# Friedrichs systems 

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WeConMApp


Why should one be interested in Friedrichs systems?
Symmetric hyperbolic systems
Symmetric positive systems
Classical theory
Boundary conditions for Friedrichs systems
Existence, uniqueness, well-posedness
Abstract formulation
Graph spaces
Cone formalism of Ern, Guermond and Caplain
Interdependence of different representations of boundary conditions
Kreĭn space formalism
Kreĭn spaces
Equivalence of boundary conditions
What can we say for the Friedrichs operator now?
Sufficient assumptions
An example: elliptic equation
Other second order equations
Two-field theory
Non-stationary theory
Homogenisation of Friedrichs systems
Homogenisation
Examples: Stationary diffusion and heat equation
Concluding remarks

Friedrichs' system (KOF1958)

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\begin{equation*}
\left(\exists \mu_{0}>0\right) \quad \mathbf{C}+\mathbf{C}^{\top}+\sum_{k=1}^{d} \partial_{k} \mathbf{A}_{k} \geqslant 2 \mu_{0} \mathbf{I} \quad(\text { ae on } \Omega) \tag{F2}
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The operator $\mathcal{L}: \mathrm{L}^{2}\left(\Omega ; \mathbf{R}^{r}\right) \longrightarrow \mathcal{D}^{\prime}\left(\Omega ; \mathbf{R}^{r}\right)$

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$$
\mathcal{L} \mathrm{u}=\mathrm{f}
$$

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## Symmetric hyperbolic systems (KOF1954)

## Summing over repeated indices:

$$
\mathrm{A}^{k} \partial_{k} \mathrm{u}+\mathrm{Bu}=\mathrm{f}
$$

In divergence form:

$$
\partial_{k}\left(\mathbf{A}^{k} \mathbf{u}\right)+\left(\mathbf{B}-\partial_{k} \mathbf{A}^{k}\right) \mathbf{u}=\mathbf{f} .
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$$

It is symmetric if all matrices $\mathbf{A}^{k}$ are symmetric; and hyperbolic (Friedrichs) if one of the matrices is even positive definite.

## The wave equation

In $d$-dimensional space:

$$
\left(\rho u^{\prime}\right)^{\prime}-\operatorname{div}(\mathbf{A} \nabla u)=g .
$$

Time $t=x^{0}$ and $\partial_{0}:=\frac{\partial}{\partial t}$ :
(*)

$$
\partial_{0}\left(\rho \partial_{0} u\right)-\sum_{i, j=1}^{d} \partial_{i}\left(a^{i j} \partial_{j} u\right)=g .
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New variables: $v_{j}:=\partial_{j} u, j \in 0 . . d$ give vector unknown $u=\left[u, v_{0}, \ldots, v_{d}\right]^{\top}$, and with: $a^{00}:=-\rho, a^{0 i}:=a^{i 0}:=0$ we have

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\end{equation*}
$$

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$$

This transformation gives us only one equation. For a system with $d+2$ unknowns to be formally deterministic, we need $d+1$ more equations. Clearly, defining equations for $v^{i}$ would lead to a formally deterministic system, which is not symmetric.

The wave equation (cont.)

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This will be the second equation of the system.

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The remaining $d$ equations will be the Schwarz symmetry relations, with one index being 0 , but multiplied by $\mathbf{A}^{\top}$ :

$$
\begin{aligned}
\partial_{0} u-v_{0} & =0 \\
\rho \partial_{0} v_{0}-a^{i j} \partial_{i} v_{j}+b^{j} v_{j} & =g \\
a^{i j} \partial_{0} v_{i}-a^{i j} \partial_{i} v_{0} & =0
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where $b^{0}:=\partial_{0} \rho, b^{j}:=-\partial_{i} a^{i j}=\left[-\operatorname{div} \mathbf{A}^{\top}\right]^{j}$, for $j \in 1 . . d$.

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where $b^{0}:=\partial_{0} \rho, b^{j}:=-\partial_{i} a^{i j}=\left[-\operatorname{div} \mathbf{A}^{\top}\right]^{j}$, for $j \in 1 . . d$.
Actually, we can take $v_{0}=\partial_{0} u$ as a definition of $u$, and solve first for the remaining unknowns.

The wave equation in the required form

$$
\begin{gathered}
{\left[\begin{array}{cccc}
\rho & 0 & \cdots & 0 \\
0 & & & \\
\vdots & \mathbf{A}^{\top} & \\
0 &
\end{array}\right] \partial_{0} \mathbf{u}+\sum_{i=1}^{d}\left[\begin{array}{cccc}
0 & -a^{i 1} & \cdots & -a^{i n} \\
-a^{i 1} & & & \\
\vdots & & \mathbf{0} &
\end{array}\right] \partial_{i} \mathbf{u}} \\
-a^{i n} \\
\\
+\left[\begin{array}{cccc}
b^{0} & b^{1} & \cdots & b^{n} \\
0 & & & \\
\vdots & \mathbf{0} & \\
0 & &
\end{array}\right] \mathbf{u}=\left[\begin{array}{c}
g \\
0 \\
\vdots \\
0
\end{array}\right]
\end{gathered}
$$

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$$
\left[\begin{array}{cccc}
\rho & 0 & \cdots & 0 \\
0 & & & \\
\vdots & & \mathbf{A}^{\top} & \\
0 & & \partial_{0} \mathbf{u}+ & \sum_{i=1}^{d}\left[\begin{array}{cccc}
0 & -a^{i 1} & \cdots & -a^{i n} \\
-a^{i 1} & & & \\
\vdots & & \mathbf{0} & \\
-a^{i n} & &
\end{array}\right] \partial_{i} \mathbf{u} \\
& +\left[\begin{array}{cccc}
b^{0} & b^{1} & \cdots & b^{n} \\
0 & & \\
\vdots & \mathbf{0} \\
0 & &
\end{array}\right] \mathbf{u}=\left[\begin{array}{c}
g \\
0 \\
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0
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\end{array}\right.
$$

$\mathbf{A}^{i}$ are symmetric, $\mathbf{A}^{0}$ is even positive definite ( $\rho>0$ and $\mathbf{A}$ is p.d.). In particular, the system to which we reduced the wave equation is hyperbolic in the sense of Petrovski.

The wave equation (cont.)

For initial data $u(0,)=.u_{0}$ and $u^{\prime}(0,)=.u_{1}$, take:

$$
\begin{aligned}
(u(0, .) & \left.=u_{0} \quad\right) \\
\partial_{0} u(0, .) & =u_{1} \\
\partial_{i} u(0, .) & =\partial_{i} u_{0} \quad, \text { for } i \in 1 . . d
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as the initial data for the system.
$u_{0}$ is defined on $\mathbf{R}^{d}$, so we can compute its derivatives in the spatial directions.

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To check:
the identities defining $v_{i}$ (and therefore the symmetry relations).
For $i \in 1 . . d$ :

$$
\partial_{0} v_{i}=\partial_{i} v_{0}=\partial_{i} \partial_{0} u=\partial_{0} \partial_{i} u
$$

(The first equality follows from the regularity of $\mathbf{A}^{\top}$, because
$\mathbf{A}^{\top}\left(\partial_{0} \vee-\nabla v_{0}\right)=0$ implies $\left.\partial_{0} v_{i}=\partial_{i} v_{0}.\right)$
Now, we have that $\partial_{0}\left(v_{i}-\partial_{i} u\right)=0$, and $v_{i}-\partial_{i} u=0$ at $t=0$, and we conclude that the last identity holds for any $t>0$.

## Maxwell's systems

In a material with electric permeability $\epsilon$, conductivity $\sigma$ and magnetic susceptibility $\boldsymbol{\mu}$

$$
\begin{aligned}
& \mathrm{D}^{\prime}=\operatorname{rot} \mathrm{H}-\mathrm{J}+\mathrm{F} \\
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together with $\operatorname{div} D=\rho$ and $\operatorname{div} B=0$, and with the constitutive laws:

$$
\begin{aligned}
\mathrm{D}(., t) & =\epsilon \mathrm{E}(., t) \\
\mathrm{J}(., t) & =\sigma \mathrm{E}(., t) \\
\mathrm{B}(., t) & =\boldsymbol{\mu} \mathrm{H}(., t) .
\end{aligned}
$$

## Maxwell's systems (cont.)

$E$ and $H$ as variables, $u:=\left[\begin{array}{l}E \\ H\end{array}\right]$, the system can be written in the form of a symmetric system:

$$
\sum_{i=0}^{3} \mathbf{A}^{i} \partial_{i} \mathbf{u}+\mathbf{B u}=\mathbf{f}
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where:

$$
\mathbf{A}^{0}=\left[\begin{array}{ll}
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\end{array}\right], \mathbf{A}^{1}:=\left[\begin{array}{cc}
\mathbf{0} & \mathbf{Q}_{1}^{\top} \\
\mathbf{Q}_{1} & \mathbf{0}
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\mathbf{0} & \mathbf{Q}_{3}^{\top} \\
\mathbf{Q}_{3} & \mathbf{0}
\end{array}\right]
$$

The constant antisymmetric matrices $\mathbf{Q}_{k}$ are given by:

$$
\mathbf{Q}_{1}:=\left[\begin{array}{ccc}
0 & 0 & 0 \\
0 & 0 & -1 \\
0 & 1 & 0
\end{array}\right], \mathbf{Q}_{2}:=\left[\begin{array}{ccc}
0 & 0 & 1 \\
0 & 0 & 0 \\
-1 & 0 & 0
\end{array}\right], \mathbf{Q}_{3}:=\left[\begin{array}{ccc}
0 & -1 & 0 \\
1 & 0 & 0 \\
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\end{array}\right] .
$$

## Maxwell's systems (cont.)

$$
\mathbf{B}=\left[\begin{array}{ll}
\boldsymbol{\sigma} & \mathbf{0} \\
\mathbf{0} & \mathbf{0}
\end{array}\right], \quad \text { while the right hand side is } \mathrm{f}=\left[\begin{array}{l}
\mathrm{F} \\
\mathrm{G}
\end{array}\right] .
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In the above we have used the fact that the rotator (curl) of a vector field E can be written as:

$$
\begin{aligned}
\operatorname{rot} \mathrm{E}=\left[\begin{array}{c}
\partial_{2} E^{3}-\partial_{3} E^{2} \\
\partial_{3} E^{1}-\partial_{1} E^{3} \\
\partial_{1} E^{2}-\partial_{2} E^{1}
\end{array}\right] & =\left[\begin{array}{ccc}
0 & 0 & 0 \\
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0 & 1 & 0
\end{array}\right] \partial_{1} \mathrm{E} \\
& +\left[\begin{array}{ccc}
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0 & 0 & 0 \\
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\end{array}\right] \partial_{2} \mathrm{E}+\left[\begin{array}{ccc}
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\end{aligned}
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If we assume the uniform boundedness and symmetry of the permeability and susceptibility tensors, the above system is even symmetric hyperbolic.

Friedrichs systems

Introduced in:
K. O. Friedrichs: Symmetric positive linear differential equations, Communications on Pure and Applied Mathematics 11 (1958), 333-418

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Goal:

- treating the equations of mixed type, such as the Tricomi equation:

$$
y \frac{\partial^{2} u}{\partial x^{2}}+\frac{\partial^{2} u}{\partial y^{2}}=0
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- unified treatment of equations and systems of different type.
- still it does not cover all of Gårding's theory of general elliptic equations, or

Lerray's of general hyperbolic equations.

Example - heat equation, first form

Heat equation with lower order terms ( $\Omega \subseteq \mathbf{R}^{d}, T>0$ and $\Omega_{T}:=\langle 0, T\rangle \times \Omega$ ):

$$
\partial_{t} u-\operatorname{div}(\mathbf{A} \nabla u)+\mathbf{b} \cdot \nabla u+c u=f \quad \text { in } \Omega_{T}
$$

where $f \in \mathrm{~L}^{2}\left(\Omega_{T}\right), c \in \mathrm{~L}^{\infty}\left(\Omega_{T}\right), \mathbf{b} \in \mathrm{L}^{\infty}\left(\Omega_{T} ; \mathbf{R}^{d}\right)$ and $\mathbf{A} \in \mathrm{L}^{\infty}\left(\Omega_{T} ; \mathrm{M}_{d}(\mathbf{R})\right)$ is symmetric with eigenvalues between $\alpha>0$ and $\beta \geqslant \alpha$ a.e. on $\Omega_{T}$.

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Similarly as for the wave equation: $\mathbf{A}=\left[a^{1} \cdots a^{d}\right]$ and $w=\nabla u$

$$
\left[\begin{array}{cc}
1 & 0^{\top} \\
0 & \mathbf{0}
\end{array}\right] \partial_{t}\left[\begin{array}{c}
u \\
\mathbf{w}
\end{array}\right]-\sum_{i=1}^{d}\left[\begin{array}{cc}
\operatorname{div}^{i} a^{i} & \left(\mathrm{a}^{i}\right)^{\top} \\
\mathrm{a}^{i} & \mathbf{0}
\end{array}\right] \partial_{x^{i}}\left[\begin{array}{l}
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c & \mathrm{~b}^{\top} \\
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\end{array}\right]\left[\begin{array}{l}
u \\
\mathrm{w}
\end{array}\right]=\left[\begin{array}{l}
f \\
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\end{array}\right] .
$$

## Example - heat equation, first form

Heat equation with lower order terms ( $\Omega \subseteq \mathbf{R}^{d}, T>0$ and $\Omega_{T}:=\langle 0, T\rangle \times \Omega$ ):

$$
\partial_{t} u-\operatorname{div}(\mathbf{A} \nabla u)+\mathbf{b} \cdot \nabla u+c u=f \quad \text { in } \Omega_{T},
$$

where $f \in \mathrm{~L}^{2}\left(\Omega_{T}\right), c \in \mathrm{~L}^{\infty}\left(\Omega_{T}\right), \mathbf{b} \in \mathrm{L}^{\infty}\left(\Omega_{T} ; \mathbf{R}^{d}\right)$ and $\mathbf{A} \in \mathrm{L}^{\infty}\left(\Omega_{T} ; \mathrm{M}_{d}(\mathbf{R})\right)$ is symmetric with eigenvalues between $\alpha>0$ and $\beta \geqslant \alpha$ a.e. on $\Omega_{T}$.
Similarly as for the wave equation: $\mathbf{A}=\left[a^{1} \cdots a^{d}\right]$ and $w=\nabla u$

$$
\left[\begin{array}{cc}
1 & 0^{\top} \\
0 & \mathbf{0}
\end{array}\right] \partial_{t}\left[\begin{array}{c}
u \\
\mathrm{w}
\end{array}\right]-\sum_{i=1}^{d}\left[\begin{array}{cc}
\operatorname{div}^{i} & \left(\mathrm{a}^{i}\right)^{\top} \\
\mathrm{a}^{i} & \mathbf{0}
\end{array}\right] \partial_{x^{i}}\left[\begin{array}{c}
u \\
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$$

It is clearly symmetric; positivity should be checked.

## Example - heat equation, second form

New unknown vector function taking values in $\mathbf{R}^{d+1}$ :

$$
\mathrm{u}=\left[\begin{array}{l}
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Then the heat equation can be written as a first-order system

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\left\{\begin{array}{l}
\partial_{t} u+\operatorname{div} \mathbf{v}+c u-\mathbf{A}^{-1} \mathbf{b} \cdot \mathbf{v}=f \\
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which is a Friedrichs system

$$
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1 & 0^{\top} \\
0 & \mathbf{0}
\end{array}\right] \partial_{t}\left[\begin{array}{c}
u \\
\mathrm{v}
\end{array}\right]+\sum_{i=1}^{d}\left[\begin{array}{ccccc}
1 & \cdots & 1 & \cdots & 0 \\
\vdots & \ddots & 0 & \cdots & 0 \\
1 & \cdots & 0 & \cdots & 0 \\
\vdots & \cdots & 0 & \cdots & 0 \\
0 & \cdots & 0 & \cdots & 0
\end{array}\right] \partial_{x^{i}}\left[\begin{array}{c}
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$$

The condition (F1) holds. The positivity condition $\mathbf{C}+\mathbf{C}^{\top} \geqslant 2 \mu_{0} \mathbf{I}$ is fulfilled if and only if $c-\frac{1}{4} \mathbf{A}^{-1} \mathbf{b} \cdot \mathbf{b}$ is uniformly positive.

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\begin{aligned}
v & :=\partial_{x} u \\
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lead to the form:

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The Schwarz symmetries give us more equations, and the following choice leads to a symmetric system:

$$
\begin{aligned}
\partial_{x} u-v & =0 \\
-y \partial_{x} v-\partial_{y} w & =0 \\
\partial_{x} w-\partial_{y} v & =0 .
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Any solution of this equation satisfies the symmetric system:

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\mathbf{A}^{1} \partial_{x} \mathbf{u}+\mathbf{A}^{2} \partial_{y} \mathbf{u}=0,
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where the matrices are given by:

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\mathbf{A}^{1}:=\left[\begin{array}{cc}
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Thus, a symmetric hyperbolic system corresponds to the Tricomi's equation in the lower half plane.
It is not positive ([KOF1958] - a transformation providing the right form).

Why should one be interested in Friedrichs systems?
Symmetric hyperbolic systems
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Boundary conditions for Friedrichs systems
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allows the treatment of different types of usual boundary conditions.

Assumptions on the boundary matrix $\mathbf{M}$

We assume (for ae $\mathbf{x} \in \Gamma$ )
[KOF1958]
(FM1) $\quad\left(\forall \boldsymbol{\xi} \in \mathbf{R}^{r}\right) \quad \mathbf{M}(\mathbf{x}) \boldsymbol{\xi} \cdot \boldsymbol{\xi} \geqslant 0$,

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\mathbf{R}^{r}=\operatorname{ker}\left(\mathbf{A}_{\nu}(\mathbf{x})-\mathbf{M}(\mathbf{x})\right)+\operatorname{ker}\left(\mathbf{A}_{\nu}(\mathbf{x})+\mathbf{M}(\mathbf{x})\right) \tag{FM2}
\end{equation*}
$$

Such M is called the admissible boundary condition.
The boundary problem: for given $\mathrm{f} \in \mathrm{L}^{2}\left(\Omega ; \mathbf{R}^{r}\right)$ find u such that

$$
\left\{\begin{array}{l}
\mathcal{L} \mathrm{u}=\mathrm{f} \\
\left(\mathbf{A}_{\nu}-\mathbf{M}\right) \mathrm{u}_{\left.\right|_{\Gamma}}=0
\end{array}\right.
$$

## Different ways to enforce boundary conditions

Instead of

$$
\left(\mathbf{A}_{\nu}-\mathbf{M}\right) \mathbf{u}=0 \quad \text { on } \Gamma,
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Lax proposed boundary conditions with

$$
\mathrm{u}(\mathrm{x}) \in N(\mathrm{x}), \quad \mathrm{x} \in \Gamma,
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## Assumptions on $N$

maximal boundary conditions: (for ae $\mathbf{x} \in \Gamma$ )
(FX1)
$N(\mathbf{x})$ is non-negative with respect to $\mathbf{A}_{\nu}(\mathbf{x})$ :

$$
(\forall \boldsymbol{\xi} \in N(\mathbf{x})) \quad \mathbf{A}_{\nu}(\mathbf{x}) \boldsymbol{\xi} \cdot \boldsymbol{\xi} \geqslant 0
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(FX2) there is no non-negative subspace with respect to $\mathbf{A}_{\nu}(\mathbf{x})$, which contains $N(\mathbf{x}) ;$

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there is no non-negative subspace with respect to

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\end{equation*}
$$

or
[RSP\&LS1966]
Let $N(\mathbf{x})$ and $\tilde{N}(\mathbf{x}):=\left(\mathbf{A}_{\nu}(\mathbf{x}) N(\mathbf{x})\right)^{\perp}$ satisfy (for ae $\left.\mathbf{x} \in \Gamma\right)$
(FV1)

$$
\begin{array}{ll}
(\forall \boldsymbol{\xi} \in N(\mathbf{x})) & \mathbf{A}_{\nu}(\mathbf{x}) \boldsymbol{\xi} \cdot \boldsymbol{\xi} \geqslant 0 \\
(\forall \boldsymbol{\xi} \in \tilde{N}(\mathbf{x})) & \mathbf{A}_{\nu}(\mathbf{x}) \boldsymbol{\xi} \cdot \boldsymbol{\xi} \leqslant 0
\end{array}
$$

(FV2)

$$
\tilde{N}(\mathbf{x})=\left(\mathbf{A}_{\nu}(\mathbf{x}) N(\mathbf{x})\right)^{\perp} \quad \text { and } \quad N(\mathbf{x})=\left(\mathbf{A}_{\nu}(\mathbf{x}) \tilde{N}(\mathbf{x})\right)^{\perp}
$$

Equivalence of different descriptions of boundary conditions

Theorem. It holds
$(F M 1)-(F M 2) \quad \Longleftrightarrow \quad(F X 1)-(F X 2) \quad \Longleftrightarrow \quad(F V 1)-(F V 2)$,
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N(\mathbf{x}):=\operatorname{ker}\left(\mathbf{A}_{\nu}(\mathbf{x})-\mathbf{M}(\mathbf{x})\right)
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In fact, for a weak existence result some additional assumptions are needed [JR1994], [MJ2004].

## Classical results on well-posedness

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Contributions:
C. Morawetz, P. Lax, L. Sarason, R. S. Phillips, J. Rauch, ...

- the meaning of traces for functions in the graph space
- weak well-posedness results under additional assumptions (on $\mathbf{A}_{\nu}$ )
- regularity of solution
- numerical treatment

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... and new open questions.


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$$

$$
\begin{equation*}
(\exists c>0)(\forall \varphi \in \mathcal{D}) \quad\|(T+\tilde{T}) \varphi\|_{L} \leqslant c\|\varphi\|_{L} ; \tag{T2}
\end{equation*}
$$

$$
\begin{equation*}
\left(\exists \mu_{0}>0\right)(\forall \varphi \in \mathcal{D}) \quad\langle(T+\tilde{T}) \varphi \mid \varphi\rangle_{L} \geqslant 2 \mu_{0}\|\varphi\|_{L}^{2} . \tag{T3}
\end{equation*}
$$

Let $\mathcal{D}:=\mathrm{C}_{c}^{\infty}\left(\Omega ; \mathbf{R}^{r}\right), L=\mathrm{L}^{2}\left(\Omega ; \mathbf{R}^{r}\right)$ and $T, \tilde{T}: \mathcal{D} \longrightarrow L$ be defined by

$$
\begin{aligned}
T \mathbf{u} & :=\sum_{k=1}^{d} \partial_{k}\left(\mathbf{A}_{k} \mathbf{u}\right)+\mathbf{C u}, \\
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where $\mathbf{A}_{k}$ and $\mathbf{C}$ are as above (they satisfy (F1)-(F2)).

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Let $\mathcal{D}:=\mathrm{C}_{c}^{\infty}\left(\Omega ; \mathbf{R}^{r}\right), L=\mathrm{L}^{2}\left(\Omega ; \mathbf{R}^{r}\right)$ and $T, \tilde{T}: \mathcal{D} \longrightarrow L$ be defined by

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... fits in this framework.

## Prolongations

$\left(\mathcal{D},\langle\cdot \mid \cdot\rangle_{T}\right)$ is an inner product space, where

$$
\langle\cdot \mid \cdot\rangle_{T}:=\langle\cdot \mid \cdot\rangle_{L}+\langle T \cdot \mid T \cdot\rangle_{L} .
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Therefore $T=\tilde{T}_{\left.\right|_{W_{0}}}^{*}$, and analogously $\tilde{T}=T_{\left.\right|_{W_{0}}}^{*}$.
Abusing notation: $T, \tilde{T} \in \mathcal{L}\left(L ; W_{0}^{\prime}\right) \ldots(\mathrm{T} 1)-(\mathrm{T} 3)$

Formulation of the problem

Lemma. The graph space

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W:=\{u \in L: T u \in L\}=\{u \in L: \tilde{T} u \in L\},
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Problem: for given $f \in L$ find $u \in W$ such that $T u=f$.

Find sufficient conditions on $V \leqslant W$ such that $T_{\left.\right|_{V}}: V \longrightarrow L$ is an isomorphism.

## Boundary operator

Boundary operator $D \in \mathcal{L}\left(W ; W^{\prime}\right)$ :

$$
{ }_{W^{\prime}}\langle D u, v\rangle_{W}:=\langle T u \mid v\rangle_{L}-\langle u \mid \tilde{T} v\rangle_{L}, \quad u, v \in W
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Lemma. $\quad D$ is symmetric and satisfies

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\operatorname{ker} D=W_{0}
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\operatorname{im} D=W_{0}^{0}:=\left\{g \in W^{\prime}:\left(\forall u \in W_{0}\right) \quad W^{\prime}\langle g, u\rangle_{W}=0\right\}
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If $T$ is the Friedrichs operator $\mathcal{L}$, then for $\mathrm{u}, \mathrm{v} \in \mathrm{C}_{c}^{\infty}\left(\mathbf{R}^{d} ; \mathbf{R}^{r}\right)$ we have

$$
{ }_{W^{\prime}}\langle D \mathbf{u}, \mathrm{v}\rangle_{W}=\int_{\Gamma} \mathbf{A}_{\nu}(\mathbf{x}) \mathbf{u}_{\left.\right|_{\Gamma}}(\mathbf{x}) \cdot \mathbf{v}_{\left.\right|_{\Gamma}}(\mathbf{x}) d S(\mathbf{x})
$$

## Well-posedness theorem

Let $V$ and $\tilde{V}$ be subspaces of $W$ that satisfy

$$
\begin{array}{ll}
(\forall u \in V) & W^{\prime}\langle D u, u\rangle_{W} \geqslant 0 \\
(\forall v \in \tilde{V}) & { }_{W}\left\langle\langle D v, v\rangle_{W} \leqslant 0\right. \\
V=D(\tilde{V})^{0}, & \tilde{V}=D(V)^{0} . \tag{V2}
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(cone formalism)

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(cone formalism)
Theorem. Under assumptions (T1) - (T3) and (V1) - (V2), the operators $T_{\left.\right|_{V}}: V \longrightarrow L$ and $\tilde{T}_{\left.\right|_{\tilde{V}}}: \tilde{V} \longrightarrow L$ are isomorphisms.
[AE\&JLG\&GC2007]

Correspondence with classical assumptions

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\begin{array}{ll}
(\forall u \in V) & { }_{W}\left\langle\langle D u, u\rangle_{W} \geqslant 0,\right. \\
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(V2)
(FV1)

$$
(\forall \boldsymbol{\xi} \in N(\mathbf{x})) \quad \mathbf{A}_{\nu}(\mathbf{x}) \boldsymbol{\xi} \cdot \boldsymbol{\xi} \geqslant 0,
$$

$$
(\forall \boldsymbol{\xi} \in \tilde{N}(\mathbf{x})) \quad \mathbf{A}_{\nu}(\mathbf{x}) \boldsymbol{\xi} \cdot \boldsymbol{\xi} \leqslant 0
$$

$$
\begin{equation*}
\tilde{N}(\mathbf{x})=\left(\mathbf{A}_{\nu}(\mathbf{x}) N(\mathbf{x})\right)^{\perp} \quad \text { and } \quad N(\mathbf{x})=\left(\mathbf{A}_{\nu}(\mathbf{x}) \tilde{N}(\mathbf{x})\right)^{\perp} \tag{FV2}
\end{equation*}
$$

(for ae $\mathbf{x} \in \Gamma$ )

## Other sets of conditions in the classical setting (recall)

maximal boundary conditions: (for ae $\mathbf{x} \in \Gamma$ )

$$
\begin{equation*}
(\forall \boldsymbol{\xi} \in N(\mathbf{x})) \quad \mathbf{A}_{\nu}(\mathbf{x}) \boldsymbol{\xi} \cdot \boldsymbol{\xi} \geqslant 0 \tag{FX1}
\end{equation*}
$$ there is no non-negative subspace with respect to $\mathbf{A}_{\nu}(\mathbf{x})$, which contains $N(\mathbf{x})$,

admissible boundary conditions: there exists a matrix function $\mathbf{M}: \Gamma \longrightarrow \mathrm{M}_{r}(\mathbf{R})$ such that (for ae $\mathbf{x} \in \Gamma$ )
(FM1)

$$
\left(\forall \boldsymbol{\xi} \in \mathbf{R}^{r}\right) \quad \mathbf{M}(\mathbf{x}) \boldsymbol{\xi} \cdot \boldsymbol{\xi} \geqslant 0
$$

(FM2)

$$
\mathbf{R}^{r}=\operatorname{ker}\left(\mathbf{A}_{\nu}(\mathbf{x})-\mathbf{M}(\mathbf{x})\right)+\operatorname{ker}\left(\mathbf{A}_{\nu}(\mathbf{x})+\mathbf{M}(\mathbf{x})\right)
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## Correspondence - maximal b.c.

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subspace $V$ is maximal non-negative with respect to $D$ :
(X1) $V$ is non-negative with respect to $D: \quad(\forall v \in V) \quad{ }_{W}\langle D v, v\rangle_{W} \geqslant 0$,
(X2) there is no non-negative subspace with respect to $D$ that contains $V$.

## Correspondence - admissible b.c.

admissible boundary condition: there exist a matrix function $\mathbf{M}: \Gamma \longrightarrow \mathrm{M}_{r}(\mathbf{R})$ such that (for ae $\mathbf{x} \in \Gamma$ )
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\end{equation*}
$$

admissible boundary condition: there exist $M \in \mathcal{L}\left(W ; W^{\prime}\right)$ that satisfy

$$
\begin{gather*}
(\forall u \in W) \quad W^{\prime}\langle M u, u\rangle_{W} \geqslant 0  \tag{M1}\\
W=\operatorname{ker}(D-M)+\operatorname{ker}(D+M)
\end{gather*}
$$

Equivalence of different descriptions of b.c.

Theorem. (classical) It holds

$$
(F M 1)-(F M 2) \quad \Longleftrightarrow \quad(F V 1)-(F V 2) \quad \Longleftrightarrow \quad(F X 1)-(F X 2),
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Theorem. (A. Ern, J.-L. Guermond, G. Caplain) It holds

$$
(M 1)-(M 2) \quad \Longrightarrow(V 1)-(V 2) \quad \Longrightarrow \quad(X 1)-(X 2)
$$

with

$$
V:=\operatorname{ker}(D-M)
$$

## $(\mathrm{M} 1)-(\mathrm{M} 2) \quad \longleftarrow \quad(\mathrm{V} 1)-(\mathrm{V} 2)$

Theorem. Let $V$ and $\tilde{V}$ satisfy (V1)-(V2), and suppose that there exist operators $P \in \mathcal{L}(W ; V)$ and $Q \in \mathcal{L}(W ; \tilde{V})$ such that

$$
\begin{array}{ll}
(\forall v \in V) & D(v-P v)=0, \\
(\forall v \in \tilde{V}) & D(v-Q v)=0, \\
& D P Q=D Q P .
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Let us define $M \in \mathcal{L}\left(W ; W^{\prime}\right)$ (for $\left.u, v \in W\right)$ with

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& \quad+{ }_{W^{\prime}}\langle D(P+Q-P Q) u, v\rangle_{W}-{ }_{W^{\prime}}\langle D u,(P+Q-P Q) v\rangle_{W} .
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Then $V:=\operatorname{ker}(D-M), \tilde{V}:=\operatorname{ker}\left(D+M^{*}\right)$, and $M$ satisfies (M1)-(M2).
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Then $V:=\operatorname{ker}(D-M), \tilde{V}:=\operatorname{ker}\left(D+M^{*}\right)$, and $M$ satisfies (M1)-(M2).
Lemma. Suppose additionally that $V+\tilde{V}$ is closed. Then the operators $P$ and $Q$ from previous theorem do exist.

## When this is satisfied?

## Lemma. (K. Burazin, N.A.)

If codim $W_{0}\left(=\operatorname{dim} W / W_{0}\right)$ is finite, then the set $V+\tilde{V}$ is closed whenever $V$ and $\tilde{V}$ satisfy (V1)-(V2).

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Let subspaces $V$ and $\tilde{V}$ of space $W$ satisfy (V1)-(V2), $V \cap \tilde{V}=W_{0}$, and $W \neq V+\tilde{V}$.

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Then $V+\tilde{V}$ is not closed in $W$.
Moreover, there do not exist operators $P$ and $Q$ with desired properties.

## Counter example

Let $\Omega \subseteq \mathbf{R}^{2}, \mu>0$ and $f \in \mathrm{~L}^{2}(\Omega)$ be given. Scalar elliptic equation

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\begin{aligned}
& V:=\left\{(\mathrm{p}, u)^{\top} \in W: \mathcal{T}_{\text {div }} \mathrm{p}=\alpha \mathcal{T}_{\mathrm{H}^{1} u} u,\right. \\
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## Lemma.

The above $V$ and $\tilde{V}$ satisfy (V1)-(V2), $V \cap \tilde{V}=W_{0}$ and $V+\tilde{V} \neq W$.

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The above $V$ and $\tilde{V}$ satisfy (V1)-(V2), $V \cap \tilde{V}=W_{0}$ and $V+\tilde{V} \neq W$. There exists an operator $M \in \mathcal{L}\left(W ; W^{\prime}\right)$, that satisfies (M1)-(M2) and $V=\operatorname{ker}(D-M)$.

Why should one be interested in Friedrichs systems?
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Abstract formulation
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## New notation

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[u \mid v]:={ }_{W^{\prime}}\langle D u, v\rangle_{W}=\langle T u \mid v\rangle_{L}-\langle u \mid \tilde{T} v\rangle_{L}, \quad u, v \in W
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$$
\begin{align*}
& (\forall v \in V) \quad[v \mid v] \geqslant 0, \\
& (\forall v \in \tilde{V}) \quad[v \mid v] \leqslant 0 ; \\
& V=\tilde{V}^{[\perp]}, \quad \tilde{V}=V^{[\perp]} . \tag{V2}
\end{align*}
$$

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( ${ }^{[\perp]}$ stands for $[\cdot \mid \cdot]$-orthogonal complement)
subspace $V$ is maximal non-negative in $(W,[\cdot \mid \cdot])$ :
(X1) $\quad V$ is non-negative in $(W,[\cdot \mid \cdot]): \quad(\forall v \in V) \quad[v \mid v] \geqslant 0$,
(X2) there is no non-negative subspace in $(W,[\cdot \mid \cdot])$ containing $V$.

## Kreĭn spaces

( $W,[\cdot \mid \cdot]$ ) is not a Krein space - it is a degenerate space, because its Gramm operator $G:=j \circ D \quad\left(j: W^{\prime} \longrightarrow W\right.$ is the canonical isomorphism $)$ has large kernel:

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Important: $\operatorname{im} D$ is closed and $\operatorname{ker} D=W_{0}$.

## Quotient Kreĭn space

Lemma. Let $U \supseteq W_{0}$ and $Y$ be subspaces of $W$. Then
a) $U$ is closed if and only if $\hat{U}:=\{\hat{v}: v \in U\}$ is closed in $\hat{W}$;
b) $(\widehat{U+Y})=\left\{u+v+W_{0}: u \in U, v \in Y\right\}=\hat{U}+\hat{Y}$;
c) $U+Y$ is closed if and only if $\hat{U}+\hat{Y}$ is closed;
d) $(\hat{Y})^{[\perp \hat{\jmath}}=\widehat{Y^{[\perp]}}$.
e) if $Y$ is maximal non-negative (non-positive) in $W$, than $\hat{Y}$ is maximal non-negative (non-positive) in $\hat{W}$;
f) if $\hat{U}$ is maximal non-negative (non-positive) in $\hat{W}$, then $U$ is maximal non-negative (non-positive) in $W$.

Theorem. a) If subspaces $V$ and $\tilde{V}$ satisfy (V1)-(V2), then $V$ is maximal non-negative in $W$ (satisfies (X1)-(X2)) and $\tilde{V}$ is maximal non-positive in $W$.
b) If $V$ is maximal non-negative in $W$, then $V$ and $\tilde{V}:=V^{[\perp]}$ satisfy (V1)-(V2).


Theorem. [EGC] (T1)-(T3) and $M \in \mathcal{L}\left(W ; W^{\prime}\right)$ satisfy (M) imply

$$
V:=\operatorname{ker}(D-M) \quad \text { and } \quad \tilde{V}:=\operatorname{ker}\left(D+M^{*}\right) \quad \text { satisfy }(V) .
$$

Corollary. Under above assumptions
$T_{\left.\right|_{\operatorname{ker}(D-M)}}: \operatorname{ker}(D-M) \longrightarrow L \quad i \quad \tilde{T}_{\left.\right|_{\operatorname{ker}\left(D+M^{*}\right)}}: \operatorname{ker}\left(D+M^{*}\right) \longrightarrow L$ are isomorphisms.
$(\mathrm{M} 1)-(\mathrm{M} 2) \quad \longleftarrow \quad(\mathrm{V} 1)-(\mathrm{V} 2) \quad$ (recall $)$
Theorem. Let $V$ and $\tilde{V}$ satisfy (V1)-(V2), and suppose that there exist operators $P \in \mathcal{L}(W ; V)$ and $Q \in \mathcal{L}(W ; \tilde{V})$ such that

$$
\begin{array}{ll}
(\forall v \in V) & D(v-P v)=0, \\
(\forall v \in \tilde{V}) & D(v-Q v)=0, \\
& D P Q=D Q P .
\end{array}
$$

Let us define $M \in \mathcal{L}\left(W ; W^{\prime}\right)$ (for $\left.u, v \in W\right)$ with

$$
\begin{aligned}
& W_{W^{\prime}}\langle M u, v\rangle_{W}={ }_{W^{\prime}}\langle D P u, P v\rangle_{W}-{ }_{W^{\prime}}\langle D Q u, Q v\rangle_{W} \\
& \quad+{ }_{W^{\prime}}\langle D(P+Q-P Q) u, v\rangle_{W}-{ }_{W^{\prime}}\langle D u,(P+Q-P Q) v\rangle_{W} .
\end{aligned}
$$

Then $V:=\operatorname{ker}(D-M), \tilde{V}:=\operatorname{ker}\left(D+M^{*}\right)$, and $M$ satisfies (M1)-(M2).
Lemma. Suppose additionally that $V+\tilde{V}$ is closed. Then the operators $P$ and $Q$ from previous theorem do exist.
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Closedness of $V+\tilde{V}$ is actually equivalent to the existence of operators $P$ and $Q$.

## On existence of $P$ and $Q$

Our original approach was indirect:
Firstly, the existence of $P$ and $Q$ implies the existence of certain projectors in the quotient Kreĭn space; more precisely:

$$
\hat{P} \hat{w}:=\widehat{P w}, \quad \hat{Q} \hat{w}:=\widehat{Q w}, \quad w \in W
$$

the projectors $\hat{P}, \hat{Q}: \hat{W} \longrightarrow \hat{W}$ are defined, satisfying

$$
\begin{gathered}
\hat{P}^{2}=\hat{P} \quad \text { and } \quad \hat{Q}^{2}=\hat{Q}, \\
\text { im } \hat{P}=\hat{V} \quad \text { and } \quad \operatorname{im} \hat{Q}=\hat{\tilde{V}}, \\
\hat{P} \hat{Q}=\hat{Q} \hat{P} .
\end{gathered}
$$

Secondly, this allowed us to prove the existence of corresponding projectors on $W$.
$(\mathrm{M} 1)-(\mathrm{M} 2) \Longleftarrow(\mathrm{V} 1)-(\mathrm{V} 2) \quad$ (direct proof)

Theorem. If $V, \tilde{V}$ are two closed subspaces of $W$ that satisfy $W_{0} \subseteq V \cap \tilde{V}$, then the following statements are equivalent:
a) There exist operators $P \in \mathcal{L}(W ; V)$ and $Q \in \mathcal{L}(W ; \tilde{V})$, such that

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b) There exist projectors $P^{\prime}, Q^{\prime} \in \mathcal{L}(W ; W)$, such that

$$
\begin{gathered}
P^{\prime 2}=P^{\prime} \quad \text { and } \quad Q^{\prime 2}=Q^{\prime} \\
\operatorname{im} P^{\prime}=V \quad \text { and } \quad \operatorname{im} Q^{\prime}=\tilde{V} \\
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(b) is equivalent to closedness of $V+\tilde{V}$.

## $(\mathrm{M} 1)-(\mathrm{M} 2) \Longleftarrow(\mathrm{V} 1)-(\mathrm{V} 2) \quad$ (cont. $)$

Theorem.
a) $V, \tilde{V} \leqslant W$ satisfy (V), and exists a closed subspace $W_{2} \subseteq C^{-}$of $W$, $V \dot{+} W_{2}=W$, then there exist an operator $M \in \mathcal{L}\left(W ; W^{\prime}\right)$ satisfying (M) and $V=\operatorname{ker}(D-M)$.
If we define $W_{1}$ as orthogonal complement of $W_{0}$ in $V$, so that $W=W_{1} \dot{+} W_{0} \dot{+} W_{2}$, and denote by $R_{1}, R_{0}, R_{2}$ projectors that correspond to above direct sum, then one such operator is given with $M=D\left(R_{1}-R_{2}\right)$.

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b) $M \in \mathcal{L}\left(W ; W^{\prime}\right)$ an operator satisfying (M1)-(M2), $V:=\operatorname{ker}(D-M)$.

For $W_{2}$, the orthogonal complement of $W_{0}$ in $\operatorname{ker}(D+M), W_{2} \subseteq C^{-}$is closed, $V \dot{+} W_{2}=W$, and $M$ coincide with the operator in (a).

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Lemma. Let $W_{2}^{\prime \prime} \leqslant W$ satisfies $W_{2}^{\prime \prime} \subseteq C^{-}$and $W_{2}^{\prime \prime}+V=W$.
Then there is a closed subspace $W_{2}$ of $W$, such that $W_{2} \subseteq C^{-}$and $W_{2} \dot{+} V=W$.

## $(\mathrm{M} 1)-(\mathrm{M} 2) \Longleftarrow(\mathrm{V} 1)-(\mathrm{V} 2) \quad$ (cont. $)$

Lemma. If $U_{1}+U_{2}=W$ for some subspaces $U_{1} \subseteq C^{+}$and $U_{2} \subseteq C^{-}$of $W$, then $U_{1} \cap U_{2} \subseteq W_{0}$.
If additionally $U_{1}$ is maximal nonnegative and $U_{2}$ maximal nonpositive, then $U_{1} \cap U_{2}=W_{0}$.

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Theorem. For a maximal nonnegative subspace $V$ of $W$, it is equivalent:
a) There is a maximal nonpositive subspace $W_{2}$ of $W$, such that $W_{2}+V=W$;
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Corollary. The conditions $(V)$ and $(M)$ are equivalent.

## Some used properties

Theorem. a) [.| $\cdot]$-orthogonal complement of a maximal non-negative (non-positive) subspace is non-positive (non-negative).
b) Each maximal semi-definite subspace contains all isotropic vectors in $W$.
c) If $L$ is a non-negative (non-positive) subspace of a Krein space, such that $L^{[\perp]}$ is non-positive (non-negative), then $\mathrm{Cl} L$ is maximal non-negative (non-positive).
d) Each maximal semi-definite subspace of a Krein space is closed.
e) A subspace $L$ of a Krein space is closed if and only if $L=L^{[\perp][\perp]}$.
f) For a subspace $L$ of a Krein space $W$ it holds

$$
L \cap L^{[\perp]}=\{0\} \quad \Longleftrightarrow \quad \mathrm{Cl}\left(L+L^{[\perp]}\right)=W
$$

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$$

Theorem. Assume (T1) - (T3) and the existence of $M \in \mathcal{L}\left(W ; W^{\prime}\right)$ satisfying

$$
\begin{gather*}
(\forall u \in W) \quad W^{\prime}\langle M u, u\rangle_{W} \geqslant 0,  \tag{M1}\\
W=\operatorname{ker}(D-M)+\operatorname{ker}(D+M) .
\end{gather*}
$$

Then the operator $T_{\left.\right|_{\operatorname{ker}(D-M)}}: \operatorname{ker}(D-M) \longrightarrow L$ is an isomorphism.

Application to the classical theory

Let $\mathcal{D}:=\mathrm{C}_{c}^{\infty}\left(\Omega ; \mathbf{R}^{r}\right), L=\mathrm{L}^{2}\left(\Omega ; \mathbf{R}^{r}\right)$ and $T, \tilde{T}: \mathcal{D} \longrightarrow L$ be defined by

$$
\begin{aligned}
T \mathbf{u} & :=\sum_{k=1}^{d} \partial_{k}\left(\mathbf{A}_{k} \mathbf{u}\right)+\mathbf{C u}, \\
\tilde{T} \mathbf{u} & :=-\sum_{k=1}^{d} \partial_{k}\left(\mathbf{A}_{k}^{\top} \mathbf{u}\right)+\left(\mathbf{C}^{\top}+\sum_{k=1}^{d} \partial_{k} \mathbf{A}_{k}^{\top}\right) \mathbf{u},
\end{aligned}
$$

where $\mathbf{A}_{k}$ and $\mathbf{C}$ are as before (they satisfy (F1)-(F2)).
Then $T$ and $\tilde{T}$ satisfy (T1)-(T3)

Application to the classical theory

Let $\mathcal{D}:=\mathrm{C}_{c}^{\infty}\left(\Omega ; \mathbf{R}^{r}\right), L=\mathrm{L}^{2}\left(\Omega ; \mathbf{R}^{r}\right)$ and $T, \tilde{T}: \mathcal{D} \longrightarrow L$ be defined by

$$
\begin{aligned}
T \mathbf{u} & :=\sum_{k=1}^{d} \partial_{k}\left(\mathbf{A}_{k} \mathbf{u}\right)+\mathbf{C u}, \\
\tilde{T} \mathbf{u} & :=-\sum_{k=1}^{d} \partial_{k}\left(\mathbf{A}_{k}^{\top} \mathbf{u}\right)+\left(\mathbf{C}^{\top}+\sum_{k=1}^{d} \partial_{k} \mathbf{A}_{k}^{\top}\right) \mathbf{u},
\end{aligned}
$$

where $\mathbf{A}_{k}$ and $\mathbf{C}$ are as before (they satisfy (F1)-(F2)).
Then $T$ and $\tilde{T}$ satisfy (T1)-(T3) and

$$
W=\left\{\mathbf{u} \in \mathrm{L}^{2}\left(\Omega ; \mathbf{R}^{r}\right): \sum_{k=1}^{d} \partial_{k}\left(\mathbf{A}_{k} \mathbf{u}\right)+\mathbf{C u} \in \mathrm{L}^{2}\left(\Omega ; \mathbf{R}^{r}\right)\right\} .
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For given matrix field M is there an operator $M$ determined by M in some natural way?

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meaning that any smooth weak solution is also a classical solution
i.e. smooth $u \in \operatorname{ker}(A-M)$ should satisfy $\left(\mathbf{A}_{\nu}-\mathbf{M}\right) u_{u_{\partial \Omega}}=0$

Representation of $D$ and $M$ via matrix fields

For $\mathrm{u}, \mathrm{v} \in \mathrm{C}_{c}^{\infty}\left(\mathbf{R}^{d} ; \mathbf{R}^{r}\right)$ we have

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Lemma. If $\mathbf{M}$ satisfies ( $F M$ ), then (for ae $\mathbf{x} \in \Gamma$ ) there is a pair of projectors $\mathbf{S}_{+}(\mathbf{x}), \mathbf{S}_{-}(\mathbf{x})$ (i.e. $\mathbf{S}_{+}(\mathbf{x})+\mathbf{S}_{-}(\mathbf{x})=\mathbf{I}$ and $\left.\mathbf{S}_{+}(\mathbf{x}) \mathbf{S}_{-}(\mathbf{x})=\mathbf{S}_{-}(\mathbf{x}) \mathbf{S}_{+}(\mathbf{x})=\mathbf{0}\right)$, s.t.

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... not good enough for applications to hyperbolic equations

## $\mathbf{P}$ is not necessarily a projector

## Lemma

For a matrix field M the following statements are equivalent.

- M satisfies (FM2).
- There is a matrix field $\mathbf{P}$ such that $\mathbf{M}=\mathbf{A}_{\nu}(\mathbf{I}-2 \mathbf{P})$ and $\operatorname{ker}\left(\mathbf{A}_{\nu} \mathbf{P}\right)+\operatorname{ker}\left(\mathbf{A}_{\nu}(\mathbf{I}-\mathbf{P})\right)=\mathbf{R}^{r}$ ae in $\partial \Omega$.


## Main result for Friedrichs systems

Theorem. Let matrix field $\mathrm{M} \in \mathrm{L}^{\infty}\left(\Gamma ; \mathrm{M}_{r}(\mathbf{R})\right)$ satisfy (FM), and let $\mathrm{S}_{-}$be extendable to a measurable function on $\mathrm{Cl} \Omega$, and satisfy: (S1) The multiplication operator $\mathcal{S}_{-, p}$ is in $\mathcal{L}(W)$.

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Test on examples ... assumptions are reasonable ...

An example - scalar elliptic equation
$\Omega \subseteq \mathbf{R}^{2}, \mu>0$ and $f \in \mathrm{~L}^{2}(\Omega)$ given.

$$
-\triangle u+\mu u=f
$$

## An example - scalar elliptic equation

$\Omega \subseteq \mathbf{R}^{2}, \mu>0$ and $f \in \mathrm{~L}^{2}(\Omega)$ given.

$$
-\Delta u+\mu u=f
$$

can be written as a first-order system

$$
\left\{\begin{array}{l}
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which is a Friedrichs system with the choice of

$$
\mathbf{A}_{1}=\left[\begin{array}{lll}
0 & 0 & 1 \\
0 & 0 & 0 \\
1 & 0 & 0
\end{array}\right], \quad \mathbf{A}_{2}=\left[\begin{array}{lll}
0 & 0 & 0 \\
0 & 0 & 1 \\
0 & 1 & 0
\end{array}\right], \quad \mathbf{C}=\left[\begin{array}{lll}
1 & 0 & 0 \\
0 & 1 & 0 \\
0 & 0 & \mu
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1 & 0 & 0 \\
0 & 1 & 0 \\
0 & 0 & \mu
\end{array}\right]
$$

Note

$$
\mathbf{A}_{\nu}=\nu_{1} \mathbf{A}_{1}+\nu_{2} \mathbf{A}_{2}=\left[\begin{array}{ccc}
0 & 0 & \nu_{1} \\
0 & 0 & \nu_{2} \\
\nu_{1} & \nu_{2} & 0
\end{array}\right] .
$$

Elliptic equation - different boundary conditions

$$
\begin{array}{ccc}
\mathbf{M} & \mathbf{A}_{\nu}-\mathbf{M} & \left(\mathbf{A}_{\nu}-\mathbf{M}\right)\left[\begin{array}{l}
\mathbf{p} \\
u
\end{array}\right]_{\left.\right|_{\Gamma}}=0 \\
{\left[\begin{array}{ccc}
0 & 0 & -\nu_{1} \\
0 & 0 & -\nu_{2} \\
\nu_{1} & \nu_{2} & 0
\end{array}\right]} & {\left[\begin{array}{ccc}
0 & 0 & 2 \nu_{1} \\
0 & 0 & 2 \nu_{2} \\
0 & 0 & 0
\end{array}\right]} & u_{\mid \Gamma}=0
\end{array}
$$

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\begin{gathered}
\mathbf{M}
\end{gathered} \begin{array}{cc}
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$$
\begin{aligned}
& \text { M } \\
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& {\left[\begin{array}{ccc}
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\end{array}\right] \quad u_{\left.\right|_{\Gamma}}=0} \\
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\end{array}\right] \quad\left[\begin{array}{ccc}
0 & 0 & 0 \\
0 & 0 & 0 \\
2 \nu_{1} & 2 \nu_{2} & 0
\end{array}\right]} \\
& \left.\boldsymbol{\nu} \cdot(\nabla u)\right|_{\Gamma}=0 \\
& {\left[\begin{array}{ccc}
0 & 0 & \nu_{1} \\
0 & 0 & \nu_{2} \\
-\nu_{1} & -\nu_{2} & 2 \alpha
\end{array}\right]\left[\begin{array}{ccc}
0 & 0 & 0 \\
0 & 0 & 0 \\
2 \nu_{1} & 2 \nu_{2} & 2 \alpha
\end{array}\right]} \\
& \boldsymbol{\nu} \cdot(\nabla u)_{\left.\right|_{\Gamma}}+\left.\alpha u\right|_{\Gamma}=0
\end{aligned}
$$

Elliptic equation - different boundary conditions

$$
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& {\left[\begin{array}{ccc}
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0 & 0 & 0
\end{array}\right]} \\
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\end{array}\right] \quad \nu \cdot(\nabla u)_{\Gamma}+\alpha u_{\left.\right|_{\Gamma}}=0}
\end{aligned}
$$

All above matrices M satisfy (FM).

Elliptic equation - projector $\mathbf{S}_{-}$

Dirichlet:

$$
\mathbf{S}_{-}=\left[\begin{array}{lll}
1 & 0 & 0 \\
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\end{array}\right]
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Robin:

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Constants can easily be extended, but we need $\boldsymbol{\nu}: \Gamma \longrightarrow \mathbf{R}^{r}$ to be Lipschitz in order to have bounded multiplication for the Robin b.c.

## Practical sufficient conditions

## Lemma

For constant $\mathbf{A}_{k} \in \mathrm{M}_{r}(\mathbf{R})$ and $\mathbf{P} \in \mathrm{M}_{r}(\mathbf{R})$ the multiplication operator $\mathrm{u} \mapsto \mathbf{P u}$ belongs to $\mathcal{L}(W)$ if and only if there exists $\mathbf{S} \in \mathrm{M}_{r}(\mathbf{R})$ such that $\mathbf{A}_{k} \mathbf{P}=\mathbf{S A}_{k}$ for $k \in 1 . . d$.

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Theorem (sufficient conditions)
Let $\mathbf{P}: \mathrm{Cl} \Omega \longrightarrow \mathrm{M}_{r}(\mathbf{R})$ be a Lipschitz matrix function satisfying:
$-\left(\exists \mathbf{S} \in \mathrm{W}^{1, \infty}\left(\Omega ; \mathrm{M}_{r}(\mathbf{R})\right)\right)(\forall k \in 1 . . d) \quad \mathbf{A}_{k} \mathbf{P}=\mathbf{S A}_{k}$

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$\operatorname{ker}\left(\mathbf{A}_{\nu}(\mathbf{x}) \mathbf{P}(\mathbf{x})\right)+\operatorname{ker}\left(\left(\mathbf{A}_{\nu}(\mathbf{x})(\mathbf{I}-\mathbf{P}(\mathbf{x}))\right)=\mathbf{R}^{r}\right.$.


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$\operatorname{ker}\left(\mathbf{A}_{\boldsymbol{\nu}}(\mathbf{x}) \mathbf{P}(\mathbf{x})\right)+\operatorname{ker}\left(\left(\mathbf{A}_{\boldsymbol{\nu}}(\mathbf{x})(\mathbf{I}-\mathbf{P}(\mathbf{x}))\right)=\mathbf{R}^{r}\right.$.
Then formula (m), for $\mathbf{M}(\mathbf{x}):=\mathbf{A}_{\nu}(\mathbf{x})(\mathbf{I}-2 \mathbf{P}(\mathbf{x}))$ on $\partial \Omega$, defines a bounded operator $M \in \mathcal{L}\left(W ; W^{\prime}\right)$ satisfying (M).


## Tests on examples

Applications on hyperbolic equations (transport and wave equation)

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N. Antonić, K. Burazin, M. Vrdoljak: Second-order equations as Friedrichs systems, Nonlin. Analysis B: Real World Appl. 14 (2014) 290-305.

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Applications on hyperbolic equations (transport and wave equation)
N. Antonić, K. Burazin, M. Vrdoljak: Second-order equations as Friedrichs systems, Nonlin. Analysis B: Real World Appl. 14 (2014) 290-305.
. . still unable do get good results for mixed type problems

## Heat equation

... with zero initial and Dirichlet boundary condition:

$$
\left\{\begin{array}{l}
\partial_{t} u-\operatorname{div}_{\mathbf{x}}\left(\mathbf{A} \nabla_{\mathbf{x}} u\right)+\mathbf{b} \cdot \nabla_{\mathbf{x}} u+c u=f \text { in } \Omega_{T} \\
u=0 \text { on } \partial \Omega \times\langle 0, T\rangle \\
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$$

...as a Friedrichs system:

$$
\left\{\begin{array}{l}
\nabla_{\mathbf{x}} u_{d+1}+\mathbf{A}^{-1} \mathbf{u}_{d}=0 \\
\partial_{t} u_{d+1}+\operatorname{div}_{\mathbf{x}} \mathbf{u}_{d}+c u_{d+1}-\mathbf{A}^{-1} \mathbf{b} \cdot \mathbf{u}_{d}=f
\end{array}\right.
$$

(note that we use $\mathrm{u}=\left(\mathrm{u}_{d}, u_{d+1}\right)^{\top}$ ).

Friedrichs operator and the graph space

The operator $T$ is given by

$$
T\left[\begin{array}{c}
\mathbf{u}_{d} \\
u_{d+1}
\end{array}\right]=\left[\begin{array}{c}
\nabla_{\mathbf{x}} u_{d+1}+\mathbf{A}^{-1} \mathbf{u}_{d} \\
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\end{array}\right],
$$

while the corresponding graph space is

$$
\begin{aligned}
& W=\left\{\mathrm{u} \in \mathrm{~L}^{2}\left(\Omega_{T} ; \mathbf{R}^{d+1}\right): \nabla_{\mathbf{x}} u_{d+1} \in \mathrm{~L}^{2}\left(\Omega_{T} ; \mathbf{R}^{d}\right)\right. \\
&\left.\& \partial_{t} u_{d+1}+\operatorname{div}_{\mathbf{x}} \mathrm{u}_{d} \in \mathrm{~L}^{2}\left(\Omega_{T}\right)\right\} \\
&=\left\{\mathrm{u} \in \mathrm{~L}_{\mathrm{div}}^{2}\left(\Omega_{T}\right): \nabla_{\mathbf{x}} u_{d+1} \in \mathrm{~L}^{2}\left(\Omega_{T} ; \mathbf{R}^{d}\right)\right\} \\
&=\left\{\mathrm{u} \in \mathrm{~L}_{\mathrm{div}}^{2}\left(\Omega_{T}\right): u_{d+1} \in \mathrm{~L}^{2}\left(0, T ; \mathrm{H}^{1}(\Omega)\right)\right\} .
\end{aligned}
$$

## Properties of the last component

Lemma. The projection $\mathrm{u}=\left(\mathrm{u}_{d}, u_{d+1}\right)^{\top} \mapsto u_{d+1}$ is a continuous linear operator from $W$ to $W(0, T)$, which is continuously embedded to $\mathrm{C}\left([0, T] ; \mathrm{L}^{2}(\Omega)\right)$.

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The space

$$
W(0, T)=\left\{u \in \mathrm{~L}^{2}\left(0, T ; \mathrm{H}^{1}(\Omega)\right): \partial_{t} u \in \mathrm{~L}^{2}\left(0, T ; \mathrm{H}^{-1}(\Omega)\right)\right\}
$$

is a Banach space when equipped by norm

$$
\|\mathbf{u}\|_{W(0, T)}=\sqrt{\|u\|_{\mathrm{L}^{2}\left(0, T ; \mathrm{H}^{1}(\Omega)\right)}^{2}+\left\|\partial_{t} u\right\|_{\mathrm{L}^{2}\left(0, T ; \mathrm{H}^{-1}(\Omega)\right)}^{2}} .
$$

## Main result

Let

$$
\begin{aligned}
& V=\left\{\mathrm{u} \in W: u_{d+1} \in \mathrm{~L}^{2}\left(0, T ; \mathrm{H}_{0}^{1}(\Omega)\right), \quad u_{d+1}(\cdot, 0)=0 \text { a.e. on } \Omega\right\} \\
& \tilde{V}=\left\{\mathrm{v} \in W: v_{d+1} \in \mathrm{~L}^{2}\left(0, T ; \mathrm{H}_{0}^{1}(\Omega)\right), \quad v_{d+1}(\cdot, T)=0 \text { a.e. on } \Omega\right\} .
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Do they satisfy (V1)-(V2)?

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\end{aligned}
$$

Do they satisfy (V1)-(V2)? Technical...
Theorem
The above $V$ and $\widetilde{V}$ satisfy (V1)-(V2), and therefore the operator $T_{\left.\right|_{V}}: V \longrightarrow L$ is an isomorphism.

## Two-field theory...

Heat equation with $\mathrm{b}=0$ and $c=0$ :

$$
\left\{\begin{aligned}
& \partial_{t} u-\operatorname{div}_{\mathbf{x}}\left(\mathbf{A} \nabla_{\mathbf{x}} u\right)=f \text { in } \Omega_{T} \\
& u=0 \text { on } \Gamma \times\langle 0, T\rangle \\
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Two field theory:
developed by Ern and Guermond for elliptic problems
matrices need to be of the form

$$
\mathbf{A}^{k}=\left[\begin{array}{cc}
\mathbf{0} & \mathrm{B}^{k} \\
\left(\mathrm{~B}^{k}\right)^{\top} & a^{k}
\end{array}\right] \quad \text { and } \quad \mathbf{C}=\left[\begin{array}{cc}
\mathbf{C}^{d} & 0 \\
0^{\top} & c^{d+1}
\end{array}\right]
$$

where $\mathrm{B}^{k} \in \mathbf{R}^{d}$ are constant vectors, $a^{k} \in \mathrm{~W}^{1, \infty}\left(\Omega_{T}\right), \mathbf{C}^{d} \in \mathrm{~L}^{\infty}\left(\Omega_{T} ; \mathrm{M}_{d}(\mathbf{R})\right)$ and $c^{d+1} \in \mathrm{~L}^{\infty}\left(\Omega_{T}\right), k \in 1 . .(d+1)$.

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For the heat equation matrices have this form!

Instead of coercivity (positivity) condition (F2), the following is required:

$$
\begin{aligned}
& \left(\exists \mu_{1}>0\right)\left(\forall \boldsymbol{\xi}=\left(\boldsymbol{\xi}_{d}, \xi_{d+1}\right) \in \mathbf{R}^{d+1}\right) \\
& \left.\qquad\left(\mathbf{C}+\mathbf{C}^{\top}+\sum_{k=1}^{d+1} \partial_{k} \mathbf{A}_{k}\right) \xi \cdot \boldsymbol{\xi} \geqslant 2 \mu_{1}\left|\boldsymbol{\xi}_{d}\right|^{2} \quad \text { (a.e. on } \Omega\right),
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& \left(\exists \mu_{2}>0\right)(\forall \mathbf{u} \in V \cup \tilde{V})
\end{aligned}
$$

$$
\sqrt{\langle\mathcal{L} \mathbf{u} \mid \mathbf{u}\rangle_{\mathrm{L}^{2}\left(\Omega_{T} ; \mathbf{R}^{d+1}\right)}}+\left\|\mathrm{B} u_{d+1}\right\|_{\mathrm{L}^{2}\left(\Omega_{T} ; \mathbf{R}^{d}\right)} \geqslant \mu_{2}\left\|u_{d+1}\right\|_{\mathrm{L}^{2}\left(\Omega_{T}\right)},
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For our system both conditions are trivially fulfilled.
Therefore, we have the well-posedness result.

An example - stationary diffusion equation
We consider the equation

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-\operatorname{div}(\mathbf{A} \nabla u)+c u=f
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$\beta^{\prime} \geq \alpha^{\prime}>0$, and
$\mathbf{A} \in \mathcal{M}_{d}\left(\alpha^{\prime}, \beta^{\prime} ; \Omega\right):=\left\{\mathbf{A} \in \mathrm{L}^{\infty}\left(\Omega ; \mathrm{M}_{d}(\mathbf{R})\right):\right.$

$$
\left.\left(\forall \boldsymbol{\xi} \in \mathbf{R}^{d}\right) \mathbf{A} \boldsymbol{\xi} \cdot \boldsymbol{\xi} \geq \alpha^{\prime}|\boldsymbol{\xi}|^{2} \& \mathbf{A} \boldsymbol{\xi} \cdot \boldsymbol{\xi} \geq \frac{1}{\beta^{\prime}}|\mathbf{A} \boldsymbol{\xi}|^{2}\right\}
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New unknown vector function taking values in $\mathbf{R}^{d+1}$ :

$$
\mathrm{u}=\left[\begin{array}{c}
\mathrm{u}_{d} \\
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$$

Then the starting equation can be written as a first-order system

$$
\left\{\begin{array}{l}
\nabla_{\mathbf{x}} u_{d+1}+\mathbf{A}^{-1} \mathbf{u}_{d}=0 \\
\operatorname{div} \mathbf{u}_{d}+c u_{d+1}=f
\end{array}\right.
$$

## An example - stationary diffusion equation (cont.)

which is a Friedrichs system with the choice of

$$
\mathbf{A}_{k}=\mathbf{e}_{k} \otimes \mathbf{e}_{d+1}+\mathbf{e}_{d+1} \otimes \mathbf{e}_{k} \in \mathrm{M}_{d+1}(\mathbf{R}), \quad \mathbf{C}=\left[\begin{array}{cc}
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The graph space: $W=\mathrm{L}_{\text {div }}^{2}(\Omega) \times \mathrm{H}^{1}(\Omega)$.
Dirichlet, Neumann and Robin boundary conditions are imposed by the following choice of $V$ and $\widetilde{V}$ :

$$
\begin{aligned}
V_{D}=\widetilde{V}_{D} & :=\mathrm{L}_{\mathrm{div}}^{2}(\Omega) \times \mathrm{H}_{0}^{1}(\Omega), \\
V_{N}=\widetilde{V}_{N} & :=\left\{\left(\mathrm{u}_{d}, u_{d+1}\right)^{\top} \in W: \boldsymbol{\nu} \cdot \mathrm{u}_{d}=0\right\}, \\
V_{R} & :=\left\{\left(\mathrm{u}_{d}, u_{d+1}\right)^{\top} \in W: \boldsymbol{\nu} \cdot \mathbf{u}_{d}=\left.a u_{d+1}\right|_{\Gamma}\right\}, \\
\widetilde{V}_{R} & :=\left\{\left(\mathrm{u}_{d}, u_{d+1}\right)^{\top} \in W: \boldsymbol{\nu} \cdot \mathrm{u}_{d}=-\left.a u_{d+1}\right|_{\Gamma}\right\} .
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$L$ real Hilbert space, as before $\left(L^{\prime} \equiv L\right), T>0$
We consider an abstract Cauchy problem in $L$ :
(P)

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\left\{\begin{aligned}
\mathbf{u}^{\prime}(t)+T \mathbf{u}(t) & =\mathbf{f}(t) \\
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where

- $\mathrm{f}:\langle 0, T\rangle \longrightarrow L, \mathrm{u}_{0} \in L$ are given,
- $T$ (not depending on $t$ ) satisfies (T1), (T2) and
(T3')

$$
(\forall \varphi \in \mathcal{D}) \quad\langle(T+\tilde{T}) \varphi \mid \varphi\rangle_{L} \geqslant 0
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Numerics:
E. Burman, A. Ern, M. A. Fernandez, SIAM JNA, 2010.
D. A. Di Pietro, A. Ern, 2012.

## Semigroup setting

A priori estimate:

$$
(\forall t \in[0, T]) \quad\|\mathbf{u}(t)\|_{L}^{2} \leqslant e^{t}\left(\left\|\mathbf{u}_{0}\right\|_{L}^{2}+\int_{0}^{t}\|\mathbf{f}(s)\|_{L}^{2}\right) .
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$$

Let $\mathcal{A}: V \subseteq L \longrightarrow L, \mathcal{A}:=-\left.T\right|_{V}$
Then ( P ) becomes:
( $\mathrm{P}^{\prime}$ )

$$
\left\{\begin{array}{rl}
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Theorem. The operator $\mathcal{A}$ is an infinitesimal generator of a $C_{0}$-semigroup on $L$.

Existence and uniqueness result

Corollary. Let $T$ be an operator that satisfies (T1)-(T2) and (T3)', let $V$ be a subspace of its graph space satisfying (V1)-(V2), and $\mathrm{f} \in \mathrm{L}^{1}(\langle 0, T\rangle ; L)$.

## Existence and uniqueness result

Corollary. Let $T$ be an operator that satisfies (T1)-(T2) and (T3)', let $V$ be a subspace of its graph space satisfying (V1)-(V2), and $\mathrm{f} \in \mathrm{L}^{1}(\langle 0, T\rangle ; L)$. Then for every $\mathrm{u}_{0} \in L$ the problem $(P)$ has the unique mild solution $\mathrm{u} \in \mathrm{C}([0, T] ; L)$ given with

$$
\mathrm{u}(t)=\mathcal{T}(t) \mathrm{u}_{0}+\int_{0}^{t} \mathcal{T}(t-s) \mathrm{f}(s) d s, \quad t \in[0, T]
$$

where $(\mathcal{T}(t))_{t \geqslant 0}$ is the semigroup generated by $\mathcal{A}$.

## Existence and uniqueness result

Corollary. Let $T$ be an operator that satisfies (T1)-(T2) and (T3)', let $V$ be a subspace of its graph space satisfying (V1)-(V2), and $\mathrm{f} \in \mathrm{L}^{1}(\langle 0, T\rangle ; L)$.
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where $(\mathcal{T}(t))_{t \geqslant 0}$ is the semigroup generated by $\mathcal{A}$.
If additionally $\mathrm{f} \in \mathrm{C}([0, T] ; L) \cap\left(\mathrm{W}^{1,1}(\langle 0, T\rangle ; L) \cup \mathrm{L}^{1}(\langle 0, T\rangle ; V)\right)$ with $V$ equipped with the graph norm and $\mathrm{u}_{0} \in V$, then the above mild solution is the classical solution of $(P)$ on $[0, T\rangle$.

## Mild solution

Theorem. Let $\mathrm{u}_{0} \in L, \mathrm{f} \in \mathrm{L}^{1}(\langle 0, T\rangle ; L)$ and let

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be the mild solution of $(P)$.
Then $\mathrm{u}^{\prime}, T \mathrm{u}, \mathrm{f} \in \mathrm{L}^{1}\left(\langle 0, T\rangle ; W_{0}^{\prime}\right)$ and

$$
\mathbf{u}^{\prime}+T \mathbf{u}=\mathbf{f}
$$

in $\mathrm{L}^{1}\left(\langle 0, T\rangle ; W_{0}^{\prime}\right)$.

## Bound on solution

From

$$
\mathrm{u}(t)=\mathcal{T}(t) \mathbf{u}_{0}+\int_{0}^{t} \mathcal{T}(t-s) \mathfrak{f}(s) d s, \quad t \in[0, T]
$$

we get:

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Non-stationary Maxwell system 1/5

Let $\Omega \subseteq \mathbf{R}^{3}$ be open and bounded with a Lipschitz boundary $\Gamma$, $\mu, \varepsilon \in \mathrm{W}^{1, \infty}(\Omega)$ positive and away from zero, $\boldsymbol{\Sigma}_{i j} \in \mathrm{~L}^{\infty}\left(\Omega ; \mathrm{M}_{3}(\mathbf{R})\right)$, $i, j \in\{1,2\}$, and $\mathrm{f}_{1}, \mathrm{f}_{2} \in \mathrm{~L}^{1}\left(\langle 0, T\rangle ; \mathrm{L}^{2}\left(\Omega ; \mathbf{R}^{3}\right)\right)$.

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We consider a generalized non-stationary Maxwell system
(MS) $\left\{\begin{aligned} \mu \partial_{t} \mathrm{H}+\operatorname{rot} \mathrm{E}+\boldsymbol{\Sigma}_{11} \mathrm{H}+\boldsymbol{\Sigma}_{12} \mathrm{E}=\mathrm{f}_{1} \\ \varepsilon \partial_{t} \mathrm{E}-\operatorname{rot} \mathrm{H}+\boldsymbol{\Sigma}_{21} \mathrm{H}+\boldsymbol{\Sigma}_{22} \mathrm{E}=\mathrm{f}_{2}\end{aligned} \quad\right.$ in $\langle 0, T\rangle \times \Omega$,

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$$

where $\mathrm{E}, \mathrm{H}:[0, T\rangle \times \Omega \longrightarrow \mathbf{R}^{3}$ are unknown functions.
Change of variable

$$
\mathrm{u}:=\left[\begin{array}{l}
\mathbf{u}_{1} \\
\mathbf{u}_{2}
\end{array}\right]=\left[\begin{array}{c}
\sqrt{\mu} \mathrm{H} \\
\sqrt{\varepsilon} \mathrm{E}
\end{array}\right], \quad c:=\frac{1}{\sqrt{\mu \varepsilon}} \in \mathrm{~W}^{1, \infty}(\Omega),
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$$

turns (MS) to the Friedrichs system

$$
\partial_{t} \mathbf{u}+T \mathbf{u}=\mathrm{F},
$$

Non-stationary Maxwell system 2/5
with

$$
\begin{aligned}
& \mathbf{A}_{1}:=c\left[\begin{array}{cccccc} 
& & & 0 & 0 & 0 \\
& \mathbf{0} & & 0 & 0 & -1 \\
0 & 0 & 0 & 0 & 1 & 0 \\
0 & 0 & 1 & & \mathbf{0} & \\
0 & -1 & 0 & & &
\end{array}\right], \quad \mathbf{A}_{2}:=c\left[\begin{array}{ccccc} 
& & & 0 & 0 \\
& 0 & & 0 & 0 \\
& & & & 0 \\
& & 0 & 0 \\
0 & 0 & -1 & & \\
0 & 0 & 0 & & \mathbf{0} \\
1 & 0 & 0 & &
\end{array}\right], \\
& \mathbf{A}_{3}:=c\left[\begin{array}{cccccc} 
& & & 0 & -1 & 0 \\
& \mathbf{0} & & 1 & 0 & 0 \\
0 & 1 & 0 & & 0 & 0 \\
0 & 1 & 0 & 0 & & \mathbf{0}
\end{array}\right], \quad \mathbf{F}=\left[\begin{array}{c}
\frac{1}{\sqrt{\mu}} f_{1} \\
\frac{1}{\sqrt{\varepsilon}} f_{2}
\end{array}\right], \quad \mathbf{C}:=\ldots .
\end{aligned}
$$

Non-stationary Maxwell system 2/5
with

$$
\begin{aligned}
& \mathbf{A}_{3}:=c\left[\begin{array}{cccccc} 
& & & 0 & -1 & 0 \\
& 0 & & 0 & 0 \\
0 & 0 & 0 \\
0 & 1 & 0 & & & \\
-1 & 0 & 0 & & \mathbf{0} \\
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\end{aligned}
$$

(F1) and (F2) are satisfied (with change $v:=e^{-\lambda t} \mathbf{u}$ for large $\lambda>0$, if needed)

Non-stationary Maxwell system 3/5

The spaces involved:

$$
\begin{aligned}
L & =\mathrm{L}^{2}\left(\Omega ; \mathbf{R}^{3}\right) \times \mathrm{L}^{2}\left(\Omega ; \mathbf{R}^{3}\right) \\
W & =\mathrm{L}_{\mathrm{rot}}^{2}\left(\Omega ; \mathbf{R}^{3}\right) \times \mathrm{L}_{\mathrm{rot}}^{2}\left(\Omega ; \mathbf{R}^{3}\right) \\
W_{0} & =\mathrm{L}_{\mathrm{rot}, 0}^{2}\left(\Omega ; \mathbf{R}^{3}\right) \times \mathrm{L}_{\mathrm{rot}, 0}^{2}\left(\Omega ; \mathbf{R}^{3}\right)=\mathrm{Cl}_{W} \mathrm{C}_{c}^{\infty}\left(\Omega ; \mathbf{R}^{6}\right),
\end{aligned}
$$

where $\mathrm{L}_{\mathrm{rot}}^{2}\left(\Omega ; \mathbf{R}^{3}\right)$ is the graph space of the rot operator.

Non-stationary Maxwell system 3/5

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$$
\nu \times \mathrm{E}_{\left.\right|_{\Gamma}}=0
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$$
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V=\tilde{V} & =\left\{\mathbf{u} \in W: \boldsymbol{\nu} \times \mathbf{u}_{2}=0\right\} \\
& =\{\mathbf{u} \in W: \boldsymbol{\nu} \times \mathrm{E}=0\} \\
& =\mathrm{L}_{\mathrm{rot}}^{2}\left(\Omega ; \mathbf{R}^{3}\right) \times \mathrm{L}_{\mathrm{rot}, 0}^{2}\left(\Omega ; \mathbf{R}^{3}\right)
\end{aligned}
$$

Non-stationary Maxwell system 4/5

Theorem. Let $\mathrm{E}_{0} \in \mathrm{~L}_{\text {rot }, 0}^{2}\left(\Omega ; \mathbf{R}^{3}\right), \mathrm{H}_{0} \in \mathrm{~L}_{\text {rot }}^{2}\left(\Omega ; \mathbf{R}^{3}\right)$ and let
$\mathrm{f}_{1}, \mathrm{f}_{2} \in \mathrm{C}\left([0, T] ; \mathrm{L}^{2}\left(\Omega ; \mathbf{R}^{3}\right)\right)$ satisfy at least one of the following conditions:
$-\mathrm{f}_{1}, \mathrm{f}_{2} \in \mathrm{~W}^{1,1}\left(\langle 0, T\rangle ; \mathrm{L}^{2}\left(\Omega ; \mathbf{R}^{3}\right)\right)$;
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Then the abstract initial-boundary value problem

$$
\left\{\begin{array}{l}
\mu \mathrm{H}^{\prime}+\operatorname{rot} \mathrm{E}+\boldsymbol{\Sigma}_{11} H+\boldsymbol{\Sigma}_{12} E=\mathrm{f}_{1} \\
\varepsilon \mathrm{E}^{\prime}-\operatorname{rot} \mathrm{H}+\boldsymbol{\Sigma}_{21} H+\boldsymbol{\Sigma}_{22} E=\mathrm{f}_{2} \\
\mathrm{E}(0)=\mathrm{E}_{0} \\
\mathrm{H}(0)=\mathrm{H}_{0} \\
\boldsymbol{\nu} \times \mathrm{E}_{\left.\right|_{\Gamma}}=0
\end{array}\right.
$$

Non-stationary Maxwell system 5/5

Theorem. ... has unique classical solution given by

$$
\begin{aligned}
{\left[\begin{array}{c}
\mathbf{H} \\
\mathbf{E}
\end{array}\right](t)=} & {\left[\begin{array}{cc}
\frac{1}{\sqrt{\mu}} \mathbf{I} & \mathbf{0} \\
\mathbf{0} & \frac{1}{\sqrt{\varepsilon}} \mathbf{I}
\end{array}\right] \mathcal{T}(t)\left[\begin{array}{c}
\sqrt{\mu} \mathbf{H}_{0} \\
\sqrt{\varepsilon} \mathrm{E}_{0}
\end{array}\right] } \\
& +\left[\begin{array}{cc}
\frac{1}{\sqrt{\mu}} \mathbf{I} & \mathbf{0} \\
\mathbf{0} & \frac{1}{\sqrt{\varepsilon}} \mathbf{I}
\end{array}\right] \int_{0}^{t} \mathcal{T}(t-s)\left[\begin{array}{c}
\frac{1}{\sqrt{\mu}} \mathrm{f}_{1}(s) \\
\frac{1}{\sqrt{\varepsilon}} \mathrm{f}_{2}(s)
\end{array}\right] d s, \quad t \in[0, T],
\end{aligned}
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where $(\mathcal{T}(t))_{t \geqslant 0}$ is the contraction $C_{0}$-semigroup generated by $-T$.

## Other examples

- Symmetric hyperbolic system

$$
\left\{\begin{array}{l}
\partial_{t} \mathbf{u}+\sum_{k=1}^{d} \partial_{k}\left(\mathbf{A}_{k} \mathbf{u}\right)+\mathbf{C u}=\mathrm{f} \quad \text { in }\langle 0, T\rangle \times \mathbf{R}^{d} \\
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- Non-stationary div-grad problem

$$
\left\{\begin{array}{l}
\partial_{t} \mathbf{q}+\nabla p=\mathrm{f}_{1} \quad \text { in }\langle 0, T\rangle \times \Omega, \quad \Omega \subseteq \mathbf{R}^{d}, \\
\frac{1}{c_{0}^{2}} \partial_{t} p+\operatorname{div} \mathbf{q}=f_{2} \quad \text { in }\langle 0, T\rangle \times \Omega \\
p_{\left.\right|_{\partial \Omega}}=0, \quad p(0)=p_{0}, \quad \mathrm{q}(0)=\mathrm{q}_{0}
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p_{\left.\right|_{\partial \Omega}}=0, \quad p(0)=p_{0}, \quad \mathbf{q}(0)=\mathrm{q}_{0}
\end{array}\right.
$$

- Wave equation

$$
\left\{\begin{array}{l}
\partial_{t t} u-c^{2} \triangle u=f \quad \text { in }\langle 0, T\rangle \times \mathbf{R}^{d} \\
u(0, \cdot)=u_{0}, \quad \partial_{t} u(0, \cdot)=u_{0}^{1}
\end{array}\right.
$$

Friedrichs systems in a complex Hilbert space

Let $L$ be a complex Hilbert space, $L^{\prime} \equiv L$ its antidual, $\mathcal{D} \subseteq L, T, \tilde{T}: L \longrightarrow L$ linear operators that satisfy (T1)-(T3) (or T3' instead of T3).

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Technical differences with respect to the real case, but results remain the same...

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For the classical Friedrichs operator we require

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\begin{equation*}
\text { matrix functions } \mathbf{A}_{k} \text { are selfadjoint: } \mathbf{A}_{k}=\mathbf{A}_{k}^{*} \tag{F1}
\end{equation*}
$$

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$$
\begin{equation*}
\left(\exists \mu_{0}>0\right) \quad \mathbf{C}+\mathbf{C}^{*}+\sum_{k=1}^{d} \partial_{k} \mathbf{A}_{k} \geqslant 2 \mu_{0} \mathbf{I} \quad(\text { ae on } \Omega) \tag{F1}
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For the classical Friedrichs operator we require
(F1) matrix functions $\mathbf{A}_{k}$ are selfadjoint: $\mathbf{A}_{k}=\mathbf{A}_{k}^{*}$,
(F2) $\quad\left(\exists \mu_{0}>0\right) \quad \mathbf{C}+\mathbf{C}^{*}+\sum_{k=1}^{d} \partial_{k} \mathbf{A}_{k} \geqslant 2 \mu_{0} \mathbf{I} \quad($ ae on $\Omega)$,
and again (F1)-(F2) imply (T1)-(T3).

Application to Dirac system 1/2

We consider the Cauchy problem

## Application to Dirac system 1/2

We consider the Cauchy problem
(DS) $\left\{\begin{array}{l}\partial_{t} \mathbf{u}+\sum_{k=1}^{3} \mathbf{A}_{k} \partial_{k} \mathbf{u}+\mathbf{C u}=\mathrm{f} \quad \text { in }\langle 0, T\rangle \times \mathbf{R}^{3}, \\ \mathbf{u}(0)=\mathbf{u}_{0},\end{array}\right.$

## Application to Dirac system 1/2

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$$
\text { (DS) } \quad\left\{\begin{array}{l}
\partial_{t} \mathbf{u}+\overbrace{\sum_{k=1}^{3} \mathbf{A}_{k} \partial_{k} \mathbf{u}+\mathbf{C u}}^{T \mathrm{u}}=\mathrm{f} \quad \text { in }\langle 0, T\rangle \times \mathbf{R}^{3}, \\
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where u: $[0, T\rangle \times \mathbf{R}^{3} \longrightarrow \mathbf{C}^{4}$ is an unknown function, $\mathrm{u}_{0}: \mathbf{R}^{3} \longrightarrow \mathbf{C}^{4}$, $\mathrm{f}:[0, T\rangle \times \mathbf{R}^{3} \longrightarrow \mathbf{C}^{4}$ are given, and

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$$
\mathbf{A}_{k}:=\left[\begin{array}{cc}
\mathbf{0} & \boldsymbol{\sigma}_{k} \\
\boldsymbol{\sigma}_{k} & \mathbf{0}
\end{array}\right], k \in 1 . .3, \quad \mathbf{C}:=\left[\begin{array}{cc}
c_{1} \mathbf{I} & \mathbf{0} \\
\mathbf{0} & c_{2} \mathbf{I}
\end{array}\right],
$$

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\end{array}\right],
$$

where

$$
\begin{align*}
& \qquad \boldsymbol{\sigma}_{1}:=\left[\begin{array}{ll}
0 & 1 \\
1 & 0
\end{array}\right], \quad \boldsymbol{\sigma}_{2}:=\left[\begin{array}{cc}
0 & -i \\
i & 0
\end{array}\right], \quad \boldsymbol{\sigma}_{3}:=\left[\begin{array}{cc}
1 & 0 \\
0 & -1
\end{array}\right], \\
& \text { are Pauli matrices, and } c_{1}, c_{2} \in \mathrm{~L}^{\infty}\left(\mathbf{R}^{3} ; \mathbf{C}\right) \tag{F1}
\end{align*}
$$

## Application to Dirac system 2/2

Theorem. Let $\mathrm{u}_{0} \in W$ and let $\mathrm{f} \in \mathrm{C}\left([0, T] ; \mathrm{L}^{2}\left(\mathbf{R}^{3} ; \mathbf{C}^{4}\right)\right)$ satisfies at least one of the following conditions:

- $\mathrm{f} \in \mathrm{W}^{1,1}\left(\langle 0, T\rangle ; \mathrm{L}^{2}\left(\mathbf{R}^{3} ; \mathbf{C}^{4}\right)\right)$;
- $\mathrm{f} \in \mathrm{L}^{1}(\langle 0, T\rangle ; W)$.

Then the abstract Cauchy problem

$$
\left\{\begin{array}{l}
\partial_{t} \mathbf{u}+\sum_{k=1}^{3} \mathbf{A}_{k} \partial_{k} \mathbf{u}+\mathbf{C u}=\mathrm{f} \\
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\end{array}\right.
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has unique classical solution given with

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\mathrm{u}(t)=\mathcal{T}(t) \mathrm{u}_{0}+\int_{0}^{t} \mathcal{T}(t-s) \mathrm{f}(s) d s, \quad t \in[0, T]
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## TODO: Time-dependent coefficients

The operator $T$ depends on $t$ (i.e. the matrix coefficients $\mathbf{A}_{k}$ and $\mathbf{C}$ depend on $t$ if $T$ is a classical Friedrichs operator):

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\left\{\begin{aligned}
\mathbf{u}^{\prime}(t)+T(t) \mathbf{u}(t) & =\mathbf{f}(t) \\
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- Semigroup theory can treat time-dependent case, but conditions that ensure existence/uniqueness result are rather complicated to verify...


## TODO: Semilinear problem

Consider

$$
\left\{\begin{aligned}
\mathbf{u}^{\prime}(t)+T \mathbf{u}(t) & =\mathbf{f}(t, \mathbf{u}(t)) \\
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where $\mathrm{f}:[0, T\rangle \times L \longrightarrow L$.

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- semigroup theory gives existence and uniqueness of solution
- it requires (locally) Lipschitz continuity of $f$ in variable $u$
- if $L=\mathrm{L}^{2}$ it is not appropriate assumption, as power functions do not satisfy it; $L=\mathrm{L}^{\infty}$ is better...

TODO: Banach space setting

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Let $L$ be a reflexive complex Banach space, $L^{\prime}$ its antidual, $\mathcal{D} \subseteq L$, $T, \tilde{T}: \mathcal{D} \longrightarrow L^{\prime}$ linear operators that satisfy a modified versions of (T1)-(T3)

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(T1)

$$
(\forall \varphi, \psi \in \mathcal{D}) \quad{ }_{L^{\prime}}\langle T \varphi, \psi\rangle_{L}=\overline{L_{L^{\prime}}\langle\tilde{T} \psi, \phi\rangle_{L}} .
$$

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$$

Technical differences with Hilbert space case, but results remain essentially the same for the stationary case...

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Problems:

- in the classical case (F1)-(F2) need not to imply (T2) and (T3): instead of (T3) we get

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\mathrm{L}_{p^{p^{\prime}}}\langle(\tilde{T}) \varphi, \varphi\rangle_{\mathrm{L}^{p}} \geqslant\|\varphi\|_{\mathrm{L}^{2}}
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- for semigroup treatment of non-stationary case we need to have $T: \mathcal{D} \subseteq L \longrightarrow L$

Why should one be interested in Friedrichs systems?
Symmetric hyperbolic systems
Symmetric positive systems
Classical theory
Boundary conditions for Friedrichs systems
Existence, uniqueness, well-posedness
Abstract formulation
Graph spaces
Cone formalism of Ern, Guermond and Caplain
Interdependence of different representations of boundary conditions
Kreĭn space formalism
Kreĭn spaces
Equivalence of boundary conditions
What can we say for the Friedrichs operator now?
Sufficient assumptions
An example: elliptic equation
Other second order equations
Two-field theory
Non-stationary theory
Homogenisation of Friedrichs systems
Homogenisation
Examples: Stationary diffusion and heat equation
Concluding remarks

## Homogenisation setting

Krešimir Burazin, Marko Vrdoljak: Homogenisation theory for Friedrichs systems, Comm. Pure Appl. Analysis 13 (2014) 1017-1044.

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W:=\{\mathbf{u} \in L: T \mathbf{u} \in L\}=\left\{\mathbf{u} \in L: T_{0} \mathbf{u} \in L\right\}
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Moreover, we have equivalence of norms $\left(\gamma=\sqrt{\max \left\{2,1+2 \beta^{2}\right\}}\right)$ :

$$
\|\mathbf{u}\|_{T} \leqslant \gamma\|\mathbf{u}\|_{T_{0}} \leqslant \gamma^{2}\|\mathbf{u}\|_{T}, \quad \text { for any } \mathbf{C} .
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## Boundary operator and a priori bound

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If $V$ is a subspace of $W$ that satisfies $(\mathrm{V})$, well-posedness result implies that $T_{\left.\right|_{V}}: V \longrightarrow L$ is an isomorphism, with

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Therefore, for fixed $T_{0}$ and $V$ satisfying $(\mathrm{V})$, we have a priori bound

$$
(\exists c>0)\left(\forall \mathbf{C} \in \mathcal{M}_{r}(\alpha, \beta ; \Omega)\right)(\forall \mathbf{u} \in V) \quad\|\mathbf{u}\|_{T_{0}} \leq c\left\|\left(\mathcal{L}_{0}+\mathbf{C}\right) \mathbf{u}\right\|_{L}
$$

Note that constant $c$ depends only on $T_{0}, \alpha$ and $\beta$.

## $H$-convergence

In the sequel $\mathcal{L}_{0}=\sum_{k=1}^{d} \mathbf{A}_{k} \partial_{k}$ and $V$ are fixed.
Definition ( $H$-convergence for Friedrichs systems)
We say that a sequence $\left(\mathbf{C}_{n}\right)$ in $\mathcal{M}_{r}(\alpha, \beta ; \Omega) H$-converges to $\mathbf{C} \in \mathcal{M}_{r}\left(\alpha^{\prime}, \beta^{\prime} ; \Omega\right)$ with respect to $T_{0}$ and $V$ if, for any $\mathrm{f} \in L$, the sequence $\left(\mathrm{u}_{n}\right)$ in $V$ defined by $\mathrm{u}_{n}:=T_{n}^{-1} \mathrm{f} \in V$, with $T_{n}=\mathcal{L}_{0}+\mathbf{C}_{n}$, satisfies

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## $H$-convergence and topology. . .

## Theorem

Let $F=\left\{\mathrm{f}_{n}: n \in \mathbf{N}\right\}$ be a dense countable family in $\mathrm{L}^{2}\left(\Omega ; \mathbf{R}^{r}\right)$,
$\mathbf{C}, \mathbf{D} \in \mathcal{M}_{r}(\alpha, \beta ; \Omega)$, and $\mathrm{u}_{n}, \mathrm{v}_{n} \in V$ solutions of $\left(T_{0}+\mathbf{C}\right) \mathrm{u}_{n}=\mathrm{f}_{n}$ and $\left(T_{0}+\mathbf{D}\right) \mathrm{v}_{n}=\mathrm{f}_{n}$, respectively. Furthermore, let

$$
d(\mathbf{C}, \mathbf{D}):=\sum_{n=1}^{\infty} 2^{-n} \frac{\left\|\mathbf{u}_{n}-\mathrm{v}_{n}\right\|_{\mathrm{H}^{-1}\left(\Omega ; \mathbf{R}^{r}\right)}+\left\|\mathbf{C u}_{n}-\mathbf{D} \mathrm{v}_{n}\right\|_{\mathrm{H}^{-1}\left(\Omega ; \mathbf{R}^{r}\right)}}{\left\|\mathbf{f}_{n}\right\|_{\mathrm{L}^{2}\left(\Omega ; \mathbf{R}^{r}\right)}}
$$

Then the function $d: \mathcal{M}_{r}(\alpha, \beta ; \Omega) \times \mathcal{M}_{r}(\alpha, \beta ; \Omega) \longrightarrow \mathbf{R}$ forms a metric on the set $\mathcal{M}_{r}(\alpha, \beta ; \Omega)$, and the $H$-convergence is equivalent to the sequential convergence in this metric space.

## Compactness assumptions

Additional assumptions: for every sequence $\mathbf{C}_{n} \in \mathcal{M}_{r}(\alpha, \beta ; \Omega)$ and every $\mathrm{f} \in L$, the sequence $\mathrm{u}_{n} \in V$ defined by $\mathrm{u}_{n}:=\left(T_{0}+\mathbf{C}_{n}\right)^{-1} \mathrm{f}$ satisfies the following: if ( $\mathrm{u}_{n}$ ) weakly converges to u in $W$, then also

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\begin{equation*}
{ }_{W}\left\langle D \mathrm{u}_{n}, \mathrm{u}_{n}\right\rangle_{W} \longrightarrow{ }_{W^{\prime}}\langle D \mathrm{u}, \mathrm{u}\rangle_{W}, \tag{K1}
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or
(K2)

$$
\left(\forall \varphi \in \mathrm{C}_{c}^{\infty}(\Omega)\right) \quad\left\langle T_{0} \mathbf{u}_{n} \mid \varphi \mathbf{u}_{n}\right\rangle_{L} \longrightarrow\left\langle T_{0} \mathbf{u} \mid \varphi \mathbf{u}\right\rangle_{L} .
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Theorem
For fixed $T_{0}$ and $V$, if family $\mathcal{M}_{r}(\alpha, \beta ; \Omega)$ satisfies (K1) and (K2), then it is compact with respect to $H$-convergence, i.e. from any sequence $\left(\mathbf{C}_{n}\right)$ in $\mathcal{M}_{r}(\alpha, \beta ; \Omega)$ one can extract a $H$-converging subsequence whose limit belongs to $\mathcal{M}_{r}(\alpha, \beta ; \Omega)$.

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The proof follows the original proof of Spagnolo in the case of parabolic $G$-convergence.

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-\operatorname{div}(\mathbf{A} \nabla u)+c u=f
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V_{R} & :=\left\{\left(\mathbf{u}_{d}, u_{d+1}\right)^{\top} \in W: \boldsymbol{\nu} \cdot \mathbf{u}_{d}=\left.a u_{d+1}\right|_{\Gamma}\right\}, \\
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$$

$$
\begin{aligned}
& W^{\prime}\langle D \mathrm{u}, \mathbf{u}\rangle_{W}=2_{{ }_{\mathrm{H}^{-\frac{1}{2}}}\left\langle\boldsymbol{\nu} \cdot \mathbf{u}_{d}, u_{d+1}\right\rangle_{\mathrm{H}^{\frac{1}{2}}}} \\
&=\left\{\begin{aligned}
0 & \ldots
\end{aligned} \begin{array}{rl} 
& \text { Dirichlet or Neumann } \\
2 a\left\|u_{d+1}\right\|_{\mathrm{L}^{2}(\Gamma)}^{2} & \ldots
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& W^{\prime}\langle D \mathrm{u}, \mathbf{u}\rangle_{W}=2_{{ }_{\mathrm{H}^{-\frac{1}{2}}}\left\langle\boldsymbol{\nu} \cdot \mathbf{u}_{d}, u_{d+1}\right\rangle_{\mathrm{H}^{\frac{1}{2}}}} \\
&=\left\{\begin{aligned}
0 & \ldots
\end{aligned} \begin{array}{rl} 
& \text { Dirichlet or Neumann } \\
2 a\left\|u_{d+1}\right\|_{\mathrm{L}^{2}(\Gamma)}^{2} & \ldots
\end{array}\right. \\
& \text { Robin } \ldots W=\mathrm{L}_{\mathrm{div}}^{2}(\Omega) \times \mathrm{H}^{1}(\Omega)
\end{aligned}
$$

## Properties (K1) and (K2)

(K1) For any sequence ( $u_{n}$ ) in $V$

$$
\mathbf{u}_{n} \longrightarrow \mathbf{u} \quad \Longrightarrow W^{\prime}\left\langle D \mathbf{u}_{n}, \mathbf{u}_{n}\right\rangle_{W} \longrightarrow W^{\prime}\langle D \mathbf{u}, \mathbf{u}\rangle_{W}
$$

${ }_{W^{\prime}}\langle D \mathrm{u}, \mathrm{u}\rangle_{W}=2_{\mathrm{H}^{-\frac{1}{2}}}\left\langle\boldsymbol{\nu} \cdot \mathbf{u}_{d}, u_{d+1}\right\rangle_{\mathrm{H}^{\frac{1}{2}}}$

$$
=\left\{\begin{aligned}
0 & \ldots \\
& \text { Dirichlet or Neumann } \\
2 a\left\|u_{d+1}\right\|_{\mathrm{L}^{2}(\Gamma)}^{2} & \ldots
\end{aligned}\right.
$$

(K2) For any sequence ( $\mathrm{u}_{n}$ ) in $V$ and any $\varphi \in \mathrm{C}_{c}^{\infty}(\Omega)$

$$
\mathbf{u}_{n} \longrightarrow \mathbf{u} \quad \Longrightarrow \quad\left\langle T_{0} \mathbf{u}_{n} \mid \varphi \mathbf{u}_{n}\right\rangle_{L} \longrightarrow\left\langle T_{0} \mathbf{u} \mid \varphi \mathbf{u}\right\rangle_{L}
$$

## Compensated compactness

$$
\begin{aligned}
\left\langle T_{0} \mathbf{u}_{n} \mid \varphi \mathbf{u}_{n}\right\rangle_{L} & =\int_{\Omega} \sum_{k=1}^{d} \mathbf{A}_{k} \partial_{k} \mathbf{u}_{n} \cdot \varphi \mathbf{u}_{n} d \mathbf{x}, \\
& =-\frac{1}{2} \int_{\Omega} \partial_{k} \varphi \sum_{k=1}^{d} \mathbf{A}_{k} \mathbf{u}_{n} \cdot \mathbf{u}_{n} d \mathbf{x}
\end{aligned}
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$$

Theorem (Quadratic theorem)
For $\mathbf{A}_{k} \in \mathrm{M}_{q, p}(\mathbf{R})$ let $\Lambda:=\left\{\boldsymbol{\lambda} \in \mathbf{R}^{p}:(\exists \boldsymbol{\xi} \neq \mathbf{0}) \quad \sum_{k=1}^{d} \xi_{k} \mathbf{A}_{k} \boldsymbol{\lambda}=\mathbf{0}\right\}$

$$
Q(\boldsymbol{\lambda}):=\mathbf{Q} \boldsymbol{\lambda} \cdot \boldsymbol{\lambda}, \text { such that } Q=0 \text { on } \Lambda \text {, }
$$

$$
\mathrm{u}_{n} \longrightarrow \mathrm{u} \text { weakly in } \mathrm{L}^{2}\left(\Omega ; \mathbf{R}^{p}\right),
$$

$$
\left(\sum_{k=1}^{d} \mathbf{A}_{k} \partial_{k} \mathbf{u}_{n}\right) \quad \text { is relatively compact in } \mathrm{H}^{-1}\left(\Omega ; \mathbf{R}^{q}\right) .
$$

Then $Q \circ \mathbf{u}_{n} \longrightarrow Q \circ \mathbf{u}$ in $\mathcal{D}^{\prime}(\Omega)$.

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& =-\frac{1}{2} \int_{\Omega} \partial_{k} \varphi \sum_{k=1}^{d} \overbrace{\mathbf{A}_{k} \mathbf{u}_{n} \cdot \mathbf{u}_{n}} d \mathbf{x}
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## Proof of (K2)

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\begin{aligned}
\sum_{k=1}^{d} \xi_{k} \mathbf{A}_{k} \boldsymbol{\lambda} & =\left[\begin{array}{c}
\lambda_{d+1} \xi_{1} \\
\vdots \\
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\sum_{k=1}^{d} \lambda_{k} \xi_{k}
\end{array}\right] \quad \Longrightarrow \quad \Lambda \ldots \lambda_{d+1}=0 \\
Q(\lambda) & =\mathbf{A}_{i} \boldsymbol{\lambda} \cdot \boldsymbol{\lambda}=2 \lambda_{i} \lambda_{d+1}=0, \quad \lambda \in \Lambda
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## Comparison with classical $H$-convergence

$$
\begin{aligned}
\mathbf{C}_{n}=\left[\begin{array}{cc}
\left(\mathbf{A}^{n}\right)^{-1} & 0 \\
0^{\top} & c_{n}
\end{array}\right] & \in \mathcal{M}_{d+1}(\alpha, \beta ; \Omega) \\
& \Longleftrightarrow\left\{\begin{array}{lll}
\mathbf{C}_{n}(\mathbf{x}) \boldsymbol{\xi} \cdot \boldsymbol{\xi} & \geq \alpha|\boldsymbol{\xi}|^{2} \\
\mathbf{C}_{n}(\mathbf{x}) \boldsymbol{\xi} \cdot \boldsymbol{\xi} & \geq & \frac{1}{\beta}\left|\mathbf{C}_{n}(\mathbf{x}) \boldsymbol{\xi}\right|^{2}
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\mathbf{A}^{n}(\mathbf{x}) \boldsymbol{\xi} \cdot \boldsymbol{\xi} & \geq & \alpha\left|\mathbf{A}^{n}(\mathbf{x}) \boldsymbol{\xi}\right|^{2}
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At a subsequence $\mathbf{C}_{n} \xrightarrow{H} \mathbf{C}$, by compactness theorem.

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- Has the limit $\mathbf{C}$ the same structure?


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$$

At a subsequence $\mathbf{C}_{n} \xrightarrow{H} \mathbf{C}$, by compactness theorem.

- Has the limit $\mathbf{C}$ the same structure?
- Can we make a connection with $H$-converging (in classical sense) subsequence of $\left(\mathbf{A}^{n}\right)$ ?


## Characterisation of the $H$-limit

## Theorem

For the Friedrichs system corresponding to the stationary diffusion equation, a sequence $\left(\mathbf{C}_{n}\right)$ in $\mathcal{M}_{d+1}(\alpha, \beta ; \Omega)$ of the form

$$
\mathbf{C}_{n}=\left[\begin{array}{cc}
\left(\mathbf{A}^{n}\right)^{-1} & 0 \\
0^{\top} & c_{n}
\end{array}\right]
$$

$H$-converges with respect to $\mathcal{L}_{0}$ and $V_{D}$ if and only if $\left(\mathbf{A}^{n}\right)$ classically $H$-converges to some $\mathbf{A}$ and $\left(c_{n}\right) L^{\infty}$ weakly $*$ converges to some $c$. In that case, the $H$-limit is the matrix function

$$
\mathbf{C}=\left[\begin{array}{cc}
\mathbf{A}^{-1} & 0 \\
0^{\top} & c
\end{array}\right]
$$

Heat equation as Friedrichs system
$\Omega \subseteq \mathbf{R}^{d}$ open and bounded set with Lipschitz boundary $\Gamma, T>0$ and $\Omega_{T}:=\Omega \times\langle 0, T\rangle$

$$
\partial_{t} u_{n}-\operatorname{div}_{\mathbf{x}}\left(\mathbf{A}^{n} \nabla_{\mathbf{x}} u_{n}\right)+c u_{n}=f \quad \text { in } \Omega_{T}
$$

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\mathbf{u}_{n}=\left[\begin{array}{c}
\mathbf{u}_{d n} \\
u_{d+1}
\end{array}\right]=\left[\begin{array}{c}
-\mathbf{A}^{n} \nabla_{\mathbf{x}} u_{n} \\
u_{n}
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\end{gathered}
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$$

The matrices $\mathbf{A}_{k}=\mathbf{e}_{k} \otimes \mathbf{e}_{d+1}+\mathbf{e}_{d+1} \otimes \mathbf{e}_{k} \in \mathrm{M}_{d+1}(\mathbf{R}), k=1, \ldots d$, $\mathbf{A}_{d+1}=\mathbf{e}_{d+1} \otimes \mathbf{e}_{d+1}$ and

$$
\begin{aligned}
\mathbf{C}_{n} & =\left[\begin{array}{cc}
\left(\mathbf{A}^{n}\right)^{-1} & 0 \\
0^{\top} & c
\end{array}\right] \\
T_{0}\left[\begin{array}{c}
\mathrm{u}_{d} \\
u_{d+1}
\end{array}\right] & =\left[\begin{array}{c}
\nabla_{\mathbf{x}} u_{d+1} \\
\partial_{t} u_{d+1}+\operatorname{div}_{\mathbf{x}} \mathbf{u}_{d}
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\end{aligned}
$$

Graph space

$$
W=\left\{\mathrm{u} \in \mathrm{~L}_{\mathrm{div}}^{2}\left(\Omega_{T}\right): u_{d+1} \in \mathrm{~L}^{2}\left(0, T ; \mathrm{H}^{1}(\Omega)\right)\right\}
$$

## Compactness result

Dirichlet boundary conditions with zero initial value:

$$
\begin{gathered}
V=\left\{\mathrm{u} \in W: u_{d+1} \in \mathrm{~L}^{2}\left(0, T ; \mathrm{H}_{0}^{1}(\Omega)\right), \quad u_{d+1}(\cdot, 0)=0 \text { a.e. on } \Omega\right\}, \\
\widetilde{V}=\left\{\mathrm{v} \in W: v^{u} \in \mathrm{~L}^{2}\left(0, T ; \mathrm{H}_{0}^{1}(\Omega)\right), \quad v^{u}(\cdot, T)=0 \text { a.e. on } \Omega\right\} .
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(K1):

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{ }_{W}\langle D \mathrm{u}, \mathrm{u}\rangle_{W}=\left\|u_{d+1}(\cdot, T)\right\|_{\mathrm{L}^{2}(\Omega)}^{2} .
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(K2): similarly to stationary diffusion equation: $\Lambda=\left\{\boldsymbol{\lambda} \in \mathbf{R}^{d+1}: \lambda_{d+1}=0\right\}$

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$\Longrightarrow \mathcal{M}_{d+1}(\alpha, \beta ; \Omega)$ is compact with $H$-topology for given $\mathcal{L}_{0}$ and $V$
Comparison with classical parabolic H -convergence. . .

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Comparison with classical parabolic H -convergence. . . similarly as for stationary diffusion equation.

## $G$-convergence

Instead of $\mathbf{C}_{n} \in \mathcal{M}_{r}(\alpha, \beta ; \Omega)$ we take

$$
\begin{aligned}
\mathcal{C}_{n} \in \mathcal{F}(\alpha, \beta ; \Omega):= & \{\mathcal{C} \in \mathcal{L}(L):(\forall \mathbf{u} \in L) \\
& \left.\langle\mathcal{C} \mathbf{u} \mid \mathbf{u}\rangle_{L} \geq \alpha\|\mathbf{u}\|_{L}^{2} \quad \& \quad\langle\mathcal{C} \mathbf{u} \mid \mathbf{u}\rangle_{L} \geq \frac{1}{\beta}\|\mathcal{C} \mathbf{u}\|_{L}^{2}\right\} .
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\end{aligned}
$$

Definition ( $G$-convergence for Friedrichs systems)
For $\mathcal{C}_{n} \in \mathcal{F}(\alpha, \beta ; \Omega)$, we say that a sequence of isomorphisms $T_{n}:=T_{0}+\mathcal{C}_{n}: V \rightarrow L G$-converges to an isomorphism $T:=T_{0}+\mathcal{C}: V \rightarrow L$, for some $\mathcal{C} \in \mathcal{F}\left(\alpha^{\prime}, \beta^{\prime} ; \Omega\right)$ if

$$
(\forall \mathrm{f} \in L) \quad T_{n}^{-1} \mathrm{f} \longrightarrow T^{-1} \mathrm{f} \text { in } W
$$

## Theorem

For fixed $T_{0}$ and $V$, if family $\mathcal{F}(\alpha, \beta ; \Omega)$ satisfies (K1), then for any sequence $\left(\mathcal{C}_{n}\right)$ in $\mathcal{F}(\alpha, \beta ; \Omega)$ there exists a subsequence of $T_{n}:=T_{0}+\mathcal{C}_{n}$ which $G$-converges to $T:=T_{0}+\mathcal{C}$ with $\mathcal{C} \in \mathcal{F}(\alpha, \beta ; \Omega)$.

Why should one be interested in Friedrichs systems?
Symmetric hyperbolic systems
Symmetric positive systems
Classical theory
Boundary conditions for Friedrichs systems
Existence, uniqueness, well-posedness
Abstract formulation
Graph spaces
Cone formalism of Ern, Guermond and Caplain
Interdependence of different representations of boundary conditions
Kreĭn space formalism
Kreĭn spaces
Equivalence of boundary conditions
What can we say for the Friedrichs operator now?
Sufficient assumptions
An example: elliptic equation
Other second order equations
Two-field theory
Non-stationary theory
Homogenisation of Friedrichs systems
Homogenisation
Examples: Stationary diffusion and heat equation
Concluding remarks

## Open problems ...

- Find all representations of a particular equation in the form of a Friedrichs system.
- Application to other equations of practical importance (mixed-type problems).
- Compare the results to those already known in the classical setting.


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