin{equation*}
|a(u, v)| \leq C_{1}\|u\|\|v\|, \quad u, v \in V \tag{2}
\end{equation*}
$$

which together with coercivity (i.e., the existence of another constant $C_{2}>0$ such that

$$
\begin{equation*}
\left.a(u, u) \geq C_{2}\|u\|^{2}, \quad u \in V\right) \tag{3}
\end{equation*}
$$

guarantees unique solvability of the following variational problem: Given any $\Phi \in V^{\prime}$, find $u=u_{\Phi} \in V$ such that

$$
\begin{equation*}
a(u, v)=\Phi(v) \quad \forall v \in V . \tag{4}
\end{equation*}
$$

This follows from the Lax-Milgram theorem. In other words, the operator equation $A u=\Phi$ has a unique solution in $V$, for any $\Phi \in V^{\prime}$, i.e., $A^{-1}: V^{\prime} \rightarrow V$ exists. According to (1-3), we have

$$
\|A\|_{V \rightarrow V^{\prime}} \leq C_{1}, \quad\left\|A^{-1}\right\|_{V^{\prime} \rightarrow V} \leq C_{2}^{-1}
$$

A bilinear form which satisfies both (2) and (3) is called $V$-elliptic, it is symmetric if

$$
a(u, v)=a(v, u) \quad \forall u, v \in V
$$

The optimal constants $C_{1}, C_{2}$ are called ellipticity constants.
As an immediate consequence of these poperties we mention that any $V$-elliptic symmetric bilinear form $a(\cdot, \cdot)$ generates an equivalent scalar product in $V$, i.e.,

$$
\{V ;(\cdot, \cdot)\} \cong\{V ; a(\cdot, \cdot)\}, \quad C_{2}\|u\|^{2} \leq\|u\|_{a}^{2} \equiv a(u, u) \leq C_{1}\|u\|^{2},
$$

for all $u \in V$. Thus, on a theoretical level and up to constants, we can always switch from the energy norm $\|u\|_{a}$ associated with the variational problem (4) resp. with $A$ to the canonical norm $\|u\|$ in $V$ if $a$ is $V$-elliptic and symmetric.

## 2 Frames

The abstract notion of a frame (or, in other words, stable representation system) in a Hilbert space was introduced in [6]. A first survey with emphasis on frames was [8], see also [10, Chapter 3], [11, Chapter 3]. A more recent and comprehensive source is the collection [12] which we recommend for further reading.

For the purpose of this section, we will consider the following definitions. Let $F \equiv$ $\left\{f_{k}\right\} \subset V$ be an at most countable system of elements in $V$.

- We will call $F$ a frame system in $V$ if there are two constants $0<A \leq B<\infty$ such that

$$
\begin{equation*}
A\|f\|^{2} \leq \sum_{k}\left|\left(f, f_{k}\right)\right|^{2} \leq B\|f\|^{2} \quad \forall f \in[F] \tag{5}
\end{equation*}
$$

A frame system $F$ is called frame in $V$ if it is dense, i.e., if $[F]=V$. The optimal constants $A, B$ in (5) are the lower and upper frame bounds, respectively, their ratio $B / A$ defines the condition of $F$ and will be denoted by $\kappa(F)$. A frame (system) $F$ is called tight if $A=B$, i.e., if $\kappa(F)=1$. Finally, a system $F$ is minimal if $f_{k} \notin\left[F \backslash\left\{f_{k}\right\}\right]$ for all $k$, i.e., if the deletion of any $f_{k}$ from the system reduces the span.

- $F$ is a Riesz system in $V$ if there are constants $0<\tilde{A} \leq \tilde{B}<\infty$ such that for all finite linear combinations $f=\sum_{k} c_{k} f_{k}$

$$
\begin{equation*}
\tilde{A}\|f\|^{2} \leq \sum_{k} c_{k}^{2} \leq \tilde{B}\|f\|^{2} \tag{6}
\end{equation*}
$$

If $[F]=V$ then a Riesz system is a Riesz basis (the reader should check that in this case indeed any $f \in V$ possesses a unique decomposition

$$
\begin{equation*}
f=\sum_{k} c_{k} f_{k} \tag{7}
\end{equation*}
$$

which $V$-converges unconditionally (with respect to rearrangements) to f). Again, the optimal constants $\tilde{A}, \tilde{B}$, and $\kappa(F)=\tilde{B} / \tilde{A}$ are called lower/upper Riesz bounds and condition of the Riesz system, respectively.

Any finite set in $V$ is a frame system, any finite set of linearly independent elements of $V$ is a Riesz system (the question reduces then just to the size of the frame bounds). Therefore, most of the following discussion is substantial only for infinite $F$ and $\operatorname{dim} V=$ $\infty$.

Orthonormal systems (ONS) in $V$ are obviously frame systems and Riesz systems at the same time (if we have a CONS then it is a frame and Riesz basis), use the orthogonality and Bessel equality. Transformations of a CONS $\left\{e_{j}\right\} \subset V$ give rise to more examples ('real-world' examples are given later), cf. [8, 11].

Example 1. After scaling, a CONS (Riesz or frame system) may loose these properties. E.g.,

$$
\left\{e_{1}, \frac{1}{2} e_{2}, \ldots, \frac{1}{k} e_{k}, \ldots\right\}
$$

is not a frame system $(A=0!)$ nor a Riesz system $(\tilde{B}=\infty!)$ although it remains an (orthogonal) Schauder basis. More generally, $\left\{\alpha_{k} e_{k}\right\}$ is a Riesz basis (a frame) in $V$ if and only if $0<A=\inf _{k} \alpha_{k} \leq \sup _{k} \alpha_{k}=B<\infty$. Scaled CONS are clearly minimal, as frames they are tight only if $\left\{\alpha_{k}\right\}$ is a constant sequence.

Example 2. The system

$$
\left\{e_{1} ; \frac{1}{\sqrt{2}} e_{2}, \frac{1}{\sqrt{2}} e_{2} ; \frac{1}{\sqrt{3}} e_{3}, \frac{1}{\sqrt{3}} e_{3}, \frac{1}{\sqrt{3}} e_{3} ; \ldots\right\}
$$

is a first example of a tight frame with $A=B=1$ which is not minimal (e.g., the 'copies' of scaled $e_{k}$ can be deleted without changing the density in $V$ ). This is the typical feature of a frame: it contains redundancy. The subsystem $\left\{e_{1}, \frac{1}{\sqrt{2}} e_{2}, \frac{1}{\sqrt{3}} e_{3}, \ldots\right\}$ is not a frame. This is a bad property, and in sharp contrast to Riesz systems: It is obvious from definition (6) that any subsystem of a Riesz system is again a Riesz system, with the same (or better) Riesz bounds.

Example 3 [15]. Set

$$
F=\left\{f_{1}=e_{1}, f_{2}=e_{1}+\frac{1}{2} e_{2}, \ldots, f_{k}=e_{k-1}+\frac{1}{k} e_{k}, \ldots\right\} .
$$

Since

$$
\frac{1}{2}\left(f, e_{k}\right)^{2}-\frac{1}{(k+1)^{2}}\left(f, e_{k+1}\right)^{2} \leq\left(f, f_{k}\right)^{2} \leq 2\left(\left(f, e_{k}\right)^{2}+\frac{1}{(k+1)^{2}}\left(f, e_{k+1}\right)^{2}\right), k \geq 1
$$

this is a frame (Exercise: find good frame bounds!). The crazy thing about this frame is that for any increasing sequence of finite subsystems $F_{n}$ such that $F_{n} \rightarrow F$ (i.e., any $f_{k}$ belongs to all $F_{n}$ with sufficiently large $n \geq n_{0}(k)$ ) the lower frame bound $A_{n}$ deteriorates: $A_{n} \rightarrow 0$. To see this in the specific case of $F_{n}=\left\{f_{1}, \ldots, f_{n}\right\}$, take

$$
f=e_{1}-2!e_{2}+3!e_{3}-\ldots+(-1)^{n+1} n!e_{n} \in\left[F_{n}\right]
$$

as test function in (5). Show that $\kappa\left(F_{n}\right) \geq(n!)^{2}$ which is a dramatic blow-up. This example is important to have in mind in connection with discrete algorithms based on a frame (compare [15] and our considerations below).

In general, one may consider the whole class of systems $F$ generated by suitable linear transformations $T$ from a CONS (i.e., $f_{k}=\sum_{j} t_{k j} e_{j}$ ) and study various properties of a system in a systematic way. This is not too hard on an abstract level (compare $[8,16])$ and left upon the reader. Without proof, let us formulate the following theorem on the connections between CONS, Riesz bases, and frames (for historical references and more details, see [8] and the article by Benedetto/Walnut in [13]).

Theorem 1 a) $F$ is a CONS (ONS) in $V$ if and only if it is a tight frame (frame system) with $A=B=1$ and $\left\|f_{k}\right\|=1$.
b) $F$ is a Riesz basis (Riesz system) in $V$ if and only if it is a minimal frame (frame system). In this case, $\tilde{A}=1 / B, \tilde{B}=1 / A$, and the definitions of frame and Riesz condition are consistent.

We turn to the operators associated with a frame which allow us to formulate the main properties and applications of frames for representation purposes. Proofs will be given in a more general context in Section 3. Let $F$ be a frame in $V$ (the consideration of frame systems is similar). Then the synthesis operator $R$ given by

$$
c=\left(c_{k}\right) \in \ell^{2} \quad \longmapsto \quad R c=\sum_{k} c_{k} f_{k} \in V
$$

is well-defined on the sequence space $\ell^{2}$ (with index set inherited from $F$ ) and a bounded as linear operator from $\ell^{2}$ to $V$. Its adjoint $R^{*}: V \rightarrow \ell^{2}$ takes the form

$$
f \in V \quad \longmapsto \quad R^{*} f=\left(\left(f, f_{k}\right)\right) \in \ell^{2}
$$

and is called analysis operator. The boundedness of $R$ and $R^{*}$ follows exclusively from the upper estimate in the definition (5). The two-sided inequality (5) can be rephrased as

$$
A(f, f) \leq\left\|R^{*} f\right\|_{\ell^{2}}^{2}=\left(R R^{*} f, f\right) \leq B(f, f) \quad \forall f \in V
$$

which shows that the symmetric operator $\mathcal{P}=R R^{*}: V \rightarrow V$ is boundedly invertible with

$$
\begin{equation*}
A \operatorname{Id} \leq \mathcal{P} \leq B \operatorname{Id},\|\mathcal{P}\|_{V \rightarrow V}=B, \quad \frac{1}{B} \operatorname{Id} \leq \mathcal{P}^{-1} \leq \frac{1}{A} \operatorname{Id},\left\|\mathcal{P}^{-1}\right\|_{V \rightarrow V}=\frac{1}{A} \tag{8}
\end{equation*}
$$

It is assumed in (8) that $A, B$ are the best possible constants in (5), i.e. the frame bounds of $F$. As a consequence, the spectral condition number of $\mathcal{P}$ coincides with the frame condition:

$$
\begin{equation*}
\kappa(P)=\|\mathcal{P}\|_{V \rightarrow V}\left\|\mathcal{P}^{-1}\right\|_{V \rightarrow V}=\kappa(F) . \tag{9}
\end{equation*}
$$

The operator $\mathcal{P}$ is called frame operator and is central to all applications. Obviously,

$$
\begin{equation*}
f=\sum_{k}\left(f, f_{k}\right) \mathcal{P}^{-1} f_{k}=\sum_{k}\left(\mathcal{P}^{-1} f, f_{k}\right) f_{k}=\sum_{k}\left(f, \mathcal{P}^{-1} f_{k}\right) f_{k} \quad \forall f \in V \tag{10}
\end{equation*}
$$

The system $\tilde{F}=\left\{\tilde{f}_{k}=\mathcal{P}^{-1} f_{k}\right\}$ is called dual frame. It is easy to see that $\tilde{F}$ is indeed a frame, with frame operator $\mathcal{P}^{-1}$. Since for tight frames $\mathcal{P}=A \operatorname{Id}(=B \operatorname{Id})$, in this (and only this) case $\tilde{F}$ concides with $F$ up to a constant multiple $1 / A$ which means that tight frames are essentially self-dual (up to constant scaling).

The equation (10) shows why frames deserve the name stable representation system: There is an $\ell^{2}$-stable procedure of computing coefficients $c_{k}$ by linear functionals (more precisely, $\left.c_{k}=\left(f, \tilde{f}_{k}\right)\right)$ for representing arbitrary $f \in V$ by a series with respect to $F$.

Linearity and stability of representations are essential for many applications. Clearly, uniqueness of representation (which holds only if $F$ is a Riesz basis) is an additional desire but not essential in some other application areas.

The frame decomposition is optimal in a certain sense as was already observed in the paper [6].

Theorem 2 Let $F$ be a frame in $V$. The representation of an arbitrary $f \in V$ in (10) is optimal in the sense that if $f=\sum_{k} c_{k} f_{k}$ for some coefficient sequence $c$ then

$$
\sum_{k} c_{k}^{2} \geq \sum_{k}\left|\left(f, \tilde{f}_{k}\right)\right|^{2}
$$

Thus,

$$
\left\|\left.\left|f\left\|^{2} \equiv \inf _{c: f=\sum_{k} c_{k} f_{k}}\right\| c \|_{\ell^{2}}^{2}=\sum_{k}\right|\left(f, \tilde{f}_{k}\right)\right|^{2}=\left(\mathcal{P}^{-1} f, f\right) \quad \forall f \in V .\right.
$$

Theorem 2 and (8) show that in the definition of a frame the basic inequalities (5) can be replaced by the requirement

$$
\begin{equation*}
\frac{1}{B}\|f\|^{2} \leq\| \| f\left\|^{2} \leq \frac{1}{A}\right\| f \|^{2} \quad \forall f \in V \tag{11}
\end{equation*}
$$

which makes no use of the specific scalar product, shows the robustness (up to constants) of the frame property with respect to spectrally equivalent changes of the Hilbert space structure on $V$, and is easier to compare with the definition (6) of a Riesz basis (in the latter case, due to the uniqueness of representation the infimum in the definition of the triple bar norm is superflous). Note that $\|\mid f\| \|=+\infty$ if $f$ has no representation with $\ell^{2}$-coefficients or has no representation at all (by definition of the infimum taken over an empty set). Thus, density (and a bit more) of $F$ is implicitly assumed in (11).

Computing with frame representations means, in one or the other way, to compute $\mathcal{P}^{-1}$ on certain elements of $V$, or equivalently, to solve the operator equation

$$
\mathcal{P} g=f
$$

for given $f$. It was already proposed in [6] (and repeated by other authors, compare, e.g., [13, Section 8.2]) that simple Richardson iteration

$$
g^{(n+1)}=g^{(n)}-\omega\left(\mathcal{P} g^{(n)}-f\right), \quad n \geq 0
$$

with parameter $\omega=2 /(A+B)$ and arbitrary starting element $g^{(0)}$ could be used. This gives (in the best case) a convergence rate

$$
\rho_{R}=\rho(\operatorname{Id}-\omega \mathcal{P})=1-\frac{2}{1+\kappa(F)} .
$$

Note that this convergence rate is exclusively depending on the frame condition (we have used (9)!), which makes tight frames particularly interesting. The method heavily
depends on knowledge about good bounds for $A, B$ which is considered a nontrivial task. Application of the slightly more involved conjugate gradient method is possible since $\mathcal{P}$ is symmetric, and would overcome this difficulty resulting in a even better rate if the frame is far from being a tight one. Other iterative methods might be tried as well. There is another tricky point. In many applications, the theoretical investigations are for infinite frames (in infinite-dimensional $V$ ) while the real algorithms work with sections of the frame. Example 3 above shows that one has to care about the conditioning of these finite sections (here, using a Riesz basis would simplify the matter).

Finally, note that there is another interesting operator

$$
\tilde{\mathcal{P}}=R^{*} R: \ell^{2} \rightarrow \ell^{2},
$$

which is also symmetric (in $\ell^{2}$ ) but not necessarily invertible. Its matrix representation (with respect to the index set of $F$ ) is

$$
\tilde{\mathcal{P}}=\left(\tilde{\mathcal{P}}_{k, j}=\left(f_{j}, f_{k}\right)\right),
$$

which makes the name Gramian of $F$ plausible for $\tilde{\mathcal{P}}$. Again, $\tilde{\mathcal{P}}$ can be used for characterizing properties of a frame (see [16] for this kind of analysis in a special case).

We conclude with a list of practically important examples of frames appearing in the literature. All of them are function frames, and we will present their simplest onedimensional versions (with respect to (subspaces of) $L_{2}(\mathbb{R})$ ). Generalizations to more general Hilbert and Banach spaces (atomic decompositions, $\phi$-transform) can be found in [9] and [13, Chapters 3, 6, and 16].

- Irregular sampling. Frames have originated [6] from sampling bandlimited signals at general, irregularly spaced locations. A comprehensive survey of the mathematical and computational aspects of this problem is given by Groechenig and Feichtinger [13, Chapter 8]. The problem is as follows: For which sequences of locations $x_{n}$ are there constants $A, B$ such that

$$
A\|f\|^{2} \leq \sum_{n}\left|f\left(x_{n}\right)\right|^{2} \leq B\|f\|_{2}
$$

for all $f \in B_{\omega}$ where $B_{\omega}$ is the subspace of all functions $f \in L_{2}(\mathbb{R})$ such that $\operatorname{supp} \hat{f} \subset[-\omega, \omega]$ (in other words, $B_{\omega}$ can be obtained by taking the inverse Fourier transform of all functions in $\left.L_{2}(-\omega, \omega)\right)$ ? For $f \in B_{\omega}$ we have

$$
f(x)=\frac{1}{2 \pi} \int_{\mathbf{R}} \hat{f}(\xi) e^{i x \xi} \chi_{[-\omega, \omega]} d \xi=\int_{\mathbf{R}} f(t) \frac{\sin (\omega(t-x))}{\pi(t-x)} d t
$$

where $\hat{f}$ denotes the Fourier transform of $f$. Thus, setting $f_{n}(t)=\sin (\omega(t-$ $\left.\left.x_{n}\right)\right) /\left(\pi\left(t-x_{n}\right)\right)$ and comparing with the frame definition, we see that the above question of stable reconstruction of $f \in B_{\omega}$ from the sample values $\left(f\left(x_{n}\right)\right)$ is equivalent to $F=\left\{f_{n}\right\}$ being a frame in $B_{\omega}$. A necessary and sufficient condition
for this has recently be found by Jaffard whose work is based on previous contributions by Duffin/Schaeffer and Landau, for details, see [13, Chapter 8]. Going to the Fourier transform domain, we see that there is yet another equivalent formulation (in terms of non-harmonic Fourier series) of the above irregular sampling problem which goes back to Wiener/Paley (see [6]): Under which conditions on the frequencies $x_{n}$ is the system $\left\{e^{-i x_{n} t}\right\}$ a frame in $L_{2}(-\omega, \omega)$ ?

- Gabor frames. Gabor frames are function systems of the form

$$
g_{m n}(t)=e^{2 \pi i m b t} g(t-n a), \quad m, n \in \mathbb{Z}
$$

which are generated by translation and modulation from a single function $g \in$ $L_{2}(\mathbb{R})$. Gabor's original proposal was to use the Gaussian

$$
g(t)=C e^{-s t^{2}}
$$

where $s>0$ is a fixed parameter, and $C>0$ some scaling constant. The point is that then $g_{m n}$ is a 'minimizer' (i.e., realizes equality) of the following uncertainty principle for localization in time-frequency domain: For any $f \in L_{2}(\mathbb{R})$,

$$
\|f\|^{2} \leq 4 \pi\|(t-n a) f(t)\|\|(\xi-m b) \hat{f}(\xi)\|
$$

Thus, looking for decompositions

$$
f(t)=\sum_{m, n} c_{m n} g_{m n}(t)
$$

means to look for a decomposition with respect to functions which are best localized with respect to the different points in a lattice $\{(n a, m b)\}$ in the time-frequency plane. Again, stable decomposition is equivalent to the frame property of the systems $F_{a, b ; g}=\left\{g_{m n}\right\}$. A detailed discussion of Gabor frames can be found in [13, Chapter 3 and 7]. E.g., Theorem 7.8 there states that for the Gaussian $g(t)$ as defined above $F_{a, b ; g}$ is a frame in $L_{2}(\mathbb{R})$ if and only if $a b<1$. In addition, [13, Chapter 5] introduces to a class of similar systems, the so-called Malvar-Wilson bases or local Fourier bases which have found applications in signal analysis and, more recently, to operator equations (compare work by the Coifman group).

- Wavelet frames. If modulation is replaced by dilation we arrive at wavelet systems. More precisely, given a normalized function $\psi \in L_{2}(\mathbb{R})$, we define

$$
\psi_{j, i}(t)=2^{j / 2} \psi\left(2^{j} t-i\right), \quad j, i \in \mathbb{Z}
$$

The classical counterparts of this construction are the Haar and the Faber-Schauder system. These are obtained if the functions depicted in Figure 1 a), b) are used as the basic $\psi$. Both choices lead to minimal systems. In the Haar case, the resulting wavelet system $F_{\psi}=\left\{\psi_{j, i}\right\}$ is even a CONS in $L_{2}(\mathbb{R})$. The other system is not a
frame in $L_{2}(\mathbb{R})$. The system $F_{\psi}$ resulting from the hat function in Figure 1 c) is the prototype of a multilevel frame in Sobolev spaces which has generalizations to finite element multigrid schemes in higher dimensions. Since wavelet systems will be discussed in later sections in connection with efficient adaptive solution strategies for operator equations, we will stop with these examples. More information can be found in $[11,10,12,13]$.

a) Haar wavelet

b) Faber-Schauder 'wavelet' (hierarchical basis)

c) BPX 'wavelet' (multilevel nodal basis frame)

Figure 1: Three basic 'wavelets' $\psi$

## 3 Stable space splittings

We will start with some notation which will be consistently used throughout the exposition. Again, $V$ is the basic Hilbert space, with $(\cdot, \cdot)$ resp. $\langle\cdot, \cdot\rangle \equiv\langle\cdot, \cdot\rangle_{V^{\prime} \times V}$ as basic scalar product resp. duality pairing. Consider a symmetric $V$-elliptic variational problem (4) to be solved. As discussed in Section 1, $\{V ; a(\cdot, \cdot)\}$ is an isomorphic copy of $V$. Let $V_{j}, j=1,2, \ldots$, be an at most countable family of Hilbert spaces, with $(\cdot, \cdot)_{j},\langle\cdot, \cdot\rangle_{j}$ introduced similarly. To each $V_{j}$ we assign its own symmetric $V_{j}$-elliptic bilinear form $b_{j}(\cdot, \cdot): V_{j} \times V_{j} \rightarrow \mathbb{R}$ which in particular means that $\left\{V_{j} ; b_{j}(\cdot, \cdot)\right\}$ are Hilbert spaces. The $V_{j}$ and $b_{j}(\cdot, \cdot)$ will be used to create simpler (both in structure and size) auxiliary problems and to compose from their solution operators an approximate inverse to $A$. The latter is then used as a preconditioner in an iterative method for solving (4), see Section 4 for the details. It is not assumed that the $V_{j}$ are subspaces of $V$ (but it is implicit that they correspond to certain portions of $V$, see below).

Denote the Hilbert sum of this family by $\tilde{V}$, i.e., for

$$
\tilde{u}=\left(u_{j}\right), \tilde{v}=\left(v_{j}\right), \quad u_{j}, v_{j} \in V_{j} \forall j
$$

set

$$
\tilde{a}(\tilde{u}, \tilde{v})=\sum_{j} b_{j}\left(u_{j}, v_{j}\right)
$$

which makes sense as a scalar product on

$$
\tilde{V}=\{\tilde{u}: \tilde{a}(\tilde{u}, \tilde{u})<\infty\} .
$$

Finally, consider bounded linear mappings $R_{j}: V_{j} \rightarrow V$. Formally, they can be considered as the components of an operator $R: \tilde{V} \rightarrow V$ given by $R \tilde{u}=\sum_{j} R_{j} u_{j}$.

Definition 3 ([25]) The system $\left\{\left\{V_{j} ; b_{j}\right\}, R_{j}\right\}$ gives rise to a stable splitting of $\{V ; a\}$ which will be expressed by the short-hand notation

$$
\begin{equation*}
\{V ; a\}=\sum_{j} R_{j}\left\{V_{j} ; b_{j}\right\} \tag{12}
\end{equation*}
$$

if there are two constants $0<\tilde{A} \leq \tilde{B}<\infty$ such that

$$
\begin{equation*}
\tilde{A} a(u, u) \leq\|\mid u\| \|^{2} \equiv \inf _{\tilde{u} \in \tilde{V}: u=R \tilde{u}} \tilde{a}(\tilde{u}, \tilde{u}) \leq \tilde{B} a(u, u) \quad \forall u \in V \tag{13}
\end{equation*}
$$

The optimal constants $\tilde{A}, \tilde{B}$ in (13) will be called lower and upper stability constants, and their ratio $\kappa=\tilde{B} / \tilde{A}$ condition of the splitting (12).

It should be noted that (13) implicitly requires that $R$ makes sense (convergence of the sum if infinitely many $V_{j}$ are involved) and is surjective, i.e., $\operatorname{ran}(R)=V$. A sufficient condition for $R$ being well-defined and bounded on all of $\tilde{V}$ is that on the set of all finite sums

$$
u=\sum_{j=1}^{n} R_{j} u_{j}, \quad u_{j} \in V_{j}, j=1, \ldots, n, n \geq 1
$$

the following replacement for the lower estimate in (13) is satisfied:

$$
\begin{equation*}
a\left(\sum_{j=1}^{n} R_{j} u_{j}, \sum_{j=1}^{n} R_{j} u_{j}\right) \leq \frac{1}{\tilde{A}} \sum_{j=1}^{n} b_{j}\left(u_{j}, u_{j}\right) . \tag{14}
\end{equation*}
$$

Indeed, if $\tilde{u} \in \tilde{V}$ is arbitrary then for the sections $u^{m, n}=\sum_{j=m}^{n} R_{j} u_{j}$ of the series under consideration we have from (14)

$$
a\left(u^{m, n}, u^{m, n}\right) \leq \frac{1}{\tilde{A}} \sum_{j=m}^{n} b_{j}\left(u_{j}, u_{j}\right),
$$

where the right-hand side converges to zero for $m, n \rightarrow \infty$ since the infinite sum converges due to $\tilde{u} \in \tilde{V}$. Thus, the partial sums of the series $\sum_{j} R_{j} u_{j}$ form a Cauchy sequence in $V$ and, hence, are convergent to some $u=R \tilde{u} \in V$. Taking $m=1$ and letting $n \rightarrow \infty$ the boundedness of $R: \tilde{V} \rightarrow V$ comes out: $\|R\|_{\tilde{V} \rightarrow V}^{2} \leq 1 / \tilde{A}$.

By definition, the adjoint $R^{*}: V \rightarrow \tilde{V}$ is defined as

$$
R^{*}: u \in V \quad \longmapsto R^{*} u=\left(R_{1}^{*} u, R_{2}^{*} u, \ldots\right) \in \tilde{V},
$$

where the components $R_{j}^{*}: V \rightarrow V_{j}$ are determined by solving the auxiliary variational problems:

$$
\begin{equation*}
b_{j}\left(R_{j}^{*} u, v_{j}\right)=a\left(u, R_{j} v_{j}\right) \quad \forall v_{j} \in V_{j} \tag{15}
\end{equation*}
$$

We postpone the derivation of the familiar representations of these operators (which follow if the duality pairings are used) to the beginning of Section 4 and concentrate on the pure functional-analytic setting. Clearly, $R^{*}$ is bounded as well ( $\left.\left\|R^{*}\right\|_{V \rightarrow \tilde{V}}^{2} \leq 1 / \tilde{A}\right)$, and we can introduce the two bounded linear operators

$$
\begin{equation*}
\mathcal{P}=R R^{*}: u \in V \longmapsto \mathcal{P} u=\sum_{j} T_{j} u \in V \quad\left(T_{j}=R_{j} R_{j}^{*}: V \rightarrow V\right) \tag{16}
\end{equation*}
$$

and

$$
\begin{equation*}
\tilde{\mathcal{P}}=R^{*} R: \tilde{u} \rightarrow \tilde{\mathcal{P}} \tilde{u} \in \tilde{V} \tag{17}
\end{equation*}
$$

where $\tilde{\mathcal{P}}$ can be considered as operator matrix which acts according to

$$
(\tilde{\mathcal{P}} \tilde{u})_{j}=\sum_{k} \overbrace{R_{j}^{*} R_{k}}^{\tilde{P}_{j k}} u_{k} \quad \forall j
$$

Following some tradition $[18,20], \mathcal{P}$ is called Schwarz operator associated with the stable splitting (12) while the operator matrix associated with $\tilde{\mathcal{P}}$ will be called extended Schwarz operator (it is nothing but the generalization of the Gramian for frames discussed in Section 2, and the abstract analog of the matrix of the semi-definite system [26]). We next prove an analog of Theorem 2.

Theorem 4 The Schwarz operator (16) associated with a stable splitting (12) is symmetric positive definite and has a bounded inverse. Moreover,

$$
\|\mid u\| \|^{2}=a\left(\mathcal{P}^{-1} u, u\right) \quad \forall u \in V
$$

and

$$
\frac{1}{\tilde{B}} \operatorname{Id} \leq \mathcal{P} \leq \frac{1}{\tilde{A}} \operatorname{Id}, \quad \kappa(\mathcal{P})=\kappa
$$

Proof. See [25, pp.73-75], where the proof is reduced to using Nepomnyashchich's Fictitious Space Lemma. To be self-contained, here is the argument. First, symmetry follows from

$$
a(\mathcal{P} u, v)=\tilde{a}\left(R^{*} u, R^{*} v\right)=\sum_{j} b_{j}\left(R_{j}^{*} u, R_{j}^{*} v\right)
$$

As a by-product, note the formula

$$
\begin{equation*}
a(\mathcal{P} u, u)=\tilde{a}\left(R^{*} u, R^{*} u\right)=\sum_{j} b_{j}\left(R_{j}^{*} u, R_{j}^{*} u\right) \geq 0 \tag{18}
\end{equation*}
$$

which also shows that $\mathcal{P}$ is non-negative.

Next consider any $\tilde{v} \in \tilde{V}$ such that $\mathcal{P} u=R \tilde{v}$ (one such $\tilde{v}$ can be given explicitly: $\left.\tilde{v}_{0}=R^{*} u\right)$. Then

$$
a(\mathcal{P} u, u)^{2}=a(R \tilde{v}, u)^{2}=\tilde{a}\left(\tilde{v}, R^{*} u\right)^{2} \leq \tilde{a}(\tilde{v}, \tilde{v}) \tilde{a}\left(R^{*} u, R^{*} u\right)=\tilde{a}(\tilde{v}, \tilde{v}) a(\mathcal{P} u, u),
$$

with equality attained for $\tilde{v}_{0}$. Thus,

$$
a(\mathcal{P} u, u)=\inf _{\mathcal{P} u=R \tilde{v}} \tilde{a}(\tilde{v}, \tilde{v})=\left\||\mathcal{P} u \||^{2} .\right.
$$

With this intermediate result at hand, we can establish the invertibility of $\mathcal{P}$. Indeed,

$$
a(\mathcal{P} u, u)^{2} \leq a(\mathcal{P} u, \mathcal{P} u) a(u, u) \leq \frac{1}{\tilde{A}}\|\mid \mathcal{P} u\| \|^{2} a(u, u)=\frac{1}{\tilde{A}} a(\mathcal{P} u, u) a(u, u)
$$

where the lower stability estimate has been incorporated. This gives $\mathcal{P} \leq 1 / \tilde{A} \cdot$ Id. On the other hand, consider any $\tilde{v} \in \tilde{V}$ such that $u=R \tilde{v}$ (the existence of such $\tilde{v}$ follows from the upper stability estimate). Then we have in the same way as above

$$
a(u, u)^{2}=\tilde{a}\left(\tilde{v}, R^{*} u\right)^{2} \leq \tilde{a}(\tilde{v}, \tilde{v}) a(\mathcal{P} u, u) .
$$

After taking the infimum with respect to all admissible $\tilde{v}$ and using the upper stability bound, we see that

$$
a(u, u)^{2} \leq\||u \|| a(\mathcal{P} u, u) \leq \tilde{B} a(u, u) a(\mathcal{P} u, u),
$$

from which $1 / \tilde{B} \cdot \operatorname{Id} \leq \mathcal{P}$ follows. This argument includes the proof of positive definitness, and the invertibility of $\mathcal{P}$ follows from writing

$$
\mathcal{P}=\frac{1}{\omega}(\operatorname{Id}-(\operatorname{Id}-\omega \mathcal{P})) \equiv \frac{1}{\omega}(\operatorname{Id}-B),
$$

where $B=\mathrm{Id}-\omega \mathcal{P}$ is symmetric and satisfies

$$
\left(1-\frac{\omega}{\tilde{A}}\right) a(u, u) \leq a(B u, u) \leq\left(1-\frac{\omega}{\tilde{B}}\right) a(u, u) .
$$

Thus, taking $\omega=2 \tilde{A} \tilde{B} /(\tilde{A}+\tilde{B})$, we get

$$
\|B\|_{a} \leq \frac{\tilde{B}-\tilde{A}}{\tilde{B}+\tilde{A}}<1
$$

Now, the invertibility of $\mathcal{P}$ comes from applying a Neumann series argument:

$$
\mathcal{P}^{-1}=\omega \sum_{k=0}^{\infty} B^{k} .
$$

Clearly, $\tilde{A} \leq \mathcal{P}^{-1} \leq \tilde{B}$. The arguments show that all estimates are sharp which gives

$$
\kappa(\mathcal{P})=\|\mathcal{P}\|_{a}\left\|\mathcal{P}^{-1}\right\|_{a}=\frac{\tilde{B}}{\tilde{A}}=\kappa .
$$

Finally, to get the expression for $\|\mid u\| \|^{2}$, substitute $\mathcal{P}^{-1} u$ for $u$ in the above expression of $\|\mid \mathcal{P} u\| \|^{2}$.

Example 4. Let us first make precise why frame theory is a special case (which is already expressed by the notation). Let $F=\left\{f_{j}\right\}$ be a frame in $V$, set $a(\cdot, \cdot)=(\cdot, \cdot)$. Without loss of generality, we can assume that all $f_{j} \neq 0$. Denote by $V_{j}$ the onedimensional subspace spanned by $f_{j}$, and set

$$
b_{j}\left(u_{j}, v_{j}\right)=\frac{\left(u_{j}, v_{j}\right)}{\left(f_{j}, f_{j}\right)} \quad \forall u_{j}, v_{j} \in V_{j}, \quad j \geq 1
$$

As $R_{j}$ we take the natural embeddings. Then the space $\tilde{V}$ is formed by all sequences

$$
\tilde{u}=\left(u_{j}\right)=\left(c_{j} f_{j}\right) \longleftrightarrow c=\left(c_{j}\right)
$$

with

$$
\tilde{a}(\tilde{u}, \tilde{u})=\sum_{j} c_{j}^{2} .
$$

which means that $\tilde{V}$ is an isometric copy of the sequence space $\ell^{2}$ used in the frame context. On the basis of (15) one computes

$$
\frac{\left(R_{j}^{*} u, f_{j}\right)}{\left(f_{j}, f_{j}\right)}=\left(u, f_{j}\right) \Longrightarrow R_{j}^{*} u=\left(u, f_{j}\right) f_{j}
$$

for all $j$ and after substitution into (18) one arrives at the familiar equation for frames:

$$
a(P u, u)=(P u, u)=\sum_{j}\left|\left(u, f_{j}\right)\right|^{2} \quad \forall u \in V
$$

$R$ and $R^{*}$ coincide (up to isometry) with the synthesis and analysis operator for frames while the Schwarz operator is nothing but the frame operator. Thus, stable splittings in the sense of the above definition are a generalization of frames in two directions. First, instead of decomposing with respect to a fixed system of elements, best decompositions $u=\sum_{j} R_{j} u_{j}$ are considered where the terms are chosen from auxiliary spaces $V_{j}$ of arbitrary dimension. This adds a greater flexibility, especially in connection with parallelization of computations. Secondly, the auxiliary spaces need not be subspaces which is of interest for a number of applications (e.g., for outer approximation schemes).

On the other hand, since frames (and Riesz bases) provide in some sense the most detailed decomposition of a space they are particularly useful in adaptive computations. Also, by clustering techniques (see Section 5) frames may be successfully used to construct more complicated subspace splittings. Last but not least, we may try to introduce methods and notions which have shown their usefulness in the frame case to the case of space splittings. E.g., one might call

$$
\{V ; a\}=\sum_{j} \hat{R}_{j}\left\{V_{j} ; b_{j}\right\}, \quad \hat{R}_{j}=\mathcal{P}^{-1} R_{j}
$$

dual space splitting since with this new set of operators $\hat{R}_{j}$ the generalization of (10)

$$
I d=\hat{R} R^{*}=R \hat{R}^{*}
$$

holds. It has not yet been investigated which of the frame concepts give rise to interesting applications in the area of space splittings vice versa.

Example 5. This example provides a bridge with the material of the next section. Let $V=\mathbb{R}^{n}$, and $A$ be a real $n \times n$ spd matrix. The scalar product is the Euclidean vector product in $\mathbb{R}^{n}$. Set $a(x, y)=(A x, y), x, y \in V$. Let $V_{j} \subset V$ denote the $j$ th coordinate space, i.e., the set of all $\mathbb{R}^{n}$ vectors where only the $j$-th coordinate is nonvanishing. Introduce

$$
b_{j}(x, y)=b_{j} x_{j} y_{j}, \quad x, y \in V_{j} \quad\left(b_{j}>0\right)
$$

Again, $R_{j}$ are the natural embeddings for subspaces. Since we have a splitting of a finite-dimensional space into a finite sum of subspaces, (13) is guaranteed (with some constants). Direct computation yields

$$
\left(R_{i}^{*} x\right)_{i}=\frac{1}{b_{i}} \sum_{j=1}^{n} a_{i j} x_{j}, \quad i=1, \ldots, n
$$

Thus, the action of $\mathcal{P}$ (as well as the "operator" matrix $\tilde{\mathcal{P}}$ ) is given by the matrix $D^{-1} A$ where $D=\operatorname{diag}(b)$. Formally, the introduction of the above splitting has led to a Schwarz operator whose representation is a preconditioned version of the initial matrix $A$. This is a general fact, and the theory of stable splittings can be viewed as a theoretical tool to control the condition number of the preconditioned operator $\mathcal{P}$ via the stability constants $\tilde{A}, \tilde{B}$ of the splitting. Moreover, classical iterative methods (Richardson/Jacobi, GaussSeidel/SOR) can be recovered as subspace correction methods [20]. E.g., $b_{i}=1 / \omega>0$ leads to the Richardson iteration (with relaxation parameter $\omega$ ), and $b_{i}=a_{i i}$ to the Jacobi method. In turn, when applied to the operator matrix $\tilde{\mathcal{P}}$, classical methods such as Richardson iteration or SOR result in modern algorithms for large linear systems such as domain decomposition and multigrid methods.

Example 6. This is the core example from the field of domain decomposition methods which has influenced the appearance of the present theory of stable splittings very much. Consider the Poisson equation for Laplace's equation

$$
\begin{equation*}
-\Delta u=f \quad \text { in } \Omega, \quad u=0 \quad \text { on } \partial \Omega . \tag{19}
\end{equation*}
$$

If we seek for weak solutions, this problem is turned into the form (4) with $V=H_{0}^{1}(\Omega)$. As usual,

$$
a(u, v)=\int_{\Omega} \nabla u \cdot \nabla v, \quad u, v \in V .
$$

Domain decomposition ideas come in if one tries to split this problem into a finite number of similar problems with respect to some subdomains $\Omega_{j}$. Figure 2 shows the case of two


Figure 2: Overlapping and non-overlapping domain decomposition
subdomains $\Omega_{1}$ and $\Omega_{2}$. The left picture corresponds to the case of sufficient overlap, the other one to the case of nonoverlapping domains.

Set $V_{j}=H_{0}^{1}\left(\Omega_{j}\right)$ and

$$
b_{j}(u, v)=\int_{\Omega_{j}} \nabla u \cdot \nabla v, \quad u, v \in V_{j}, j=1,2 .
$$

For functions from $V_{j}$ defined on $\Omega_{j}$, extension by zero to the whole domain $\Omega$ is a simple choice for $R_{j}, j=1,2$. Then the auxiliary problems (15) are essentially Poisson problems on the domains $\Omega_{j}$. It turns out that

$$
\left\{H_{0}^{1}(\Omega) ; a\right\}=R_{1}\left\{H_{0}^{1}\left(\Omega_{1}\right) ; b_{1}\right\}+R_{2}\left\{H_{0}^{1}\left(\Omega_{2}\right) ; b_{2}\right\}
$$

is a stable splitting in the sense of our definition (use a smooth partition of unity adapted to the domain splitting to see the non-trivial upper inequality (13)). The constants depend on the regularity and the shape of the domains involved. The most critical parameter is the "thickness" of the corridor of overlap $\Omega_{1} \cap \Omega_{2}$, reducing it means blowing up $\tilde{B}$.

The famous alternating method by H. Schwarz is essentially a (mathematical) algorithm for solving (19) which fits into the algorithms explained in the next Section. Formally, it starts with an arbitrary initial guess $u^{(0)} \in H_{0}^{1}(\Omega)$ and does alternately the following:

$$
\begin{aligned}
v^{(k)} & =u^{(k)}-R_{1} B_{1}^{-1} R_{1}^{\prime}\left(-\Delta u^{(k)}-f\right) \\
u^{(k+1)} & =v^{(k)}-R_{2} B_{2}^{-1} R_{2}^{\prime}\left(-\Delta v^{(k)}-f\right)
\end{aligned}
$$

until a stopping criteria is fulfilled. In each iteration step, the residual $r(u)=-\Delta u-f$ has to be computed twice, for $u^{(k)}$ and the intermediate iterate $v^{(k)} . R_{j}^{\prime}$ is the restriction operator to $\Omega_{j}, R_{j}$ the extension-by-zero operator from above, and $B_{j}^{-1}$ is the solution operator for the Poisson problem in $H_{0}^{1}\left(\Omega_{j}\right)$ (with homogeneous Dirichlet boundary
values). An alternative to the alternating method of Schwarz would be to work with the simpler iteration

$$
u^{(k+1)}=u^{(k)}-R_{1} B_{1}^{-1} R_{1}^{\prime} r\left(u^{(k)}\right)-R_{2} B_{2}^{-1} R_{2}^{\prime} r\left(u^{(k)}\right), \quad k \geq 0 .
$$

Now only one residual has to be computed, and part of the computation can be done in parallel. The disadvantage is that in analogy to the Jacobi and Gauss-Seidel method, the second method is usually slower (it needs roughly twice as many iterations to reach the same error reduction).

In the nonoverlapping case (right picture in Figure 2), a third space which serves the data on the interface $\Gamma$ is needed. Its design is related to the trace operator. Set $V_{3}=\tilde{H}^{1 / 2}(\Gamma)$, and take for $R_{3}: V_{3} \rightarrow V$ some kind of harmonic extension of functions defined on $\Gamma$ to the whole domain $\Omega$, satisfying

$$
\left\|R_{3} f\right\|_{H^{1}(\Omega)} \leq C\|f\|_{\tilde{H}^{1 / 2}(\Gamma)},\left.\quad R_{3} f\right|_{\Gamma}=f .
$$

The exact meaning of $\tilde{H}^{1 / 2}(\Omega)$ and the construction of a suitable auxiliary form $b_{3}(\cdot, \cdot)$ which should be $V_{3}$-elliptic and symmetric, will not be explained here (Sobolev spaces of this type incorporate the zero values at the endpoints of $\Gamma$ and can be defined by interpolation). Under suitable assumptions on the regularity of $\Gamma$ and $\Omega$, the repaired splitting is again stable. In the practical application, the construction of an (approximate) inverse $B_{3}^{-1}$ (the so-called interface solver) is now the challenge. For more details and historical references, see [19].

Example 7. Stable splittings have a long tradition in connection with approximation processes. For the understanding of the present example, the booklet [24, Chapter 1-2] gives the necessary background (only the case $q=2, X_{1}=X=H, X_{2}=V \subset H$, $\theta_{1}=0, \theta_{2}=\gamma>0$ of [24, Satz 2.4.1] is of interest in the Hilbert space context).

Let

$$
\begin{equation*}
V_{0} \subset V_{1} \subset \ldots \subset V_{j} \subset \ldots \subset V=\cos _{V}\left(\cup_{j} V_{j}\right) \tag{20}
\end{equation*}
$$

be an increasing sequence of subspaces of $V$. Assume that $V$ is continuously embedded into another Hilbert space $H$, and that $H=\operatorname{clos}_{H}(V)$. To avoid any problems, we also require all $V_{j}$ to be subspaces of $H$, too. Fix some $a>1$ and $\gamma>0$. We say that, with respect to the family $\left\{V_{j}\right\}$, the pair $(H, V)$ satisfies a Jackson inequality if

$$
\begin{equation*}
e_{j}(u)_{H} \equiv \inf _{v_{j} \in V_{j}}\left\|u-v_{j}\right\| \leq C a^{-\gamma j}\|u\|_{V} \quad \forall u \in V \tag{21}
\end{equation*}
$$

resp. a Bernstein inequality if

$$
\begin{equation*}
\left\|v_{j}\right\|_{V} \leq C a^{\gamma j}\left\|v_{j}\right\|_{H} \quad \forall v_{j} \in V_{j} \tag{22}
\end{equation*}
$$

uniformly in $j \geq 0$. Often, one uses this abstract setting with $H=L_{2}(\Omega) \supset V=H^{\gamma}(\Omega)$. Typically (e.g., for finite element spaces $V_{j}$ obtained by regular dyadic refinement or in the case of dyadic multiresolution analysis), we have $a=2$ while the choice of $\gamma$
is connected with the degree of approximation obtainable from $\left\{V_{j}\right\}$ resp. with the smoothness of the elements $v_{j} \in V_{j}$.

From [24, Satz 2.4.1] it follows that

$$
\begin{equation*}
X^{s} \equiv[H, V]_{s / \gamma, 2}=\sum_{j}\left\{V_{j} ; a^{2 s j}(\cdot, \cdot)_{H}\right\}, \quad 0<s<\gamma \tag{23}
\end{equation*}
$$

is a stable splitting for the intermediate interpolation spaces $V \subset X^{s} \subset H$ which are obtained by the $K$-method. We have dropped the $R_{j}$ in the notation which indicates natural embedding for subspaces. The power of this result lies in the fact that these interpolation spaces $X^{s}$ are often nontrivial spaces (e.g., Sobolev spaces of smoothness parameter $0<s<\gamma$ ) while the spaces $V_{j}$ in the splitting are equipped with scaled $H\left(=L_{2}(\Omega)\right)$-scalar products which means that the auxiliary problems associated with the forms $b_{j}(\cdot, \cdot)$ are simpler and $s$-independent (up to a scaling factor).

A useful consequence of (23) is as follows. Let $a_{s}(\cdot, \cdot)$ denote a symmetric $X^{s}$-elliptic bilinear form for some fixed $s \in(0, \gamma)$. Then

$$
\begin{equation*}
\left\{V_{J} ; a_{s}(\cdot, \cdot)\right\}=\sum_{j=0}^{J}\left\{V_{j} ; a^{2 s j}(\cdot, \cdot)_{H}\right\} \tag{24}
\end{equation*}
$$

is a stable splitting for each $J$, with condition $\kappa_{J}$ that is uniformly bounded in $J$ : $\kappa_{J} \leq C \kappa(s)$, where $\kappa(s)$ is the condition of the splitting (23) for that particular $s$. The constant $C$ only depends on $a, s$, and the ellipticity constants of $a_{s}(\cdot, \cdot)$.

Let us prove this result (since this is essentially a result on the conditioning of subsplittings of (23), it deserves a proof due to the counterexamples of Section 2). Denote by $\tilde{A}(s), \tilde{B}(s)$ the stability constants for the splitting (23) (with respect to a fixed scalar product in $X^{s}$ ). The lower estimate is trivial since by definition of the triple bar norms $|\|\cdot\||$ (23) resp. $|||\cdot|||_{J}$ for (24) obviously

$$
\left\|\left|u_{J}\| \|_{J}^{2} \geq\left\|\mid u_{J}\right\|\left\|^{2} \geq \tilde{A}(s)\right\| u_{J} \|_{X^{s}}^{2} \geq c \tilde{A}(s) a_{s}\left(u_{J}, u_{J}\right) \quad \forall u_{J} \in V_{J}\right.\right.
$$

For the upper bound, by definition of the infimum for any $u_{J} \in V_{J}$ there are $v_{j} \in V_{j}, j \geq$ 0 , such that $u_{J}=\sum_{j} v_{j}$ and

$$
\sum_{j} a^{2 s j}\left\|v_{j}\right\|_{H}^{2} \leq 2 \tilde{B}(s)\left\|u_{J}\right\|_{X^{s}}^{2} \leq C \tilde{B}(s) a_{s}\left(u_{J}, u_{J}\right)
$$

Again, the last step is the ellipticity assumption for $a_{s}(\cdot, \cdot)$. Define

$$
\hat{v}_{j}=v_{j}, j<J, \quad \hat{v}_{J}=\sum_{j \geq J} v_{j}=u_{J}-\sum_{j=0}^{J-1} v_{j} \in V_{J}
$$

Thus, $u_{J}=\sum_{j=0}^{J} \hat{v}_{j}$ and since

$$
\left\|\hat{v}_{J}\right\|^{2} \leq\left(\sum_{j \geq J} a^{-j s} \cdot\left(a^{j s}\left\|v_{j}\right\|_{H}\right)\right)^{2}
$$

$$
\begin{aligned}
& \leq\left(\sum_{j \geq J} a^{-2 j s}\right)\left(\sum_{j \geq J} a^{2 j s}\left\|v_{j}\right\|_{H}^{2}\right) \\
& \leq C a^{-2 j s} \sum_{j \geq J} a^{2 j s}\left\|v_{j}\right\|_{H}^{2}
\end{aligned}
$$

we immediately obtain

$$
\left\|\mid u_{J}\right\|\left\|_{J}^{2} \leq \sum_{j=0}^{J} a^{2 j s}\right\| \hat{v}_{j}\left\|^{2} \leq C \sum_{j} a^{2 j s}\right\| v_{j} \|_{H}^{2} \leq C \tilde{B}(s) a_{s}\left(u_{J}, u_{J}\right)
$$

This proves the assertion.
In the same way one shows that the stability of (23) implies the stability of

$$
\begin{equation*}
X^{s}=\sum_{j}\left\{W_{j} ; a^{2 s j}(\cdot, \cdot)_{H}\right\}, \quad 0<s<\gamma \tag{25}
\end{equation*}
$$

where $W_{0}=V_{0}, W_{j}=\operatorname{ran}\left(P_{j}-P_{j-1}\right), j \geq 1$, are "difference" spaces generated by an arbitrary sequence of uniformly bounded projectors $P_{j}: H \rightarrow V_{j}$. Note that in this case $X^{s}$ is stably decomposed into a direct sum of subspaces which is the counterpart to a Riesz basis construction.

A special case of a family of projectors with the above properties are orthogonal projectors. With this, a possible advantage of direct sum decompositions becomes transparent: they may be even good for $H$ and some of the dual spaces $X^{-s}=\left(X^{s}\right)^{\prime}$. Indeed, let the linear operators $Q_{j}: H \rightarrow V_{j}$ be such that $Q_{j} v_{j}=v_{j}$ (projection property) and

$$
\left(v-Q_{j} v, v_{j}\right)_{H}=0 \quad \forall v_{j} \in V_{j} \forall v \in V \quad \text { (orthogonality). }
$$

Then obviously

$$
v=\underbrace{Q_{0} v}_{w_{0}}+\underbrace{\left(Q_{1}-Q_{0}\right) v}_{w_{1}}+\ldots+\underbrace{\left(Q_{j}-Q_{j-1}\right) v}_{w_{j}}+\ldots
$$

with mutually orthogonal terms $w_{j} \in W_{j}$ as defined above and

$$
\|v\|_{H}^{2}=\sum_{j}\left\|w_{j}\right\|_{H}^{2}
$$

The latter equality can be reinterpreted as stability assertion for the case $s=0$ in (25) and $H=X^{0}$.

Analogous results can be obtained for $-\gamma<s<0$ by a duality argument which we leave upon the reader (Hint: Use

$$
\|\tilde{v}\|_{X^{-s}}:=\sup _{v \neq 0} \frac{(\tilde{v}, v)_{H}}{\|v\|_{X^{s}}}, \quad \tilde{v} \in H
$$

the mutual $H$-orthogonality of the components of the decompositions of $\tilde{v}$ and $v$, together with the two-sided stability estimates for the splitting (25. Note that $X^{-s}$ is defined as the closure of $H$ under the above norm.)

The corresponding assertions are not true for the splitting (23). E.g., putting $s=0$ in the right-hand side of (23) does not lead to a stable splitting for $H$. Indeed, if one takes $u=u_{0} \in V_{0}$ then due to the monotonicity assumption (20), the decompositions

$$
u_{0}=\frac{1}{J+1} \underbrace{\left(u_{0}+u_{0}+\ldots+u_{0}\right)}_{J+1 \text { times }}
$$

are admissible in the definitions for the triple bar norms associated with both (23) and (24). Thus,

$$
\left\|\left|u_{0}\| \|^{2} \leq\left\|\mid u_{0}\right\|\left\|_{J}^{2} \leq(J+1)^{-1}\right\| u_{0} \|^{2} \quad \forall J \geq 0\right.\right.
$$

This shows the non-stability of the splitting for $H$.
What concerns the splitting of $\left\{V_{J} ;(\cdot, \cdot)_{H}\right\}$ given by (24) for $s=0$, we have

$$
\left\|u_{J}\right\|_{H}^{2} \leq\left(\sum_{j=0}^{J}\left\|v_{j}\right\|_{H}\right)^{2} \leq(J+1) \sum_{j=0}^{J}\left\|v_{j}\right\|_{H}^{2} \quad \forall u_{J}=\sum_{j=0}^{J} v_{j}
$$

which shows (together with the above opposite inequality for a particular function $u_{0} \in$ $\left.V_{0} \subset V_{J}\right)$ that $\tilde{A}_{J}=1 /(J+1)$. The reader can easily verify that, on the other hand, $\tilde{B}_{J}=1$ if $V_{J-1} \subset V_{J}$ is strict. Thus, normally the condition of the finite splittings (24) is exactly $\kappa_{J}=J+1$, i.e., exhibits moderate growth in the number of levels.

Further examples will be given later.

## 4 Iterative solvers

In this section we come to some consequences of the notion of stable space splittings for the construction of iterative solution methods for solving variational problems such as (4) and its discretizations. Throughout this section, assume that the splitting (12) is stable. We will use the notation introduced in Section 3. Recall that

$$
a(u, v)=\langle A u, v\rangle, \quad b_{j}\left(u_{j}, v_{j}\right)=\left\langle B_{j} u_{j}, v_{j}\right\rangle_{j}
$$

defines invertible operators $A: V \rightarrow V^{\prime}, B_{j}: V_{j} \rightarrow V_{j}^{\prime}$, and introduce the dual operators $R_{j}^{\prime}: V^{\prime} \rightarrow V_{j}^{\prime}$ by

$$
\left\langle R_{j}^{\prime} \Phi, v_{j}\right\rangle_{j}=\left\langle\Phi, R_{j} v_{j}\right\rangle \quad \forall \Phi \in V^{\prime}, v_{j} \in V_{j}
$$

For given $\Phi \in V^{\prime}$, define $\phi_{j} \in V_{j}$ by solving the auxiliary variational problems

$$
b_{j}\left(\phi_{j}, v_{j}\right)=\Phi\left(R_{j} v_{j}\right)=\left\langle\Phi, R_{j} v_{j}\right\rangle=\left\langle R_{j}^{\prime} \Phi, v_{j}\right\rangle_{j} \quad \forall v_{j} \in V_{j},
$$

$\underset{\sim}{j} \geq 1$. Observe that $\tilde{\phi}=\left(\phi_{j}\right) \in \tilde{V}$ (to this end, represent $\Phi=A u_{\Phi}$ and check that $\left.\tilde{\phi}=R^{*} u_{\Phi}\right)$, and set $\phi=R \tilde{\phi}=\sum_{j} R_{j} \phi_{j}$.

Then we have the following obvious

Theorem 5 The unique solution $u=u_{\Phi} \in V$ of the variational problem (4) (or, what is the same, of the operator equation $A u=\Phi$ in $V^{\prime}$ ) is also the unique solution of the operator equation

$$
\begin{equation*}
\mathcal{P} u=\phi \tag{26}
\end{equation*}
$$

in $V$. Moreover, we have $u=R \tilde{u}$ for any solution $\tilde{u} \in \tilde{V}$ of the operator equation

$$
\begin{equation*}
\tilde{\mathcal{P}} \tilde{u}=\tilde{\phi} \tag{27}
\end{equation*}
$$

in $\tilde{V}$. In addition, we have the representations

$$
\begin{equation*}
\mathcal{P}=\left(\sum_{j} R_{j} B_{j}^{-1} R_{j}^{\prime}\right) A \equiv C A \tag{28}
\end{equation*}
$$

and

$$
\tilde{\mathcal{P}}=\left(\tilde{\mathcal{P}}_{i j}\right), \quad \tilde{\mathcal{P}}_{i j}=B_{i}^{-1} R_{i}^{\prime} A R_{j}: V_{j} \rightarrow V_{i}
$$

Thus, the introduction of a stable space splitting allows us to switch to several equivalent formulations, with the advantage that a proper choice of the space splitting (i.e., of $\left.V_{j}, B_{j}, R_{j}\right)$ leads via Theorem 4 to a well-conditioned operator $\mathcal{P}$. Note that the operator

$$
\begin{equation*}
C=\sum_{j} R_{j} B_{j}^{-1} R_{j}^{\prime} \equiv \sum_{j} \hat{T}_{j}: V^{\prime} \rightarrow V \tag{29}
\end{equation*}
$$

is symmetric with respect to $\langle\cdot, \cdot\rangle$ and can be considered as preconditioner or approximate inverse for $A$. Recall that $T_{j}=\hat{T}_{j} A$ in comparison with (16).

Following this setup, several iterative methods for solving (4) based on auxiliary subproblems associated with the given stable splitting can be introduced and analyzed (see [20, 22, 21, 25, 27]). To avoid discussions about situations which are of no direct practical use, we will assume from now on that the number of spaces $V_{j}$ is finite, i.e., we consider a finite stable splitting

$$
\begin{equation*}
\{V ; a\}=\sum_{j=1}^{J} R_{j}\left\{V_{j} ; b_{j}\right\} \tag{30}
\end{equation*}
$$

(Clearly, real computer implementations will also require finite-dimensionality of the $V_{j}$ ). In their abstract form, the following algorithms have been formulated in [20]:
(AS) Additive Schwarz method. Starting with an initial guess $u^{(0)} \in V$, repeat

$$
u^{(n+1)}=u^{(n)}-\omega \sum_{j=1}^{J}\left(T_{j} u^{(n)}-\phi_{j}\right) \equiv u^{(n)}-\omega \sum_{j=1}^{J} \hat{T}_{j}\left(A u^{(n)}-\Phi\right)
$$

until a stopping criteria is satisfied.
(MS) Multiplicative Schwarz method. Starting with an initial guess $u^{(0)} \in V$, repeat

$$
\begin{aligned}
& v^{(0)}=u^{(n)}, \\
& v^{(j)}=v^{(j-1)}-\omega\left(T_{j} v^{(j-1)}-\phi_{j}\right) \equiv v^{(j-1)}-\omega \hat{T}_{j}\left(A v^{(j-1)}-\Phi\right), \quad j=1, \ldots, J, \\
& \left.u^{(n+1}\right)=v^{(J)},
\end{aligned}
$$

until a stopping criteria is satisfied.
(SMS) Symmetric multiplicative Schwarz method. Starting with an initial guess $u^{(0)} \in V$, repeat

$$
\begin{aligned}
& \left.v^{(0)}=u^{(n}\right), \\
& v^{(j)}=v^{(j-1)}-\omega \hat{T}_{j}\left(A v^{(j-1)}-\Phi\right), \quad j=1, \ldots, J, \\
& v^{(2 J-j+1)}=v^{(2 J-j)}-\omega \hat{T}_{j}\left(A v^{(2 J-j)}-\Phi\right), \quad j=J, \ldots, 1, \\
& \left.u^{(n+1}\right)=v^{(2 J)},
\end{aligned}
$$

until a stopping criteria is satisfied.
Note that the ordering of the subproblems has impact only on the multiplicative methods (MS) and (SMS), and that one iteration step with (SMS) is composed of two iteration steps of (MS) where the second visits the subproblems in reverse order. The relaxation parameter $\omega>0$ has to be chosen appropriately (e.g., in dependence on knowledge on $\tilde{A}, \tilde{B})$. If the choice of $\omega$ seems to be a problem, the symmetry of $C$ can be explored and a preconditioned conjugate gradient method used (cf. [23, Section 9.4.4]):
(SCG) Schwarz-preconditioned conjugate gradient method. Choose an initial $u^{(0)} \in V$, and set

$$
r^{(0)}=A u^{(0)}-\Phi \in V^{\prime}, p^{(0)}=C r^{(0)}=\sum_{j=1}^{J} \hat{T}_{j} r^{(0)} \in V, \rho^{(0)}=\left\langle r^{(0)}, p^{(0)}\right\rangle
$$

Repeat

$$
\begin{aligned}
q & =A p^{(n)} \in V^{\prime}, \quad \lambda=\rho^{(n)} /\left\langle q, p^{(0)}\right\rangle \\
u^{(n+1)} & =u^{(n)}-\lambda p^{(n)} \\
r^{(n+1)} & =r^{(n)}-\lambda q \\
q & =C r^{(n+1)}=\sum_{j=1}^{J} \hat{T}_{j} r^{(n+1)} \in V, \quad \rho^{(n+1)}=\left\langle r^{(n+1)}, q\right\rangle \\
p^{(n+1)} & =q+\left(\rho^{(n)} / \rho^{(n+1)}\right) p^{(n)}
\end{aligned}
$$

until a stopping criteria is satisfied.

A further alternative is to derive from (SMS) a preconditioner to be used within a pcg algorithm (again, the motivation might be to avoid problems with $\omega$ or to improve the convergence rate, i.e., to add robustness to the multiplicative solver).

An elegant way to analyze the above iterations is to rewrite them in terms of classical iterative methods applied to the operator matrix $\tilde{\mathcal{P}}$ (which is now of size $J$ ) as proposed in [26, 27]. Write $\mathcal{P}$ as a sum of strictly lower triangular, diagonal and strictly upper triangular parts

$$
\tilde{\mathcal{P}}=\tilde{\mathcal{L}}+\tilde{\mathcal{D}}+\tilde{\mathcal{U}},
$$

and denote by Id the identity matrix (operator) in $\tilde{V}$. The reader may verify that (with respect to $\{\tilde{V} ; \tilde{a}\}$ )

$$
\tilde{\mathcal{D}}^{*}=\tilde{\mathcal{D}}, \tilde{\mathcal{U}}^{*}=\tilde{\mathcal{L}} \Longleftrightarrow b_{i}\left(\tilde{\mathcal{P}}_{i j} u_{j}, v_{i}\right)=b_{j}\left(u_{j}, \tilde{\mathcal{P}}_{j i} v_{i}\right) \quad \forall u_{j} \in V_{j}, v_{i} \in V_{i} .
$$

In the considerations below, we set $u^{(n)}=R \tilde{u}^{(n)}, n \geq 0$, and consider a linear iteration scheme in $\tilde{V}$ (which is now a space of $J$-vectors $\tilde{v}=\left(v_{1}, \ldots, v_{J}\right)^{T}$ ) to generate the sequence $\left(\tilde{u}^{(n)}\right)$ :

$$
\begin{equation*}
\tilde{u}^{(n+1)}=\tilde{u}^{(n)}-\tilde{\mathcal{N}}\left(\tilde{\mathcal{P}} \tilde{u}^{(n)}-\tilde{\phi}\right), \quad n \geq 0 . \tag{31}
\end{equation*}
$$

Lemma 6 a) Damped Richardson iteration: If $\tilde{\mathcal{N}} \equiv \tilde{\mathcal{N}}_{A S}=\omega \tilde{\mathrm{Id}}$ then the iteration (31) is equivalent to (AS).
b) Damped Jacobi iteration: Assume that $\tilde{\mathcal{D}}$ is invertible in $\tilde{V}$. If $\tilde{\mathcal{N}} \equiv \omega \tilde{\mathcal{D}}^{-1}$ then the iteration (31) is equivalent to (AS) for the modified stable splitting

$$
\{V ; a\}=\sum_{j=1}^{J} R_{j}\left\{V_{j} ; a\left(R_{j} \cdot, R_{j} \cdot\right)\right\}
$$

c) Richardson-SOR method: If $\tilde{\mathcal{N}} \equiv \tilde{\mathcal{N}}_{M S}=\left(\frac{1}{\omega} \tilde{\mathrm{Id}}+\tilde{\mathcal{L}}\right)^{-1}$ then the iteration (31) is equivalent to (MS). In analogy to part b) above, the original SOR-method defined by $\tilde{\mathcal{N}}=\left(\frac{1}{\omega} \tilde{\mathcal{D}}+\tilde{\mathcal{L}}\right)^{-1}$ is a particular case.
d) Richardson-SSOR method: If

$$
\tilde{\mathcal{N}} \equiv \tilde{\mathcal{N}}_{S M S}=\left(\frac{1}{\omega} \tilde{\mathrm{Id}}+\tilde{\mathcal{U}}\right)^{-1}\left(\frac{2}{\omega} \tilde{\mathrm{Id}}-\tilde{\mathcal{D}}\right)\left(\frac{1}{\omega} \tilde{\mathrm{Id}}+\tilde{\mathcal{L}}\right)^{-1}
$$

then the iteration (31) is equivalent to (SMS).
Proof. We give some hints, details of the algebraic computations are left upon the reader. To see the stated equivalences, one has to compare the iteration operator

$$
M=\operatorname{Id}-R \tilde{\mathcal{N}} R^{*}
$$

for the sequence $u^{(n)}$ in $V$ resulting from a certain choice of $\tilde{\mathcal{N}}$ with the iteration operator of the corresponding abstract method.
a) Here,

$$
M_{A S}=\operatorname{Id}-\omega R R^{*}=\operatorname{Id}-\omega \mathcal{P}
$$

Now, compare with the definition of (AS).
b) $\tilde{\mathcal{D}}$ is invertible in $\tilde{V}$ if (and only if)

$$
\begin{equation*}
\omega_{0} b_{j}\left(u_{j}, u_{j}\right) \leq \tilde{b}_{j}\left(u_{j}, u_{j}\right) \equiv a\left(R_{j} u_{j}, R_{j} u_{j}\right) \leq \omega_{1} b_{j}\left(u_{j}, u_{j}\right) \quad \forall u_{j} \in V_{j}, j=1, \ldots, J, \tag{32}
\end{equation*}
$$

holds with two constants $0<\omega_{0} \leq \omega_{1}<\infty$. The upper estimate is obvious from (12): $\omega_{1} \leq 1 / \tilde{A}$. The lower estimate does not follow from (12) and needs to be assumed (if all $V_{j}$ are finite-dimensional it is certainly true with some positive $\omega_{0}$ ). By the twosided estimate in (32) and the definition of stable splittings it follows that the modified splitting is also stable (with other $\tilde{A}, \tilde{B}$ ). Check that the Schwarz operator associated with this new splitting takes the form $R \tilde{\mathcal{D}}^{-1} R^{*}$, and compare with the result of part a). c) First of all, since $\tilde{\mathcal{L}}$ is strictly lower triangular, we can compute $\tilde{\mathcal{N}}_{M S}$ by the formula

$$
\tilde{\mathcal{N}}_{M S}=\left(\frac{1}{\omega} \tilde{\mathrm{Id}}+\tilde{\mathcal{L}}\right)^{-1}=\omega \sum_{k=0}^{J-1}(-\omega \tilde{\mathcal{L}})^{k} .
$$

Thus,

$$
M_{M S}=\mathrm{Id}+\sum_{k=1}^{J}(-\omega)^{k} R \tilde{\mathcal{L}}^{k-1} R^{*}
$$

which has to be compared with the iteration matrix for (MS):

$$
\left(\operatorname{Id}-\omega T_{J}\right) \ldots\left(\operatorname{Id}-\omega T_{1}\right)=\operatorname{Id}+\sum_{k=1}^{J}(-\omega)^{k} \sum_{J \geq j_{k}>\ldots>j_{1} \geq 1} T_{j_{k}} \ldots T_{j_{1}} .
$$

Now, prove by induction that the coefficients in the two operator polynomials (with respect to the variable $(-\omega)$ ) coincide. For $k=1$ this is certainly true since

$$
\sum_{j=1}^{J} T_{j}=\mathcal{P}=R R^{*}
$$

(this also shows that the 'linear' part of the iteration (MS) coincides with the iteration operator $M_{A S}$ of (AS)). For $k=2$,

$$
R \tilde{\mathcal{L}} R^{*}=\sum_{i=1}^{J} R_{i} \sum_{j=1}^{i-1}\left(R_{i}^{*} R_{j}\right) R_{j}^{*}=\sum_{1 \leq j<i \leq J} T_{i} T_{j},
$$

and so on. Compare [25, p.78].
d) As in part c), one has

$$
M_{M S}^{*}=\left(\operatorname{Id}-\omega T_{1}\right) \ldots\left(\operatorname{Id}-\omega T_{J}\right)=\operatorname{Id}-R\left(\frac{1}{\omega} \tilde{\operatorname{Id}}+\tilde{\mathcal{U}}\right)^{-1} R^{*}
$$

Therefore, we have to check that

$$
\begin{aligned}
M_{M S}^{*} M_{M S} & =\left(\operatorname{Id}-R\left(\frac{1}{\omega} \tilde{\operatorname{Id}}+\tilde{\mathcal{U}}\right)^{-1} R^{*}\right)\left(\operatorname{Id}-R\left(\frac{1}{\omega} \tilde{\operatorname{Id}}+\tilde{\mathcal{L}}\right)^{-1} R^{*}\right) \\
& =\operatorname{Id}-R\left(\frac{1}{\omega} \tilde{\operatorname{Id}}+\tilde{\mathcal{U}}\right)^{-1}\left(\frac{2}{\omega} \tilde{\operatorname{Id}}+\tilde{\mathcal{L}}+\tilde{\mathcal{U}}-\tilde{\mathcal{P}}\right)\left(\frac{1}{\omega} \tilde{\operatorname{Id}}+\tilde{\mathcal{L}}\right)^{-1} R^{*} \\
& =\operatorname{Id}-R \tilde{\mathcal{N}}_{M S M} R^{*}=M_{S M S}
\end{aligned}
$$

This completes the proof of the lemma.
We can now state the main result of this section. Let us denote the asymptotic convergence rate of an iterative method governed by an iteration operator $M$ by its spectral radius:

$$
\rho=\lim _{n \rightarrow \infty}\left\|M^{n}\right\|^{1 / n}
$$

Below, we choose the operator norm $\|\cdot\|_{a}$ in $\{V ; a\}$ in this definition. If $M$ is symmetric $\{V ; a\}$ then the above formula simplifies to

$$
\rho=\|M\|_{a}=\sup _{a(u, u) \leq 1} a(M u, u) .
$$

This simplifying argument applies to (AS) and (SMS) directly, for (MS) we use the obvious relationship

$$
\rho_{M S} \leq\left\|M_{M S}\right\|_{a}=\sqrt{\left\|M_{M S}^{*} M_{M S}\right\|_{a}}=\sqrt{\left\|M_{S M S}\right\|_{a}}=\rho_{S M S}^{1 / 2} .
$$

Theorem 7 Let (12) be a stable space splitting into finitely many auxiliary spaces $V_{j}$, $j=1, \ldots, J$, with stability constants $\tilde{A}, \tilde{B}$, and condition $\kappa$. condition $\kappa$.
a) The additive method (AS) converges for $0<\omega<2 \tilde{A}$, at rate

$$
\rho_{A S, \omega}=\max \{|1-\omega / \tilde{A}|,|1-\omega / \tilde{B}|\}
$$

The optimal convergence rate is achieved for $\omega^{*}=2 \tilde{A} \tilde{B} /(\tilde{A}+\tilde{B})$ :

$$
\begin{equation*}
\rho_{A S}^{*}=\rho_{A S, \omega^{*}}=\inf _{0<\omega<2 / \tilde{A}} \rho_{A S, \omega}=1-\frac{2}{1+\kappa} . \tag{33}
\end{equation*}
$$

For the related method (SCG), we have the estimate

$$
\begin{equation*}
\left\|u-u^{(n)}\right\|_{a} \leq 2\left(1-\frac{2}{1+\sqrt{\kappa}}\right)^{n}\left\|u-u^{(0)}\right\|_{a}, \quad n \geq 1 \tag{34}
\end{equation*}
$$

for the guaranteed error reduction in the energy norm.
b) For the multiplicative algorithms (MS) and (SMS), convergence is guaranteed if $0<\omega<2 / \omega_{1}$, where $\omega_{1} \leq 1 / \tilde{A}$ is defined by (32). The convergence rate can be estimated by

$$
\rho_{M S, \omega}^{2} \leq \rho_{S M S} \leq 1-\frac{\omega\left(2-\omega \omega_{1}\right)}{\tilde{B}\|\tilde{\mathrm{Id}}+\omega \tilde{\mathcal{L}}\|_{\tilde{a}}^{2}}
$$

The optimal rates satisfy

$$
\begin{equation*}
\left(\rho_{M S, \omega}^{*}\right)^{2} \leq \rho_{S M S}^{*} \leq 1-\frac{1}{\tilde{B}\left(2\|\tilde{\mathcal{L}}\|_{\tilde{a}}+\omega_{1}\right)} \tag{35}
\end{equation*}
$$

If no additional assumptions on the splitting are available, we can use

$$
\begin{equation*}
\|\tilde{\mathcal{L}}\|_{a} \leq \frac{\left[\log _{2}(2 J)\right]}{2 \tilde{A}} \tag{36}
\end{equation*}
$$

to arrive at

$$
\begin{equation*}
\left(\rho_{M S, \omega}^{*}\right)^{2} \leq \rho_{S M S}^{*} \leq 1-\frac{1}{\log _{2}(4 J) \cdot \kappa} . \tag{37}
\end{equation*}
$$

Proof (see [27]) for more details). Part a) is just a repetition of the convergence rate estimates for the Richardson resp. conjugate gradient iteration applied to the equation (26) with the spd operator $\mathcal{P}$ (sharp spectral bounds for the latter are contained in Theorem 4). The reader is recommended to look at the classical results in [23, Sections 4.4 and 9.4].

We concentrate on part b). Together with the above observations on computing convergence rates $\rho$, Lemma 6 allows us to reduce the statements on the convergence behavior for the multiplicative Schwarz methods to norm estimates in $\tilde{V}$ for the case of the symmetric iteration (SMS). Indeed,

$$
\begin{aligned}
a\left(M_{S M S} u, u\right) & =a(u, u)-a\left(R \tilde{\mathcal{N}}_{M S M} R^{*} u, u\right) \\
& =a(u, u)-\tilde{a}\left(\left(\frac{2}{\omega} \tilde{\operatorname{Id}}-\tilde{\mathcal{D}}\right) \tilde{w}, \tilde{w}\right) \\
& \leq a(u, u)-\left(\frac{2}{\omega}-\omega_{1}\right) \tilde{a}(\tilde{w}, \tilde{w}), \quad \tilde{w}=\left(\frac{1}{\omega} \tilde{\operatorname{Id}}+\tilde{\mathcal{L}}\right)^{-1} R^{*} u
\end{aligned}
$$

Here, we have used Lemma 6 d$),\left(\left(\frac{1}{\omega} \tilde{\mathrm{Id}}+\tilde{\mathcal{U}}\right)^{-1}\right)^{*}=\left(\frac{1}{\omega} \mathrm{Id}+\tilde{\mathcal{L}}\right)^{-1}$, and $\tilde{\mathcal{D}} \geq \omega_{1} \mathrm{Id}$ (see (32)). Now we continue with

$$
\tilde{a}(\tilde{w}, \tilde{w}) \geq\left\|\frac{1}{\omega} \tilde{\operatorname{Id}}+\tilde{\mathcal{L}}\right\|_{\tilde{a}}^{-2} \tilde{a}\left(R^{*} u, R^{*} u\right) \geq \frac{\omega^{2}}{\left(1+\omega\|\tilde{\mathcal{L}}\|_{\tilde{a}}\right)^{2}} \tilde{a}\left(R^{*} u, R^{*} u\right)
$$

and

$$
\tilde{a}\left(R^{*} u, R^{*} u\right)=a(\mathcal{P} u, u) \geq \frac{1}{\tilde{B}} a(u, u) .
$$

Putting things together, we get the estimate for $\rho_{S M S, \omega}$. Minimization with respect to $0<\omega<2 / \omega_{1}$ gives (35).

The proof of (36) is given in detail in [27, p. 174-176], and will not be repeated here. The estimate (37) follows if (36) and $\omega_{1} \leq 1 / \tilde{A}$ are substituted into (35). This concludes the proof of Theorem 7.

The above theorem leads to a number of questions. First, from the estimates it becomes clear that the multiplicative algorithms are never much worse than the additive Schwarz method. That the logarithmical factor $\log _{2}(4 J) \asymp \log _{2} J$ is sharp shows the following example [29] of a sequence of stable space splittings. The splittings are (in the spirit of Example 5) inherited from a sequence of real spd matrices $A_{n}=L_{n}+D_{n}+L_{n}^{T}$, where $D_{n}=\mathrm{Id}_{n}$. The spectrum of the $A_{n}$ is uniformly bounded away from 0 and $\infty$ :

$$
0<\lambda_{\min } \leq \lambda_{\min }\left(A_{n}\right) \leq \lambda_{\max }\left(A_{n}\right) \leq \lambda_{\max }<\infty
$$

On the other hand, there exist unit vectors $x_{n} \in \mathbb{C}^{n}$ such that for a sequence of real numbers $\alpha_{n} \asymp \log _{2} n$ the equality $\left(L_{n} x_{n}, x_{n}\right)=i \alpha_{n}$ holds for $n \rightarrow \infty$ (throughout this example, $(\cdot, \cdot)$ denotes the scalar product in $\left.\mathbb{C}^{n}\right)$.

For this matrix setting, it is easy to see (compare Example 5 and [23]) that (SMS) coincides with the usual SSOR-method for $A_{n}$. By the above assumptions, using the identity

$$
\left\|M_{S M S}\right\|_{a}=1-\frac{2 / \omega-1}{\left\|A_{n}^{-1 / 2}\left(1 / \omega \operatorname{Id}_{n}+L_{n}\right)\right\|^{2}}, \quad n \geq 1
$$

(see [23, Lemma 4.4.23]), one is lead to estimating

$$
\left\|A_{n}^{-1 / 2}\left(1 / \omega \operatorname{Id}_{n}+L_{n}\right)\right\|^{2} \geq \lambda_{\max }^{-1}\left\|1 / \omega \operatorname{Id}_{n}+L_{n}\right\|^{2} \geq \lambda_{\max }^{-1} r\left(1 / \omega \operatorname{Id}_{n}+L_{n}\right)^{2}
$$

where $r(B)$ denotes the numerical radius of $B$ ([23, Section 2.9.5]). Using the vectors $x_{n}$, we see that

$$
r\left(1 / \omega \operatorname{Id}_{n}+L_{n}\right)=\sup _{0 \neq x} \frac{\left|\left(\left(1 / \omega \operatorname{Id}_{n}+L_{n}\right) x, x\right)\right|}{(x, x)} \geq\left|\frac{1}{\omega}+i \alpha_{n}\right|
$$

which yields

$$
\left\|A^{-1 / 2}\left(1 / \omega \operatorname{Id}_{n}+L_{n}\right)\right\|^{2} \geq c\left(\omega^{-2}+\left(\log _{2} n\right)^{2}\right), \quad n \rightarrow \infty
$$

for some positive constant $c$. Now substitute into the formula for $\left\|M_{S M S}\right\|_{a}$ and take the infimum with respect to $0<\omega<2$.

A family of matrices which have the above properties are the finite sections

$$
A_{n}=\left(a_{i j}=a_{i-j}: i, j=1, \ldots, n\right), \quad a_{k}= \begin{cases}1 & k=0 \\ \frac{c_{0} \sin (k \pi / 2)}{k} & , \quad k \neq 0\end{cases}
$$

of a $\ell^{2}$-bounded and boundedly $\ell^{2}$-invertible Toeplitz operator $A_{\infty}\left(\left|c_{0}\right|<2 / \pi\right)$. The auxiliary vectors are given by

$$
x_{n}=\frac{1}{\sqrt{n}}\left(i, i^{2}, \ldots, i^{n}\right), \quad n \geq 1
$$

The reader may consult [29] for details and references on Toeplitz matrices. Challenge: The above argument does not show that the convergence rate for (MS) deteriorates in the same manner (even though numerical evidence provided in [29] shows that this is the case). Is there an argument to close this little gap?

On the other hand, in most of the classical cases it is well known that SOR-like methods outperform the simpler Richardson iteration. Artificial examples show that the multiplicative algorithms may be arbitrarily fast compared to the additive method (AS). Many sufficient conditions for convergence of multiplicative schemes have been discussed (see, e.g., $[20,22,21]$ ). Some of them are, at the same time, methods to establish the stability of a given splitting. The reader is recommended to learn from the above sources about the so-called strengthened Cauchy-Schwarz inequalities associated with a splitting which, in their simplest form,

$$
\begin{equation*}
a\left(R_{i} u_{i}, R_{j} v_{j}\right) \leq \gamma_{i j} \sqrt{b_{i}\left(u_{i}, u_{i}\right)} \sqrt{b_{j}\left(v_{j}, v_{j}\right)} \quad \forall u_{i} \in V_{i}, v_{j} \in V_{j} \tag{38}
\end{equation*}
$$

are equivalent to estimates for the norms of the operator entries $\tilde{\mathcal{P}}_{i j}$ of the extended Schwarz operator $\tilde{\mathcal{P}}$. Indeed, (38) is equivalent to $\left\|\tilde{\mathcal{P}}_{i j}\right\|_{V_{j} \rightarrow V_{i}}=\left\|\tilde{\mathcal{P}}_{j i}\right\|_{V_{i} \rightarrow V_{j}} \leq \gamma_{i j}$. Exercise: Consider in detail the case $J=2$ and give the best possible estimates. This would also cover the two iterative methods given in connection with Example 6.

The theory has been developed so far exclusively for linear spd problems. This is certainly a drawback. Attempts to slightly remove these assumptions will be discussed later.

## 5 Modifications of space splittings

This very short section emphasizes the fact that abstract concepts could pay off if used in a systematic way. Due to the theorems of the previous two sections we now know that various iterative solution schemes are governed by the availability of two-sided stability estimates such as (13) for a space splitting (12). This makes it possible to introduce some general operations on stable space splittings which allow us to modify a given one, in order to adapt it to a specific application or to optimize the implementation with respect to a given hardware platform. We list a few of them.

Refinement and clustering. The method of refinement consists in splitting in a given stable splitting (say, in (30)) the auxiliary spaces further:

$$
\begin{equation*}
\left\{V_{j} ; b_{j}\right\}=\sum_{k=1}^{I_{j}} R_{j k}\left\{V_{j k} ; b_{j k}\right\}, \quad j=1, \ldots, J \tag{39}
\end{equation*}
$$

Now, denote the stability constants of the $j$-th splitting in (39) by $\tilde{A}_{j}, \tilde{B}_{j}$. Then the refined splitting

$$
\begin{equation*}
\{V ; a\}=\sum_{j=1}^{J} \sum_{k=1}^{I_{j}} R_{j} R_{j k}\left\{V_{j k} ; b_{j k}\right\}, \quad j=1, \ldots, J \tag{40}
\end{equation*}
$$

again satisfies the stability definition, with constants

$$
\tilde{A}_{r e f} \geq \tilde{A} \min _{j=1, \ldots, J} \tilde{A}_{j}, \quad \tilde{B}_{r e f} \leq \tilde{B} \max _{j=1, \ldots, J} \tilde{B}_{j}
$$

Even though these are very rough estimates (not equalities for the optimal stability constants), they give first hints on which modifications are useful. If $I_{j}=1$, the statement tells us that spectrally equivalent changes of the bilinear forms in the spaces $V_{j}$ are admissible as long as they are uniformly tight and correctly scaled. This was implicitely used in the argument for Lemma 6, b).

Clustering is the inverse operation. Given a stable splitting into many component spaces (such as (40)), we will build several groups which we replace then by a single auxiliary space. Let us assume that the grouping is according to (39). Then the original
splitting turns into the clustered splitting of (40). Having access to the stabilty constants of the splittings (40) and (39) we get now

$$
\begin{aligned}
& \tilde{A} \equiv\left(\tilde{A}_{r e f}\right)_{c l u s t} \geq \tilde{A}_{r e f}\left(\max _{j=1, \ldots, J} \tilde{B}_{j}\right)^{-1} \\
& \tilde{B} \equiv\left(\tilde{B}_{r e f}\right)_{c l u s t} \leq \tilde{B}_{r e f}\left(\min _{j=1, \ldots, J} \tilde{A}_{j}\right)^{-1}
\end{aligned}
$$

Clustering is very useful if one starts with frame decompositions and clusters them such that larger spaces with certain properties result. Load balancing on parallel architectures with a small to moderate number of processors might be one motivation. Below we will show some applications of this technique. The refinement/clustering schemes are schematically illustrated in Figure 3.


Figure 3: Refinement and clustering
Selection is a critical but very important procedure. It is visualized in Figure 4. Each space $V_{j}$ in (30) is replaced by its subspace $\bar{V}_{j}$ while the bilinear form is preserved by restriction (taking the trivial subspace $\{0\}$ is allowed, in this case $V_{j}$ is simply removed from the original splitting and the index $j$ may be dropped from the set of indices). Consider the subspace

$$
\bar{V}=\sum_{j} R_{j} \bar{V}_{j}
$$

in $V$. We may hope for having produced a new stable selected space splitting from (30):

$$
\begin{equation*}
\{\bar{V} ; a\}=\sum_{j=1}^{J} R_{j}\left\{\bar{V}_{j} ; b_{j}\right\} \tag{41}
\end{equation*}
$$

The difficulty with this procedure is that it is almost equivalent to selecting subsplittings of a given splitting which can make, in analogy with the situation for frames, the
condition of a splitting much worse. However, in the multilevel applications (as was shown in Example 7) some of the most natural selections (the finite sections of an infinite splitting) are uniformly stable. Further examples of the selection procedure will be discussed later, in connection with adaptivity (see also [25, Section 4.2.2], be aware of inconsistencies there). Recall again that for subsplittings of stable splittings into direct sums of auxiliary spaces a deterioration of condition numbers is not possible.


Figure 4: Selection
Let us finally mention that we have succesfully used tensor products of stable space splittings of univariate schemes in connection with sparse grid spaces and anisotropic constant coefficient operators in higher dimensions [28]. With the above techniques, a lot of practical situations can be covered (i.e., seemingly independent applications can be derived from one and the same basic splitting). However, characterizations of all useful, i.e., stability-preserving, transformations of stable splittings and frames have not yet been studied in a general and systematic way.

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