# Analysis of a system of elliptic partial differential equations and its possible boundary conditions when discretized with Hermite and Lagrange elements 

av
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## Chapter 1

## Introduction

This thesis examines a system of elliptic partial differential equations. The system itself is not elliptic. It will be called "the simplified $k-\epsilon$ model" and is defined as:

$$
\left\{\begin{array}{cc}
\epsilon-\Delta k= & f_{1}(x)  \tag{1.1}\\
\text { on } \Omega \\
-\Delta \epsilon+k= & f_{2}(x)
\end{array} \text { on } \Omega .\right.
$$

There are two main goals in this thesis. The first goal is to determine if a previously unused set of boundary conditions is numerically stable. The second goal is to make a stable implementation (1.1) using a mixed finite element method. If these goals are reached, the lessons learned from examining the simplified model can be valuable when implementing the full $k$ - $\epsilon$-model.

The new set of boundary conditions (1.3) will be compared to a set of boundary conditions (1.2) from the literature [1]

$$
\begin{align*}
& B C_{1}= \begin{cases}\epsilon(x)= & g_{1}(x) \\
k(x)= & g_{2}(x) \\
\text { on } \partial \Omega\end{cases}  \tag{1.2}\\
& B C_{2}=\left\{\begin{array}{cc}
\epsilon(x)= & g_{1}(x) \\
\text { on } \partial \Omega
\end{array}\right.  \tag{1.3}\\
& \frac{\partial \epsilon}{\partial n}=g_{2}(x) \\
& \text { on } \partial \Omega
\end{align*}
$$

Previous work on the $k$ - $\epsilon$-model by implementing the finite element method discovered issues with convergence of the numerical scheme [2,3, Smith, R.M]. Smith used a continuous Lagrange approach to the problem, which only allows for one boundary condition for each variable.

The problem (1.1) is a simplification of the more complex turbulence problem called the $k-\epsilon$ - model (1.4):

$$
\begin{align*}
\partial_{t} k+\bar{u} \nabla k-\frac{1}{2} c_{\mu} \frac{k^{2}}{\epsilon}\left|\nabla \bar{u}+\nabla \bar{u}^{T}\right|^{2}-\nabla \cdot\left(c_{\mu} \frac{k^{2}}{\epsilon} \nabla k\right)+\epsilon= & 0 \\
\partial_{t} \epsilon+\bar{u} \nabla \epsilon-\frac{1}{2} c_{1} \frac{k^{2}}{\epsilon}\left|\nabla \bar{u}+\nabla \bar{u}^{T}\right|^{2}-\nabla \cdot\left(c_{\epsilon} \frac{k^{2}}{\epsilon} \nabla \epsilon\right)+c_{2} \frac{\epsilon^{2}}{k}= & 0 . \tag{1.4}
\end{align*}
$$

The $k-\epsilon$ - model is one of the most commonly used turbulence models. It consists of two transport equations and models the transport of the variables $k, \epsilon$. The variable $k$ represents the turbulent kinetic energy, while $\epsilon$ represents the turbulent dissipation. A new set of possible boundary conditions should be useful for using the model in engineering problems.

In the study of the simplified $k-\epsilon$ - model, the finite element method (FEM) is utilized. Both regular and mixed finite element methods are used. The work on mixed finite elements largely depends on two publications [4, 5].

The element types Lagrange-2, Lagrange-3 and Hermite elements are used. The Lagrange-elements are chosen because they are $H^{1}(\Omega)$ conforming and are among the simplest and most used elements. The Hermite elements are $H^{2}(\Omega)$ conforming and among the simplest elements to conform with $H^{2}(\Omega)$. Hermite elements are crucial because they allow for two boundary conditions to be set. They are therefore needed to examine what boundary conditions can be set. These two element types will be used when examining the stability of the simplified $k-\epsilon$ - model.

Hermite elements are not commonly used. One of the reasons is rounding error issues pertaining to mesh refinement. These issues are solved in this thesis in chapter 6 by scaling the basis functions. Since they are not commonly used, there is not a lot of literature on the matter, thus most of chapter 7 through 9 is original research.

In chapter two, the thesis enumerates a list of boundary conditions known to work. The use of a different type of boundary conditions is potentially useful when applying the $k-\epsilon$ model to real life problems where the experimental information about the boundary conditions is limited.

Time is excluded from the simplified models. To implement time, it is necessary to first calculate the current state of the system, and then use the time information to calculate the next time step. If the calculation of the current state is inaccurate, it is not possible to calculate the correct states for later times. If the approach proves useful, then adding time to the models can be done.

The analysis of condition numbers and the preconditioning of the systems is based on the papers $[6,7]$. The preconditioner is constructed so that when it is combined with the differential operator becomes an automorphi. Analyzing the preconditioned form will say something about how well defined the differential form is and how stable the solutions are.

For the Finite Element Method the books [8, 9] are used.

## Chapter 2

## Arriving at the problem

For a sense of completeness we explain the assumptions, reasoning and ideas that leads to the $k-\epsilon$ - model. We follow [1, p. 40-41].

The problem (1.1) is a simplification of the more complex turbulence problem called the $k-\epsilon$ - model (2.5). This is derived from the general Navier-Stokes equations for incompressible flows (2.1)

$$
\begin{align*}
\nabla \cdot u & =0 \\
\partial_{t} u+u \nabla u+\nabla p-v \Delta u & =\frac{f}{\rho} . \tag{2.1}
\end{align*}
$$

with $v$ as the kinematic viscosity, $p=P / \rho$ is the reduced pressure, $\rho$ is a density field, $u$ is a velocity vector field, $P$ is a pressure field and $f$ represents external forces. Consider (2.1) with random initial data $u^{0}=\bar{u}^{0}+u^{\prime 0}$, where $\bar{u}$ represents the expected value and $u^{00}$ is the random element. Taking the expected value of the entire Navier-Stokes then leads to

$$
\begin{aligned}
\nabla \cdot \bar{u} & =0 \\
\partial_{t} \bar{u}+\nabla \cdot \overline{\left(\bar{u}+u^{\prime}\right) \otimes\left(\bar{u}+u^{\prime}\right)}+\nabla \bar{p}-v \Delta \bar{u} & =\bar{f}
\end{aligned}
$$

which is the same as

$$
\begin{array}{cc}
\nabla \cdot \bar{u}=0, & R=-\overline{u^{\prime} \otimes u^{\prime}}  \tag{2.2}\\
\partial_{t} \bar{u}+\nabla \cdot(\bar{u} \otimes \bar{u})+ & \nabla \bar{p}-v \Delta \bar{u}=\bar{f}+\nabla \cdot R .
\end{array}
$$

With the following assumptions, it is possible to derive the $k-\epsilon$ equations:

- Frame invariance and 2D mean flow, $v_{T}$ a polynomial function of $k, \epsilon$.
- $u^{\prime 2}$ and $\left|\nabla \times u^{\prime}\right|^{2}$ are passive scalars when convected by $\bar{u}+u^{\prime}$.
- Ergodicity allows statistical averages to be replaced by space averages.
- Local isotropy of the turbulence at level of small scales.
- A Reynold hypothesis for $\overline{\nabla \times u^{\prime} \otimes \nabla \times u^{\prime}}$.
- A closure hypothesis $\left|\nabla \times \nabla \times u^{\prime}\right|^{2}=c_{2} \frac{\epsilon^{2}}{k}$.

The set of equations used to find $R$, the turbulence in the flow, is

$$
\begin{equation*}
k=\frac{1}{2} \overline{\left|u^{\prime}\right|^{2}} \quad \epsilon=\frac{v}{2} \overline{\left|\nabla u^{\prime}+\nabla u^{\prime T}\right|^{2}} \tag{2.3}
\end{equation*}
$$

with $k$ as the turbulent kinetic energy or small scales, and $\epsilon$ is the rate of viscous energy dissipation. Reynolds hypothesis is that the turbulence $(R)$ in flows and is a local function of $\nabla u^{\prime}+\nabla u^{\prime T}$. In two dimensional mean flows we have,

$$
\begin{equation*}
R=v_{T}\left(\nabla \bar{u}+\nabla \bar{u}^{T}\right)+\alpha I \quad v_{T}=c_{\mu} \frac{k^{2}}{\epsilon} \tag{2.4}
\end{equation*}
$$

Combining the assumptions , (2.3), (2.4) with (2.2) we arrive at (2.5)

$$
\begin{align*}
\partial_{t} k+\bar{u} \nabla k-\frac{1}{2} c_{\mu} \frac{k^{2}}{\epsilon}\left|\nabla \bar{u}+\nabla \bar{u}^{T}\right|^{2}-\nabla \cdot\left(c_{\mu} \frac{k^{2}}{\epsilon} \nabla k\right)+\epsilon= & 0 \\
\partial_{t} \epsilon+\bar{u} \nabla \epsilon-\frac{1}{2} c_{1} \frac{k^{2}}{\epsilon}\left|\nabla \bar{u}+\nabla \bar{u}^{T}\right|^{2}-\nabla \cdot\left(c_{\epsilon} \frac{k^{2}}{\epsilon} \nabla \epsilon\right)+c_{2} \frac{\epsilon^{2}}{k}= & 0 . \tag{2.5}
\end{align*}
$$

With the constants $c_{\mu}, c_{\epsilon}, c_{1}, c_{2}$ chosen so that the model is accurate for:

- The decay in time of homogeneous turbulence.
- The measurements in shear layers in local equilibrium.
- The log-wall law in boundary layers.

The suggested set of boundary conditions according to [1] is:

- $\bar{u}, k, \epsilon$ given initially everywhere.
- $\bar{u}, k, \epsilon$ given on the inflow boundaries at all $t$
- $v_{t} \partial_{n} \bar{u}, v_{t} \partial_{n} k, v_{t} \partial_{n} \in$ given on the outflow boundaries for all $t$.
- $u \cdot n=0, \frac{\bar{u} \cdot \mathrm{~s}}{\sqrt{v\left|\partial_{n} \bar{u}\right|}}-\frac{1}{\chi} \log \left(\delta \sqrt{\frac{1}{v}\left|\partial_{n} \bar{u}\right|}\right)+\beta=0$ on $\Gamma+\delta$.
- $\left.k\right|_{\Gamma+\delta}=\left|\mu \partial_{n}(\bar{u} \cdot s)\right| c_{\mu}^{-\frac{1}{2}},\left.\epsilon\right|_{\Gamma+\delta}=\frac{1}{\chi \delta}\left|\mu \partial_{n}(\bar{u} \cdot s)\right|^{\frac{3}{2}}$.

Where $\Gamma$ represents a solid wall, and $\delta$ is an adjustable artificial boundary parallel to the wall with $\delta(t, x) \in[10,100] v / u_{\tau}$. s is the tangent to the boundary, and $n$ is the normal to the boundary. In order to study other possible sets of boundary conditions for 2.5 , it is practical to do some simplifications. The
main simplification is to assume that $\frac{k^{2}}{\epsilon}$ is linear, and to not consider changes in time. Thus (2.5) reduces to (2.6)

$$
\begin{align*}
\bar{u} \nabla k-\frac{1}{2} c_{\mu} \frac{k^{2}}{\epsilon}\left|\nabla \bar{u}+\nabla \bar{u}^{T}\right|^{2}-c_{\mu} \frac{k^{2}}{\epsilon} \Delta k+\epsilon= & 0  \tag{2.6}\\
\bar{u} \nabla \epsilon-\frac{1}{2} c_{1} \frac{k^{2}}{\epsilon}\left|\nabla \bar{u}+\nabla \bar{u}^{T}\right|^{2}-c_{\epsilon} \frac{k^{2}}{\epsilon} \Delta \epsilon+c_{2} \frac{\epsilon^{2}}{k}= & 0
\end{align*}
$$

To study a problem similar to (2.6), yields valuable knowledge about possible boundary conditions for (2.5). Thus we study an as simple set of equations possible, the candidate set of equations (2.7) is chosen

$$
\left\{\begin{align*}
\epsilon(x)-\Delta k(x) & =f_{1}(x) & & \text { on } \Omega  \tag{2.7}\\
-\Delta \epsilon(x)+k(x) & =f_{2}(x) & & \text { on } \Omega
\end{align*}\right.
$$

### 2.1 A short description of FEM

The finite element method is a method of numerically approximating a function by describing it in terms of local polynomials. One usually starts out with a problem, lets say Poisson problem (2.8).

$$
\left\{\begin{array}{ccc}
-\Delta u= & f & \text { on } \Omega  \tag{2.8}\\
u= & 0 & \text { on } \partial \Omega
\end{array}\right.
$$

where $\Omega$ is an open domain, and $\partial \Omega$ is its boundary. Then a weak form is created by multiplying by a test function and integrating by parts creating (2.9)

$$
\begin{equation*}
\int_{\Omega} \nabla u \cdot \nabla v \mathrm{~d} x=\int_{\Omega} f v \mathrm{~d} x \tag{2.9}
\end{equation*}
$$

We say; Let $u \in H_{0}^{1}(\Omega), u$ is a weak solution of (2.8) if (2.9) holds for all $v \in H_{0}^{1}(\Omega) . H_{0}^{1}(\Omega)=C_{c}^{0}(\bar{\Omega}) \cap H^{1}(\Omega)$, where $H^{1}(\Omega)$ is a Sobolev space and $C_{c}^{0}(\bar{\Omega})^{1}$ is the space of all continuous functions with compact support.

For easier notation, we let $W=\left\{w \in H_{0}^{1}(\Omega)\right\}$. To discretize $u, v$ we restrict our numerical solution and test functions to a subspace $v_{h}, u_{h} \in W_{h} \subset W$. The space $W_{h}$ is spanned by a set of basis functions $\left\{\phi_{i}\right\}_{i=1}^{N}$. The selection of these basis functions is determined by which element is chosen. The domain is subdivided into a set of elements. These elements is usually triangular or square in 2D, and line segments in 1D. The element type defines local basis functions for each sub domain. Each elements have a set of evaluation points. The basis functions form a basis with respect to the evaluation points. They are usually polynomials. If two basis functions $\phi_{i, L_{1}}, \phi_{j, L_{2}}$ has the same evaluation point, then the sum of them is the global basis function.

[^0]We make an assumption that our numerical solution can be described in terms of the basis functions $\tilde{u}=\sum U_{k} \phi(x)_{k}$. We then have;

$$
\begin{align*}
\int_{\Omega} \sum U_{k} \phi(x)_{k}^{\prime} \cdot \sum V_{i} \phi(x)_{i}^{\prime} \mathrm{d} x & =\int_{\Omega} f \sum V_{i} \phi(x)_{i} \mathrm{~d} x  \tag{2.10}\\
\sum \sum U_{k} V_{i} \int_{\Omega} \phi(x)_{k}^{\prime} \phi(x)_{i}^{\prime} \mathrm{d} x & =\sum V_{i} \int_{\Omega} f \phi(x)_{i} \mathrm{~d} x \tag{2.11}
\end{align*}
$$

Since this has to hold for any test function $v$, we can choose to use a set test functions;

$$
\begin{array}{r}
v_{i}=\sum_{j=1}^{n} V_{i, j} \phi_{j} \\
\left\{\begin{array}{l}
V_{i, i}=1 \\
V_{i, j}=0
\end{array} \quad \text { for } i \neq j\right. \tag{2.13}
\end{array}
$$

Meaning that $\left(V_{i, i}\right)$ is equal to 1 , while all of the other coefficients are 0 . (2.11) then becomes a set of $n$ equations with $\left\{U_{i}\right\}$ as the only unknown values. By setting $A_{i, j}=\int_{\Omega} \phi(x)_{k}^{\prime} \phi(x)_{i}^{\prime} \mathrm{d} x$, and $b_{j}=\sum \int_{\Omega} f \phi(x)_{i} \mathrm{~d} x$, the system (2.11) is the same as the matrix equations (2.14).

$$
\begin{equation*}
A \tilde{u}=b \tag{2.14}
\end{equation*}
$$

where $\tilde{u}=\left[U_{1}, U_{2}, \ldots, U_{n}\right]^{T}$.
By solving (2.14) for $\tilde{\mathcal{U}}$ we obtain our finite element approximation
For this thesis this approach to the finite element method will be used, but in a system of equations.

## Chapter 3

## Condition Numbers

This chapter describes condition numbers and their importance in PDEs. As we know a linear PDE system $L u=f$ can be approximated by using FEM. Such that approximating the solution can be done by instead solving the linear system $A x=b$. In general one would like to solve this by taking the inverse of the matrix $A$. A natural question to ask oneself is: "Is the taking the inverse numerically stable?" To answer this question we introduce condition numbers. We let $e$ represent the error in the vector $b$ coming from the representation of $b$ in the computer. So we will have $A y=b+e$. Then $\left\|A^{-1}(e)\right\| /\left\|A^{-1} b\right\|$ will represent the relative error in the solution, and $\|e\| /\|b\|$ will be the relative error in the data. We then figure out the error in the solution relative the error in the data by dividing one by the other.

$$
\begin{aligned}
& \frac{\left\|A^{-1} e\right\| /\left\|A^{-1} b\right\|}{\|e\| /\|b\|}=\frac{\left\|A^{-1} e\right\|\|b\|}{\|e\|\left\|A^{-1} b\right\|} \\
= & \frac{\left\|A^{-1} e\right\|}{\|e\|} \frac{\|b\|}{\left\|A^{-1} b\right\|} \leq\left\|A^{-1}\right\| \cdot\|A\|
\end{aligned}
$$

on the other hand we have

$$
\begin{aligned}
& \frac{\left\|A^{-1} e\right\| /\left\|A^{-1} b\right\|}{\|e\| /\|b\|}=\frac{\left\|A^{-1} e\right\|\|b\|}{\|e\|\left\|A^{-1} b\right\|} \\
= & \frac{\left\|A^{-1} e\right\|}{\|e\|} \frac{\|b\|}{\left\|A^{-1} b\right\|} \geq \frac{1}{\|A\| \cdot\left\|A^{-1}\right\|}
\end{aligned}
$$

by defining $\kappa(A):=\|A\| \cdot\left\|A^{-1}\right\|$ we get

$$
\begin{aligned}
\frac{1}{\|A\| \cdot\left\|A^{-1}\right\|} & \leq \frac{\left\|A^{-1} e\right\| /\left\|A^{-1} b\right\|}{\|e\| /\|b\|} \leq\|A\| \cdot\left\|A^{-1}\right\| \\
\frac{\|e\|}{\|b\|} \frac{1}{\kappa(A)} & \leq \frac{\left\|A^{-1}(e+b)-A^{-1} b\right\|}{\left\|A^{-1} b\right\|} \leq \frac{\|e\|}{\|b\|} \kappa(A) \\
\frac{\|e\|}{\|b\|} \frac{1}{\kappa(A)} & \leq \frac{\|y-x\|}{\|x\|} \leq \frac{\|e\|}{\|b\|} \kappa(A)
\end{aligned}
$$

Therefore we see that the condition number impose a bound on the error of the solution of the linear problem.

We can furthermore compute the condition number. The norm of a matrix is the following: $\|A\|:=\sup \{\|A x\|:\|x\| \leq 1\}$ which in an $l_{2}$-matrix norm can be re written as $\|A\|=\sup _{x \in C^{n}} \frac{\left|x^{T} A x\right|}{|x|^{2}}$ which is the magnitude of the largest eigenvalue of $A$ called $\mu_{\text {max }}^{A}$. Similarly $\left\|A^{-1}\right\|=\sup _{x \in C^{n}} \frac{\left|x^{T} A^{-1} x\right|}{|x|^{2}}=$ $\frac{1}{\inf _{x \in C^{n}} \frac{\left|x^{T} A x\right|}{|x|^{2}}}$. And $\inf _{x \in C^{n}} \frac{\left|x^{T} A x\right|}{|x|^{2}}$ is the smallest possible eigen value of $A$ called $\mu_{\text {min }}^{A}$. By combining all this we get that $\kappa(A):=\|A\| \cdot\left\|A^{-1}\right\|=\frac{\left|\mu_{\text {max }}^{A}\right|}{\left|\mu_{\min }^{A}\right|} \geq 1$. The smallest possible condition number $A$ can have is 1 .

### 3.1 Condition number relating to the $k-\epsilon$ model

It is useful to examine simpler versions of the $k$ - $\epsilon$-model and solve any issues there before extrapolating the the full model. This thesis will look at the properties of three different weak formulations of the simplified problem in one and two dimensions. One of these formulations is the linear formulation, the two others are systems of dependent equations. The main differences between the two systems are the selection of function spaces for the trail and test functions in FEM. This will be described in detail in chapter 4. The matrices produced by FEM is then preconditioned, and the condition number of the preconditioned matrix is analyzed. The idea is that if the condition number of the preconditioned matrix remain within $O(1)$, then the unconditioned matrix makes an isomorphism between the function space of the solutions $((k, \epsilon))$ and the function space of the input data $\left.\left(f_{1}, f_{2}\right)\right)[6,7]$.

### 3.1.1 An example

Let $(k, \epsilon) \in H^{1}(\Omega) \times H^{1}(\Omega)$ and the finite element method creates the pairing $<A(k, \epsilon),\left(v_{1} \cdot v_{2}\right)>=<\left(f_{1}, f_{2}\right),\left(v_{1} \cdot v_{2}\right)>\in H^{-1}(\Omega) \times H^{-1}(\Omega)$. We use the

Riesz mapping ${ }^{1}$ to construct a preconditioner $B$. The preconditioner should be constructed so that $<B A(k, \epsilon),\left(v_{1} . v_{2}\right)>=<B\left(f_{1}, f_{2}\right),\left(v_{1} . v_{2}\right)>\in H^{1}(\Omega) \times$ $H^{1}(\Omega)$. One may then analyze the condition number $\kappa(B A)$ in order to figure out if the mapping $(k, \epsilon) \rightarrow<A(k, \epsilon),\left(v_{1} . v_{2}\right)>$ is an isomorphism.

If $c_{0} \leq \kappa(B A) \leq c_{1}$ when the mesh is refined, then $\frac{c_{0}}{\left\|(B A)^{-1}\right\|} \leq\|B A\| \leq$ $\frac{c_{1}}{\left\|(B A)^{-1}\right\|}$. Thus the inverse operator of $B A$ is well defined. Therefor the mapping $B A: H^{1}(\Omega) \times H^{1}(\Omega) \rightarrow H^{1}(\Omega) \times H^{1}(\Omega)$ is an isomorphism. Since $B$ is known to be an isomorphism, it also follows that $A$ itself is isomorphic.

[^1]
## Chapter 4

## Defining a weak formulation of the simplified $k-\epsilon$ - model

For this section we assume $\Omega$ is open and bounded, with a $C^{1}$ boundary. As described earlier, a simplification of the $k$ - $\epsilon$ problem can be defined as following.

Find $\epsilon$ and $k$ solving the equation

$$
\begin{array}{rlrl}
\epsilon(x)-\Delta k(x) & =f_{1}(x) & \text { on } \Omega \\
-\Delta \epsilon(x)+k(x) & =f_{2}(x) & & \text { on } \Omega
\end{array}
$$

Two possible sets of boundary conditions are

$$
\begin{align*}
& B C_{1}=\left\{\begin{array}{lll}
\epsilon(x)= & g_{1}(x) & \text { on } \partial \Omega \\
k(x)= & g_{2}(x) & \text { on } \partial \Omega
\end{array}\right.  \tag{4.1}\\
& B C_{2}
\end{align*}=\left\{\begin{array}{cl}
\epsilon(x)= & g_{1}(x)  \tag{4.2}\\
\text { on } \partial \Omega
\end{array}, \begin{array}{lll}
\frac{\partial \epsilon}{\partial n}= & g_{2}(x) & \text { on } \partial \Omega
\end{array}\right.
$$

Writing the problem in matrix form looks like:

$$
\left[\begin{array}{cc}
I & -\Delta  \tag{4.3}\\
-\Delta & I
\end{array}\right] \cdot\left[\begin{array}{l}
\epsilon \\
k
\end{array}\right]=\left[\begin{array}{l}
f_{1} \\
f_{2}
\end{array}\right]
$$

We can then multiply it with a test function, integrate, and integrate by parts to obtain a weak formulation

$$
\begin{aligned}
& {\left[\begin{array}{ll}
v_{1} & v_{2}
\end{array}\right] \cdot\left[\begin{array}{cc}
I & -\Delta \\
-\Delta & I
\end{array}\right] \cdot\left[\begin{array}{l}
\epsilon \\
k
\end{array}\right]=\left[\begin{array}{ll}
v_{1} & v_{2}
\end{array}\right] \cdot\left[\begin{array}{l}
f_{1} \\
f_{2}
\end{array}\right]} \\
& \int_{\Omega} v_{1} \epsilon-v_{1} \Delta k-v_{2} \Delta \epsilon+v_{2} k d x=\int_{\Omega} v_{1} f_{1}+v_{2} f_{2} d x
\end{aligned}
$$

Integration by parts can yield two different weak formulations.

$$
\begin{align*}
\int_{\Omega} v_{1} \epsilon+\nabla k \cdot \nabla v_{1}+\nabla v_{2} \cdot \nabla \epsilon+v_{2} k d x & =\int_{\Omega} v_{1} f_{1}+v_{2} f_{2} d \mathrm{x}  \tag{4.4}\\
\int_{\Omega} v_{1} \epsilon-k \Delta v_{1}-v_{2} \Delta \epsilon+v_{2} k d \mathrm{x} & =\int_{\Omega} v_{1} f_{1}+v_{2} f_{2} d \mathrm{x} \tag{4.5}
\end{align*}
$$

### 4.1 The Trace Theorem

The trace theorem [10, p. 258] states that:
Trace Theorem 4.1 Assume $\Omega$ is bounded and $\partial \Omega$ is $C^{1}$.
Then there exists a bounded linear operator

$$
T: W^{1, p}(\Omega) \rightarrow L^{p}(\partial \Omega)
$$

such that;
$T u=\left.u\right|_{\partial \Omega}$ if $u \in W^{1, p}(\Omega) \cap C(\bar{\Omega})$
and
$\|T u\|_{L^{p}(\partial \Omega)} \leq C\|u\|_{W^{1, p}(\Omega)}$ for each $u \in W^{1, p}(\Omega)$, with the constant $C$ depending only on $p$ and $\Omega$.

## $4.2 \quad H^{2} \times L^{2}$ formulation

A weak formulation reads as follows:
Weak formulation 4.2.1 Find $\epsilon \in H^{2}(\Omega)$ and $k \in L^{2}(\Omega)$ solving (4.5) $\forall\left(v_{1}, v_{2}\right) \in H^{2}(\Omega) \times L^{2}(\Omega)$.

For this formulation, we use the boundary condition (4.2). The advantage of using (4.2) is that it bounded by the trace theorem. For all functions $u \in$ $H^{2}(\Omega)$, there exists a trace Tu with $\left\|\frac{\partial u}{\partial n}\right\|_{L^{2}(\partial \Omega)} \leq C_{1}\|u\|_{H^{2}(\Omega)}$, and $\|u\|_{L^{2}(\partial \Omega)} \leq$ $C_{2}\|u\|_{H^{2}(\Omega)}$. Thus it makes sense to talk about the trace. It is not however obvious that it makes sense to talk about the trace of $L^{2}(\Omega)$, and it is not obvious that it is bounded for every bounded function $w \in L^{2}(\Omega)$.

## $4.3 \quad H^{1} \times H^{1}$ formulation

A weak formulation reads as follows:
Weak formulation 4.3.1 Find $\epsilon \in H^{1}(\Omega)$ and $k \in H^{1}(\Omega)$ solving (4.4) $\forall\left(v_{1}, v_{2}\right) \in H^{1}(\Omega) \times H^{1}(\Omega)$.

For this formulation, we use the boundary condition (4.1). The advantage of using (4.1) is that it bounded by the trace theorem. For all functions $u \in$ $H^{1}(\Omega)$, there exists a trace $T u$ where we have $\|u\|_{L^{2}(\partial \Omega)} \leq C_{2}\|u\|_{H^{1}(\Omega)}$. Thus we know it makes sense to talk about the trace. We can then construct a trace function $T(\epsilon, k)_{H^{1}(\Omega) \times H^{1}(\Omega)} \rightarrow(\epsilon, k)_{L^{2}(\partial \Omega) \times L^{2}(\partial \Omega)}$ which is bounded.

### 4.4 Argument about choice in boundary conditions

The choices of the boundary conditions 4.2 and 4.3 are made to work on as general problems as possible. Allowing $(\epsilon, k) \in H^{1}(\bar{\Omega}) \times H^{1}(\bar{\Omega})$, or $(\epsilon, k) \in$ $H^{2}(\bar{\Omega}) \times L^{2}(\bar{\Omega})$ permits the usage of 12 different boundary conditions, some of them listed here. The full table is listed in the appendix as 11.3. Most of these possible boundary conditions have the disadvantage that they have higher requirements for the function space of the solution. These requirements are usually of the type: Find $(\epsilon, k) \in H^{2}(\bar{\Omega}) \times L^{2}(\Omega)$. They require the closure to be included in the solution space. The Trace Theorem is not enough to prove the existence of the boundary condition for all functions in the solution space. The Trace theorem is used to determine the exact requirements.

The boundary conditions (4.2) and (4.1) exists for all functions in the function spaces $H^{2}(\Omega) \times L^{2}(\Omega)$ and $H^{1}(\Omega) \times H^{1}(\Omega)$ respectively.

List of boundary conditions requiring $(\epsilon, k) \in H^{2}(\bar{\Omega}) \times L^{2}(\Omega)$

$$
\begin{align*}
& \left\{\begin{array}{cc}
\frac{\partial \epsilon(x)}{\partial n}=g_{1}(x) & \text { on } \partial \Omega \\
\frac{\partial^{2} \epsilon(x)}{(\partial n)^{2}}=g_{2}(x) & \text { on } \partial \Omega
\end{array}\right.  \tag{4.6}\\
& \begin{cases}\epsilon(x)=g_{1}(x) & \text { on } \partial \Omega \\
\frac{\partial^{2} \epsilon(x)}{(\partial n)^{2}}=g_{2}(x) & \text { on } \partial \Omega\end{cases} \tag{4.7}
\end{align*}
$$

If we have $\epsilon \in H^{2}(\Omega)$ the Trace theorem says we know that the trace Tu exists and is bounded by $\|\epsilon\|_{H^{2}(\Omega)} . \epsilon \in H^{2}(\Omega)$ means $\|\epsilon\|_{H^{2}(\Omega)}<\infty$. So $\|T u\|_{H^{1}(\partial \Omega)}$ is well defined and bounded. The boundary condition

$$
\begin{equation*}
\frac{\partial^{2} \epsilon(x)}{(\partial n)^{2}}=g_{2}(x) \quad \text { on } \partial \Omega \tag{4.8}
\end{equation*}
$$

is not bounded by $\|\epsilon\|_{H^{2}(\Omega)}$. So the problem has to read as follows to be well defined:

Weak formulation 4.4.1 Find $\epsilon \in H^{2}(\bar{\Omega})$ and $k \in L^{2}(\Omega)$
solving (4.5) $\forall\left(v_{1}, v_{2}\right) \in H^{2}(\bar{\Omega}) \times L^{2}(\Omega)$.

## Chapter 5

## Uniqueness of Solutions of Bilinear forms

This section contains proofs of uniqueness of solutions of the bilinear forms. In 1969 Ivo Babuška submitted to Springer-Verlag the theorem later called "Babuška-Lax-Milgram Theorem" [11, 5]. The "Babuška-Lax-Milgram Theorem" is a generalization of the "Lax-Milgram Theorem" [10]. It will be used to prove uniqueness of the $H^{1}(\Omega) \times H^{1}(\Omega)$ formulation in this thesis.

Babuška-Lax-Milgram Theorem 5.1 Let $H_{1}$ and $H_{2}$ be two Hilbert (complex and complete) spaces with scalar product $(\cdot, \cdot)_{H_{1}}$ and $(\cdot, \cdot)_{H_{2}}$. Let $B[u, v]$ be a bilinear form on $H_{1} \times H_{2}$ for $u \in H_{1}, v \in H_{2}$ such that;

$$
\begin{align*}
&|B[u, v]| \leq C_{1}\|u\|_{H_{1}}\|v\|_{H_{2}}  \tag{5.1}\\
& \sup _{\substack{u \in H_{1} \\
\|u\| \leq 1}}|B[u, v]| \geq C_{2}\|v\|_{H_{2}}  \tag{5.2}\\
& \sup _{v \in H_{2}}|B[u, v]| \geq C_{3}\|u\|_{H_{1}}  \tag{5.3}\\
&\|v\| \leq 1
\end{align*}
$$

with $C_{2}>0, C_{3}>0, C_{1}<\infty$.
Let $f$ be a linear functional on $H_{2}\left(f \in H_{2}^{*}\right)$. Then there exists exactly one element $u_{f} \in H_{1}$ such that

$$
B\left[u_{f}, v\right]=\overline{f(v)}^{1} \quad \forall v \in H_{2}
$$

and

$$
\left\|u_{f}\right\|_{H_{1}} \leq \frac{\|f\|_{H_{2}^{*}}}{C_{3}}
$$

[^2]
### 5.1 System $H^{1} \times H^{1}$

### 5.1.1 Preliminaries

Let $\epsilon, k \in H^{1}(\Omega)$

$$
\left\{\begin{align*}
\epsilon(x)-\Delta k(x) & =f_{1}(x) & & \text { on } \Omega  \tag{5.4}\\
-\Delta \epsilon(x)+k(x) & =f_{2}(x) & & \text { on } \Omega \\
\epsilon & =g_{1} & & \text { on } \partial \Omega \\
k & =g_{2} & & \text { on } \partial \Omega .
\end{align*}\right.
$$

To examine (5.4) we look at a similar system (5.5)

$$
\left\{\begin{align*}
u_{1}(x)-\Delta u_{1}(x) & =f_{1}(x) & & \text { on } \Omega  \tag{5.5}\\
-\Delta u_{1}(x)+u_{2}(x) & =f_{2}(x) & & \text { on } \Omega \\
u_{1} & =0 & & \text { on } \partial \Omega \\
u_{2} & =0 & & \text { on } \partial \Omega
\end{align*}\right.
$$

We multiply it with a test function $\left(v_{1}, v_{2}\right) \in H^{1}(\Omega) \times H^{1}(\Omega)$ and integrate by parts to get

$$
\begin{align*}
& {\left[\begin{array}{ll}
v_{1} & v_{2}
\end{array}\right] \cdot\left[\begin{array}{cc}
I & -\Delta \\
-\Delta & I
\end{array}\right] \cdot\left[\begin{array}{l}
u_{1} \\
u_{2}
\end{array}\right]=\left[\begin{array}{l}
v_{1} \\
v_{2}
\end{array}\right] \cdot\left[\begin{array}{l}
f_{1} \\
f_{2}
\end{array}\right] } \\
& \int_{\Omega} v_{1} u_{1}-v_{1} \Delta u_{2}-v_{2} \Delta u_{1}+v_{2} u_{1} \mathrm{~d} x=\int_{\Omega} v_{1} f_{1}+v_{2} f_{2} \mathrm{~d} x  \tag{5.6}\\
& \int_{\Omega} v_{1} \epsilon+\nabla v_{1} \cdot \nabla k+\nabla v_{2} \cdot \nabla \epsilon+v_{2} k \mathrm{~d} x=\int_{\Omega} v_{1} f_{1}+v_{2} f_{2} \mathrm{~d} x . \tag{5.7}
\end{align*}
$$

If we show that (5.7) holds for $\left(u_{1}, u_{2}\right) \in H_{0}^{1}(\Omega) \times H_{0}^{1}(\Omega)$, there must exists a pair $\left(w_{1}, w_{2}\right) \in H_{g_{1}}^{1}(\Omega) \times H_{g_{2}}^{1}(\Omega)$ such that $\left(\epsilon-w_{1}, k-w_{2}\right)=\left(u_{1}, u_{2}\right) \in$ $H_{0}^{1}(\Omega) \times H_{0}^{1}(\Omega)$ and

$$
\begin{array}{cc}
\tilde{f}_{1}=f_{1}-w_{1}+\Delta w_{2} & \in H^{-1}(\Omega) \\
\tilde{f}_{2}=f_{2}+\Delta w_{2}-w_{1} & \in H^{-1}(\Omega)
\end{array}
$$

Thus with these substitutions (5.4) turns into this:

$$
\begin{aligned}
u_{1}(x)-\Delta u_{2}(x) & =\tilde{f}_{1}(x) \quad \text { on } \Omega \\
-\Delta u_{1}(x)+u_{2}(x) & =\tilde{f}_{2}(x) \quad \text { on } \Omega \\
u_{1} & =0 \quad \text { on } \partial \Omega \\
u_{2} & =0 \quad \text { on } \partial \Omega
\end{aligned}
$$

### 5.1.2 Using Babuška-Lax-Milgram

The proof in this section was created by the student.
If we assume that $\left(u_{1}, u_{2}\right) \in H_{0}^{1}(\Omega) \times H_{0}^{1}(\Omega)$

$$
\begin{aligned}
|B[u, v]| & =\sqrt{\left|\int_{\Omega} u_{1} v_{1}+\nabla u_{2} \cdot \nabla v_{1} \mathrm{~d} x\right|^{2}+\left|\int_{\Omega} u_{2} v_{2}+\nabla u_{1} \cdot \nabla v_{2} \mathrm{~d} x\right|^{2}} \\
& =\sqrt{\left(\left|\int_{\Omega} u_{1} v_{1}+\nabla u_{2} \cdot \nabla v_{1} \mathrm{~d} x\right|\right)^{2}+\left(\left|\int_{\Omega} u_{2} v_{2}+\nabla u_{1} \cdot \nabla v_{2} \mathrm{~d} x\right|\right)^{2}} \\
& \leq \sqrt{\left(\int_{\Omega}\left|u_{1} v_{1}+\nabla u_{2} \cdot \nabla v_{1}\right| \mathrm{d} x\right)^{2}+\left(\int_{\Omega}\left|u_{2} v_{2}+\nabla u_{1} \cdot \nabla v_{2}\right| \mathrm{d} x\right)^{2}} \\
& =\sqrt{\left(\left\|u_{1} v_{1}+\nabla u_{2} \cdot \nabla v_{1}\right\|_{L^{1}(\Omega)}\right)^{2}+\left(\left\|u_{2} v_{2}+\nabla u_{1} \cdot \nabla v_{2}\right\|_{L^{1}(\Omega)}\right)^{2}}
\end{aligned}
$$

applying Minkowskis inequality

$$
\begin{aligned}
& \leq \sqrt{\left(\left\|u_{1} v_{1}\right\|_{L^{1}(\Omega)}+\left\|\nabla u_{2} \cdot \nabla v_{1}\right\|_{L^{1}(\Omega)}\right)^{2}+\left(\left\|u_{2} v_{2}\right\|_{L^{1}(\Omega)}+\left\|\nabla u_{1} \cdot \nabla v_{2}\right\|_{L^{1}(\Omega)}\right)^{2}} \\
& \leq\left(\left(\left\|u_{1}\right\|_{L^{2}(\Omega)}\left\|v_{1}\right\|_{L^{2}(\Omega)}+\left\|\nabla u_{2}\right\|_{L^{2}(\Omega)}\left\|\nabla v_{1}\right\|_{L^{2}(\Omega)}\right)^{2}\right. \\
& \left.+\left(\left\|u_{2}\right\|_{L^{2}(\Omega)}\left\|v_{2}\right\|_{L^{2}(\Omega)}+\left\|\nabla u_{1}\right\|_{L^{2}(\Omega)}\left\|\nabla v_{2}\right\|_{L^{2}(\Omega)}\right)^{2}\right)^{\frac{1}{2}} \\
& =\left(\left\|u_{1}\right\|_{L^{2}(\Omega)}^{2}\left\|v_{1}\right\|_{L^{2}(\Omega)}^{2}+\left\|\nabla u_{2}\right\|_{L^{2}(\Omega)}^{2}\left\|\nabla v_{1}\right\|_{L^{2}(\Omega)}^{2}\right. \\
& +2\left\|u_{1}\right\|_{L^{2}(\Omega)}\left\|v_{1}\right\|_{L^{2}(\Omega)}\left\|\nabla u_{2}\right\|_{L^{2}(\Omega)}\left\|\nabla v_{1}\right\|_{L^{2}(\Omega)} \\
& +\left\|u_{2}\right\|_{L^{2}(\Omega)}^{2}\left\|v_{2}\right\|_{L^{2}(\Omega)}^{2}+\left\|\nabla u_{1}\right\|_{L^{2}(\Omega)}^{2}\left\|\nabla v_{2}\right\|_{L^{2}(\Omega)}^{2} \\
& \left.\quad+2\left\|u_{2}\right\|_{L^{2}(\Omega)}\left\|v_{2}\right\|_{L^{2}(\Omega)}\left\|\nabla u_{1}\right\|_{L^{2}(\Omega)}\left\|\nabla v_{2}\right\|_{L^{2}(\Omega)}\right)^{\frac{1}{2}} \\
& \leq\left(\left\|u_{1}\right\|_{H^{1}(\Omega)}^{2}\left\|v_{1}\right\|_{H^{1}(\Omega)}^{2}+\left\|u_{2}\right\|_{H^{1}(\Omega)}^{2}\left\|v_{1}\right\|_{H^{1}(\Omega)}^{2}\right. \\
& +2\left\|u_{1}\right\|_{H^{1}(\Omega)}\left\|v_{1}\right\|_{H^{1}(\Omega)}\left\|u_{2}\right\|_{H^{1}(\Omega)}\left\|v_{1}\right\|_{H^{1}(\Omega)} \\
& +\left\|u_{2}\right\|_{H^{1}(\Omega)}^{2}\left\|v_{2}\right\|_{H^{1}(\Omega)}^{2}+\left\|u_{1}\right\|_{H^{1}(\Omega)}^{2}\left\|v_{2}\right\|_{H^{1}(\Omega)}^{2} \\
& \left.\quad+2\left\|u_{2}\right\|_{H^{1}(\Omega)}\left\|v_{2}\right\|_{H^{1}(\Omega)}\left\|u_{1}\right\|_{H^{1}(\Omega)}\left\|v_{2}\right\|_{H^{1}(\Omega)}\right)^{\frac{1}{2}}
\end{aligned}
$$

$$
\begin{gathered}
=\left(\left(\left\|u_{1}\right\|_{H^{1}(\Omega)}^{2}\left\|v_{1}\right\|_{H^{1}(\Omega)}^{2}+\left\|u_{2}\right\|_{H^{1}(\Omega)}^{2}\left\|v_{1}\right\|_{H^{1}(\Omega)}^{2}\right.\right. \\
\left.+\left\|u_{2}\right\|_{H^{1}(\Omega)}^{2}\left\|v_{2}\right\|_{H^{1}(\Omega)}^{2}+\left\|u_{1}\right\|_{H^{1}(\Omega)}^{2}\left\|v_{2}\right\|_{H^{1}(\Omega)}^{2}\right) \\
+2\left\|u_{1}\right\|_{H^{1}(\Omega)}\left\|v_{1}\right\|_{H^{1}(\Omega)}\left\|u_{2}\right\|_{H^{1}(\Omega)}\left\|v_{1}\right\|_{H^{1}(\Omega)} \\
\left.+2\left\|u_{2}\right\|_{H^{1}(\Omega)}\left\|v_{2}\right\|_{H^{1}(\Omega)}\left\|u_{1}\right\|_{H^{1}(\Omega)}\left\|v_{2}\right\|_{H^{1}(\Omega)}\right)^{\frac{1}{2}} \\
=\left(\left(\left\|u_{1}\right\|_{H^{1}(\Omega)}^{2}+\left\|u_{2}\right\|_{H^{1}(\Omega)}^{2}\right)\left(\left\|v_{1}\right\|_{H^{1}(\Omega)}^{2}+\left\|v_{2}\right\|_{H^{1}(\Omega)}^{2}\right)\right. \\
\left.+2\left\|u_{1}\right\|_{H^{1}(\Omega)}\left\|u_{2}\right\|_{H^{1}(\Omega)}\left(\left\|v_{1}\right\|_{H^{1}(\Omega)}^{2}+\left\|v_{2}\right\|_{H^{1}(\Omega)}^{2}\right)\right)^{\frac{1}{2}} \\
=\left(\|u\|_{H^{1}(\Omega) \times H^{1}(\Omega)}^{2}\|v\|_{H^{1}(\Omega) \times H^{1}(\Omega)}^{2}\right. \\
\left.+2\left\|u_{1}\right\|_{H^{1}(\Omega)}\left\|u_{2}\right\|_{H^{1}(\Omega)}\|v\|_{H^{1}(\Omega) \times H^{1}(\Omega)}^{2}\right)^{\frac{1}{2}} \\
=\|v\|_{H^{1}(\Omega) \times H^{1}(\Omega)}\left(\|u\|_{H^{1}(\Omega) \times H^{1}(\Omega)}^{2}+2\left\|u_{1}\right\|_{H^{1}(\Omega)}\left\|u_{2}\right\|_{H^{1}(\Omega)}\right)^{\frac{1}{2}}
\end{gathered}
$$

by Youngs Inequality

$$
\begin{aligned}
& \leq\|v\|_{H^{1}(\Omega) \times H^{1}(\Omega)}\left(\|u\|_{H^{1}(\Omega) \times H^{1}(\Omega)}^{2}+\left\|u_{1}\right\|_{H^{1}(\Omega)}^{2}+\left\|u_{2}\right\|_{H^{1}(\Omega)}^{2}\right)^{\frac{1}{2}} \\
& =\|v\|_{H^{1}(\Omega) \times H^{1}(\Omega)}\left(\|u\|_{H^{1}(\Omega) \times H^{1}(\Omega)}^{2}+\|u\|_{H^{1}(\Omega) \times H^{1}(\Omega)}^{2}\right)^{\frac{1}{2}} \\
& =\sqrt{2}\|v\|_{H^{1}(\Omega) \times H^{1}(\Omega)}\|u\|_{H^{1}(\Omega) \times H^{1}(\Omega)} .
\end{aligned}
$$

Thus the requirement (5.1) is met by:

$$
|B[u, v]| \leq \sqrt{2}\|v\|_{H^{1}(\Omega) \times H^{1}(\Omega)}\|u\|_{H^{1}(\Omega) \times H^{1}(\Omega)}
$$

Next step is to prove the coercivity requirements (5.2) and (5.3) .

$$
\begin{aligned}
B[u, w] & =\int_{\Omega} u_{1} w_{1}+\nabla u_{2} \cdot \nabla w_{1} \mathrm{~d} x+\int_{\Omega} u_{2} w_{2}+\nabla u_{1} \cdot \nabla w_{2} \mathrm{~d} x \\
& =\left(u_{1}, w_{1}\right)_{L^{2}(\Omega)}+\left(u_{2}, w_{1}\right)_{H_{0}^{1}(\Omega)}+\left(u_{2}, w_{2}\right)_{L^{2}(\Omega)}+\left(u_{1}, w_{2}\right)_{H_{0}^{1}(\Omega)} .
\end{aligned}
$$

We prove the coercivity requirements by finding a function that fulfills it. This can be done because the coercivity requirements is a sup requirement. By careful construction of a candidate function the sup requirement is met. This is
done with

$$
\begin{align*}
w=\left[\begin{array}{l}
w_{1} \\
w_{2}
\end{array}\right] & =K\left[\begin{array}{l}
u_{2}-u_{1}+u_{1} \\
u_{1}+u_{2}-u_{2}
\end{array}\right] \quad\left(=K\left[\begin{array}{l}
u_{2} \\
u_{1}
\end{array}\right]\right)  \tag{5.8}\\
K & =\frac{1}{\sqrt{\left\|u_{1}\right\|_{H^{1}(\Omega)}^{2}+\left\|u_{2}\right\|_{H^{1}(\Omega)}^{2}}} \tag{5.9}
\end{align*}
$$

Thus $\|w\|_{H^{1}(\Omega) \times H^{1}(\Omega)}=1$, so the requirements is met.

$$
\begin{aligned}
& =K\left(u_{1}, u_{2}-u_{1}+u_{1}\right)_{L^{2}(\Omega)}+K\left(u_{2}, u_{2}-u_{1}+u_{1}\right)_{H_{0}^{1}(\Omega)} \\
& +K\left(u_{2}, u_{1}+u_{2}-u_{2}\right)_{L^{2}(\Omega)}+K\left(u_{1}, u_{1}+u_{2}-u_{2}\right)_{H_{0}^{1}(\Omega)} \\
& =K\left(u_{1}, u_{2}\right)_{L^{2}(\Omega)}-K\left(u_{1}, u_{1}\right)_{L^{2}(\Omega)}+K\left(u_{1}, u_{1}\right)_{L^{2}(\Omega)} \\
& +K\left(u_{2}, u_{2}\right)_{H_{0}^{1}(\Omega)}-K\left(u_{2}, u_{1}\right)_{H_{0}^{1}(\Omega)}+K\left(u_{2}, u_{1}\right)_{H_{0}^{1}(\Omega)} \\
& +K\left(u_{2}, u_{1}\right)_{L^{2}(\Omega)}-K\left(u_{2}, u_{2}\right)_{L^{2}(\Omega)}+K\left(u_{2}, u_{2}\right)_{L^{2}(\Omega)} \\
& +K\left(u_{1}, u_{1}\right)_{H_{0}^{1}(\Omega)}+K\left(u_{1}, u_{2}\right)_{H_{0}^{1}(\Omega)}-K\left(u_{1}, u_{2}\right)_{H_{0}^{1}(\Omega)} \\
& =K\left\|u_{1}\right\|_{H_{0}^{1}(\Omega)}^{2}+K\left\|u_{2}\right\|_{H_{0}^{1}(\Omega)}^{2}+K\left\|u_{1}\right\|_{\left.L_{( }^{2} \Omega\right)}^{2}+K\left\|u_{2}\right\|_{\left.L_{( }^{2} \Omega\right)}^{2} \\
& -K\left\|u_{2}\right\|_{L^{2}(\Omega)}^{2}-K\left\|u_{1}\right\|_{L^{2}(\Omega)}^{2}+2 K\left(u_{2}, u_{1}\right)_{L^{2}(\Omega)} \\
& =K\left\|u_{1}\right\|_{H_{0}^{1}(\Omega)}^{2}+K\left\|u_{2}\right\|_{H_{0}^{1}(\Omega)}^{2}+K\left\|u_{1}\right\|_{\left.L_{( }^{2} \Omega\right)}^{2}+K\left\|u_{2}\right\|_{\left.L_{( }^{2} \Omega\right)}^{2}-K\left\|u_{1}-u_{2}\right\|_{L^{2}(\Omega)}^{2} \\
& =K\left\|u_{1}\right\|_{H_{0}^{1}(\Omega)}^{2}+K\left\|u_{2}\right\|_{H_{0}^{1}(\Omega)}^{2}+K\left\|u_{1}\right\|_{\left.L_{( }^{2} \Omega\right)}^{2}+K\left\|-u_{2}\right\|_{\left.L_{( }^{2} \Omega\right)}^{2}-K\left\|u_{1}-u_{2}\right\|_{L^{2}(\Omega)}^{2}
\end{aligned}
$$

## minowskis inequality

$$
\begin{aligned}
& \geq K\left\|u_{1}\right\|_{H_{0}^{1}(\Omega)}^{2}+K\left\|u_{2}\right\|_{H_{0}^{1}(\Omega)}^{2}+K\left\|u_{1}-u_{2}\right\|_{\left.L_{( }^{2} \Omega\right)}^{2}-K\left\|u_{1}-u_{2}\right\|_{L^{2}(\Omega)}^{2} \\
& =K\left\|u_{1}\right\|_{H_{0}^{1}(\Omega)}^{2}+K\left\|u_{2}\right\|_{H_{0}^{1}(\Omega)}^{2} \\
& =\frac{K}{2}\left\|u_{1}\right\|_{H_{0}^{1}(\Omega)}^{2}+\frac{K}{2}\left\|u_{2}\right\|_{H_{0}^{1}(\Omega)}^{2}+\frac{K}{2}\left\|u_{1}\right\|_{H_{0}^{1}(\Omega)}^{2}+\frac{K}{2}\left\|u_{2}\right\|_{H_{0}^{1}(\Omega)}^{2}
\end{aligned}
$$

Poincares inequality
$\geq \frac{K}{2}\left\|u_{1}\right\|_{H_{0}^{1}(\Omega)}^{2}+\frac{K}{2}\left\|u_{2}\right\|_{H_{0}^{1}(\Omega)}^{2}+\frac{C_{1}}{2}\left\|u_{1}\right\|_{L^{2}(\Omega)}^{2}+\frac{C_{2}}{2}\left\|u_{2}\right\|_{L^{2}(\Omega)}^{2}$
$\geq D\left(\left\|u_{1}\right\|_{H^{1}(\Omega)}^{2}+\left\|u_{2}\right\|_{H^{1}(\Omega)}^{2}\right)$
$=D\|u\|_{H^{1}(\Omega) \times H^{1}(\Omega)}^{2}$

So the requirement (5.3). To show (5.2), note that $B[\cdot, \cdot]$ is symmetric so:

$$
\begin{aligned}
|B[u, w]| & =\left|B\left[\left[\begin{array}{l}
u_{1} \\
u_{2}
\end{array}\right], K\left[\begin{array}{l}
u_{2} \\
u_{1}
\end{array}\right]\right]\right| \\
& =\left|K B\left[\left[\begin{array}{l}
u_{1} \\
u_{2}
\end{array}\right],\left[\begin{array}{l}
u_{2} \\
u_{1}
\end{array}\right]\right]\right| \\
& =\left|K B\left[\left[\begin{array}{l}
u_{2} \\
u_{1}
\end{array}\right],\left[\begin{array}{l}
u_{1} \\
u_{2}
\end{array}\right]\right]\right| \\
& =\left|B\left[\left[\begin{array}{l}
u_{2} \\
u_{1}
\end{array}\right], K\left[\begin{array}{l}
u_{1} \\
u_{2}
\end{array}\right]\right]\right| \\
& =|B[w, u]|
\end{aligned}
$$

Which fulfills the requirement (5.2).

### 5.2 System $H^{2} \times L^{2}$

The argument in this section was created by the student. For simplicity we write $H^{-2}$ for the dual space of $H^{2}$, and $H^{-2} \times L^{2}$ will be identified as the dual space of $H^{2} \times L^{2}$, we also assume that $\Omega=(0,1)^{n}$

### 5.2.1 Preliminaries

Before looking at the problem, we can examine a candidate for preconditioning for the system. The preconditioner will be the matrix arising from the problem (5.10)

$$
\left\{\begin{align*}
\Delta \Delta u_{1}(x)+u_{1} & =f_{1}(x) & & \text { on } \Omega  \tag{5.10}\\
u_{2} & =f_{2}(x) & & \text { on } \Omega \\
u_{1} & =0 & & \text { on } \partial \Omega \\
\frac{\partial u_{1}}{\partial n} & =0 & & \text { on } \partial \Omega
\end{align*}\right.
$$

The bilinear from associated with (5.10) is (5.11).

$$
\begin{equation*}
P[u, v]=\int_{\Omega} \Delta u_{1} \Delta v_{1}+u_{1} v_{1}+u_{2} v_{2} \mathrm{~d} x \tag{5.11}
\end{equation*}
$$

$P[u, v]$ is a positive definite and symmetric, so its an inner product on $H^{2}(\Omega) \times$ $L^{2}(\Omega)$. The system (5.10) has a unique solution for every pair $f_{1}, f_{2}$ in the dual space of $H^{2}(\Omega) \times L^{2}(\Omega)$ by Riesz representation theorem. It is obvious that $P \in \mathcal{L}\left(H^{2} \times L^{2}, H^{-2} \times L^{2}\right)$ (the space of linear functionals from $H^{2} \times L^{2}$ to $H^{-2} \times L^{2}$ ). Since $P^{-1}$ exists, it is a member of $\mathcal{L}\left(H^{-2} \times L^{2}, H^{2} \times L^{2}\right)$. $P$ has a unique solution, we know that $P$ has an inverse so that that $P^{-1} \circ P=I_{H^{2} \times L^{2}}$ (the identity functional on $H^{2} \times L^{2}$ ).

### 5.2.2 The bilinear form

A similar argument as the one made in 5.1.1 can be made to show that we only need to prove uniqueness for (5.12) with $\left\{u_{1}, v_{1}\right\} \in H_{0}^{2}(\Omega),\left\{u_{2}, v_{2}\right\} \in L^{2}(\Omega)$

The bilinear form of the system looks like:

$$
\begin{equation*}
B[u, v]=\int_{\Omega}-\Delta v_{1} u_{2}-\Delta u_{1} v_{2}+u_{1} v_{1}+u_{2} v_{2} \mathrm{~d} x . \tag{5.12}
\end{equation*}
$$

With the matrix

$$
B=\left[\begin{array}{cc}
I & -\Delta  \tag{5.13}\\
-\Delta & I
\end{array}\right]
$$

If we let the operator matrix $B$ work on a pair of functions $u=\left(u_{1}, u_{2}\right) \in$ $H^{2} \times L^{2}$, then

$$
\begin{aligned}
B \cdot u & =\left[\begin{array}{cc}
I & -\Delta \\
-\Delta & I
\end{array}\right] \cdot\left[\begin{array}{l}
u_{1} \\
u_{2}
\end{array}\right] \\
& =\left[\begin{array}{c}
u_{1}-\Delta u_{2} \\
-\Delta u_{1}+u_{2}
\end{array}\right] \quad \in\left[\begin{array}{c}
H^{-2} \\
L^{2}
\end{array}\right]
\end{aligned}
$$

Thus $B \in \mathcal{L}\left(H^{2} \times L^{2}, H^{-2} \times L^{2}\right)$ as well. We conclude that $P^{-1} \circ B \in \mathcal{L}\left(H^{2} \times\right.$ $\left.L^{2}, H^{2} \times L^{2}\right)$.

We know $\left(P^{-1} \circ P\right)$ is bounded and 0 is not part of its spectrum If the same holds true for $\left(P^{-1} \circ B\right)$, we can use Lax-Milgram to prove uniqueness of solutions for $\left(P^{-1} \circ B\right)[u, v]$.

First we need to show that the maximal and the minimal eigenvalue of $P^{-1} B$ is limited by constants $c_{1}, c_{2}$.

$$
\left\{\begin{array}{c}
P=\left[\begin{array}{cc}
\Delta \Delta+I & 0 \\
0 & I
\end{array}\right]  \tag{5.14}\\
B=\left[\begin{array}{cc}
I & -\Delta \\
-\Delta & I
\end{array}\right]
\end{array}\right.
$$

The corresponding eigenvalue problem of (5.14) is (5.15)

$$
\left[\begin{array}{cc}
I & -\Delta  \tag{5.15}\\
-\Delta & I
\end{array}\right]\left[\begin{array}{l}
u \\
v
\end{array}\right]=\lambda\left[\begin{array}{cc}
\Delta \Delta+I & 0 \\
0 & I
\end{array}\right]\left[\begin{array}{l}
u \\
v
\end{array}\right]
$$

If we are working are working on $C_{c}^{4}\left([0,1]^{n}\right)$ the operator $-\Delta$ has eigenvalues $\lambda_{D}=\left\{k^{2} \pi^{2}\right\}_{k=1}^{N} \in\left(\pi^{2}, \infty\right)$ [12, p.67]. When applying FEM to 5.15 , it turns into a matrix problem; 5.16.

$$
\left[\begin{array}{cc}
I & C  \tag{5.16}\\
C & I
\end{array}\right]\left[\begin{array}{l}
u \\
v
\end{array}\right]=\lambda\left[\begin{array}{cc}
D & 0 \\
0 & I
\end{array}\right]\left[\begin{array}{l}
u \\
v
\end{array}\right]
$$

$C$ is rectangular because $u, v \in H^{2} \times L^{2}$, so instead of squaring, we use $D$.
Since the function space is $C_{c}^{4}\left([0,1]^{n}\right)$ the eigenvalues take the from $\lambda_{k}=$ $(k \pi L(\Omega))^{2}$ [12], and have single multiplicity. Let $e_{i}$ be the eigenvector associated with $\lambda_{i}$ as an eigenvalue for $-\Delta$. Then we have that $u_{i}=\left[a e_{i}, b e_{i}\right]$ is an eigenvalue for (5.15).

$$
\begin{aligned}
& {\left[\begin{array}{cc}
I & -\Delta \\
-\Delta & I
\end{array}\right]\left[\begin{array}{l}
a e_{i} \\
b e_{i}
\end{array}\right] }=\rho\left[\begin{array}{cc}
\Delta \Delta+I & 0 \\
0 & I
\end{array}\right]\left[\begin{array}{l}
a e_{i} \\
b e_{i}
\end{array}\right] \\
& {\left[\begin{array}{cc}
I & \lambda_{i} I \\
\lambda_{i} I & I
\end{array}\right]\left[\begin{array}{l}
a e_{i} \\
b e_{i}
\end{array}\right]=\rho\left[\begin{array}{cc}
\lambda_{i}^{2} I+I & 0 \\
0 & I
\end{array}\right]\left[\begin{array}{l}
a e_{i} \\
b e_{i}
\end{array}\right] . }
\end{aligned}
$$

Further reductions is possible:

$$
\begin{aligned}
{\left[\begin{array}{cc}
I & \lambda_{i} I \\
\lambda_{i} I & I
\end{array}\right]\left[\begin{array}{l}
a e_{i} \\
b e_{i}
\end{array}\right] } & =\rho\left[\begin{array}{cc}
\lambda_{i}^{2} I+I & 0 \\
0 & I
\end{array}\right]\left[\begin{array}{l}
a e_{i} \\
b e_{i}
\end{array}\right] \\
\left(\left[\begin{array}{cc}
1 & \lambda_{i} \\
\lambda_{i} & 1
\end{array}\right] \otimes I\right)\left(\left[\begin{array}{l}
a \\
b
\end{array}\right] \otimes e_{i}\right) & =\rho\left(\left[\begin{array}{cc}
\lambda_{i}^{2}+1 & 0 \\
0 & 1
\end{array}\right] \otimes I\right)\left(\left[\begin{array}{l}
a \\
b
\end{array}\right] \otimes e_{i}\right) \\
\left(\left[\begin{array}{cc}
1 & \lambda_{i} \\
\lambda_{i} & 1
\end{array}\right]\left[\begin{array}{l}
a \\
b
\end{array}\right]\right) \otimes\left(I e_{i}\right) & =\rho\left(\left[\begin{array}{cc}
\lambda_{i}^{2}+1 & 0 \\
0 & 1
\end{array}\right]\left[\begin{array}{l}
a \\
b
\end{array}\right]\right) \otimes\left(I e_{i}\right) .
\end{aligned}
$$

Where $\otimes$ is the Kronecker product. Then the eigenvalues are the eigenvalues of (5.17).

$$
\left(\left[\begin{array}{cc}
1 & \lambda_{i}  \tag{5.17}\\
\lambda_{i} & 1
\end{array}\right]\left[\begin{array}{l}
a \\
b
\end{array}\right]\right)=\rho\left(\left[\begin{array}{cc}
\lambda_{i}^{2}+1 & 0 \\
0 & 1
\end{array}\right]\left[\begin{array}{l}
a \\
b
\end{array}\right]\right) .
$$

The eigenvalues of (5.17) takes the form

$$
\begin{align*}
& \rho_{i}=\frac{\lambda_{i}^{2}+2+\left|\lambda_{i}\right| \sqrt{5 \lambda_{i}^{2}+4}}{2 \lambda_{i}^{2}+2}  \tag{5.18}\\
& \rho_{i}^{\prime}=\frac{\lambda_{i}^{2}+2-\left|\lambda_{i}\right| \sqrt{5 \lambda_{i}^{2}+4}}{2 \lambda_{i}^{2}+2} \tag{5.19}
\end{align*}
$$

Where $\left\{\rho_{i}\right\},\left\{\rho^{\prime}\right\}$ is two sets of eigenvalues. Zero is not an eigenvalue;

$$
\begin{aligned}
& \rho_{i}= \frac{\lambda_{i}^{2}+2+\left|\lambda_{i}\right| \sqrt{5 \lambda_{i}^{2}+4}}{2 \lambda_{i}^{2}+2} \\
& 0= \frac{\lambda_{i}^{2}+2+\left|\lambda_{i}\right| \sqrt{5 \lambda_{i}^{2}+4}}{2 \lambda_{i}^{2}+2} \\
& \text { no solution for } \lambda \geq 1 \\
& \rho_{i}^{\prime}= \frac{\lambda_{i}^{2}+2-\lambda_{i} \sqrt{5 \lambda_{i}^{2}+4}}{2 \lambda_{i}^{2}+2} \\
& 0= \frac{\lambda_{i}^{2}+2+\lambda_{i} \sqrt{5 \lambda_{i}^{2}+4}}{2 \lambda_{i}^{2}+2} \\
&\left\{\begin{array}{c}
\lambda_{i, 1}=1 \\
\lambda_{i, 2}=-1
\end{array}\right.
\end{aligned}
$$

Zero is not an eigenvalue of the system since $\lambda \in\left(\pi^{2}, \infty\right)$. We see that, when discretizing of the operator $-\Delta$, the set $(\pi k)^{2^{N}}$ will be all the eigenvalues of the $N \times N$ matrix representing the discretization. When the system (5.15) is discretized, it is easy to confirm that there will be $2 N$ eigenvalues in the set $\left\{\rho\left(\lambda_{i}\right)\right\} \cup\left\{\rho^{\prime}\left(\lambda_{i}\right)\right\}$. Since there is exactly as many eigenvalues as there is rows, we conclude that all of the eigenvalues are found. The eigenvalues has single multiplicity. All the eigenfunctions are linearly independent and span $\mathbb{R}^{2 N}$ (see 11.2). Therefore we conclude that we have found all the eigenvectors (discretized eigenfunctions). We find all of the eigenfunctions first by solving the systems (5.20) and (5.21).

$$
\begin{align*}
& {\left[\begin{array}{cc}
1 & \lambda_{i} \\
\lambda_{i} & 1
\end{array}\right]\left[\begin{array}{l}
a_{i} \\
b_{i}
\end{array}\right]=\left[\begin{array}{cc}
\rho_{i} \lambda_{i}^{2}+\rho_{i} & 0 \\
0 & \rho_{i}
\end{array}\right]\left[\begin{array}{c}
a_{i} \\
b_{i}
\end{array}\right]}  \tag{5.20}\\
& {\left[\begin{array}{cc}
1 & \lambda_{i} \\
\lambda_{i} & 1
\end{array}\right]\left[\begin{array}{l}
c_{i} \\
d_{i}
\end{array}\right]=\left[\begin{array}{cc}
\rho_{i}^{\prime} \lambda_{i}^{2}+\rho_{i}^{\prime} & 0 \\
0 & \rho_{i}^{\prime}
\end{array}\right]\left[\begin{array}{c}
c_{i} \\
d_{i}
\end{array}\right]} \tag{5.21}
\end{align*}
$$

The eigenfunctions will be with their respective eigenvalues will be

$$
\begin{gather*}
{\left[\begin{array}{c}
a_{i} e_{i} \\
b_{i} e_{i}
\end{array}\right], \rho_{i}}  \tag{5.22}\\
{\left[\begin{array}{c}
c_{i} e_{i} \\
d_{i} e_{i}
\end{array}\right], \rho_{i}^{\prime} .} \tag{5.23}
\end{gather*}
$$

By letting $\lambda$ tend to infinity, the eigenvalues converge towards

$$
\begin{align*}
& \lim _{i \rightarrow \infty} \rho_{i}=\frac{1+\sqrt{5}}{2}  \tag{5.24}\\
& \lim _{i \rightarrow \infty} \rho_{i}^{\prime}=\frac{1-\sqrt{5}}{2} \tag{5.25}
\end{align*}
$$

These are the golden ratio $(\phi)$, and the negative golden ratio conjugate. Since both $\left|\rho_{i}^{\prime}\right|,\left|\rho_{i}\right|$ are strictly increasing when $i$ increases, the condition number of $P^{-1} B$ is be limited by,

$$
\begin{align*}
\lim \kappa\left(P^{-1} B\right) & \geq\left|\frac{\frac{1+\sqrt{5}}{2}}{\frac{\lambda_{0}^{2}+2-\lambda_{0} \sqrt{5 \lambda_{0}^{2}+4}}{2 \lambda_{0}^{2}+2}}\right|  \tag{5.27}\\
& =\left|\frac{\frac{1+\sqrt{5}}{2}}{\frac{\pi^{4}+2-\sqrt{5 \pi^{4}+4 \pi^{2}}}{2 \pi^{4}+2}}\right| \approx \frac{1.6180}{0.606127} \approx 2.66946 \tag{5.28}
\end{align*}
$$

So we have that $c_{1} \geq\left\|P^{-1} B\right\|$ and $c_{2} \leq\left\|\left(P^{-1} B\right)^{-1}\right\|$, then $\left|\left(P^{-1} B u, v\right)\right| \leq$ $c_{1}|(u, v)| \leq c_{1}\|u\|\|v\|$. And also, $\inf _{u} \sup _{v} \frac{\left(P^{-1} B u, v\right)}{\|u\|\|v\|} \geq c_{2}^{-1}$. Since the bilinear form is symmetric, $\inf _{v} \sup _{u} \frac{\left(P^{-1} B u, v\right)}{\|u\|\|v\|} \geq c_{2}^{-1}$ is also true. All of the requirements of the Babuška-Lax-Milgram Theorem are meet, so the bilinear form $P^{-1} \circ B[u, v]=L_{p}(v)$ has an unique solution. This can only be the case if $B[u, v]=L_{b}(v)$ has an unique solution.

## Chapter 6

## Technical issues and implementation

### 6.1 Condition numbers

There are three different definitions of condition numbers used in this thesis.
Condition Number Definition 1 Let $A$ be $n \times n$ matrix with real eigenvalues $0 \leq$ $\lambda_{1} \leq \lambda_{2} \leq \cdots \leq \lambda_{n}$. Then the condition number is

$$
\kappa_{1}(A)=\frac{\lambda_{n}}{\lambda_{1}}
$$

Condition Number Definition 2 Let A be $n \times n$ matrix with real eigenvalues $\lambda_{1} \leq$ $\lambda_{2} \leq \cdots \leq \lambda_{n}$. Then the condition number is

$$
\kappa_{2}(A)=\frac{\max _{i}\left|\lambda_{i}\right|}{\min _{j}\left|\lambda_{j}\right|}
$$

Condition Number Definition 3 Let $A$ be a real $m \times n$ matrix with real and complex eigenvalues $\left\{\lambda_{i}\right\}_{i=1}^{n}$. Let $B=A^{T} A$, with real eigenvalues $0 \leq \gamma_{1} \leq \gamma_{2} \leq$ $\cdots \leq \gamma_{n}$

$$
\kappa_{3}(A)=\frac{\sqrt{\gamma_{1}}}{\sqrt{\gamma_{n}}}
$$

### 6.2 Python issues

The simplified $k-\epsilon$ model is implemented in python. The libraries numpy [13], scipy [13] and syFi [14] are used. The inverting algorithm in scipy is to unstable for the purposes of this thesis.

```
scipy.linalg.inv(A)
```

Instead the matrixes are generated in python, and written in ".m" format. Then they where implemented in octave [15], and the "qz" algorithm is used. This approach proved more stable. A sample code for doing this looks like:

```
nMax= 8
def run(N) :
    #calculating the matrixes
    fi_Str = "_"
    for key in AA. keys():
        i,j = key
        d[i][j] = float (AA[key])
```



```
    prCo = zeros([m,m],float)
    for key in PC.keys():
            i,j = key
            prCo[i][j] = float(PC[key])
            fi_Str += "Pc%d(%d, %/d)=%s;\n" % (N,i+1,j+1,prCo[i ][j])
    #eigenvalues of system
        fi_Str += "v%d_=_чеig(A%d) (_)\n" % (N,N)
```



```
    #qz for finding the list of eigenvalues
```



```
    fi_Str += "__eg%d(1)_ц_; \n" % (N)
    fi_Str += "کeg%d(size(eg%d)(1)
    #printing out the absolute values, they are close to 1
```



```
    fi_Str += " ک_minEgenverdi%d_=_min(abs(eg%d))
    fi_Str + = "_Cond%d=maxEgenverdi%d/minEgenverdi%d\n"%(N,N,N)
    return fi_Str
if __name__== " __main__" :
    f = open("testPrec.m",'w')
    fi_str = ""
    for i in range (0, nMax):
    #print "کdette_er_'i's" ,i
        print 2**i
    # N = 125 = breakdown
        a = run(2**i)
```

$$
\begin{array}{r}
\text { fi_str }+=\mathrm{a} \\
\text { f. write (fi_str) }
\end{array}
$$

### 6.3 SyFi

SyFi finite element package SyFi it is a C++ library built on top of the symbolic math library GiNaC [16]. In this thesis it is used to generate the finite element matrixes. This is a sample code used to generate a Hermite and a discontinuous Lagrange reference matrix.

```
import SyFi as SF
SF.initSyFi(2)
l = SF.Line([0.0, 0.0], [1.0/N,0.0])
fe = SF.Hermite(1)
lagOrder = 2
feLa = SF. DiscontinuousLagrange (1,lagOrder - 1)
#hermite mass matrix
A={}
#hermite biharmonic matrix
B = {}
#hermite mass matrix
M= {}
#biharmoic matrix with mass matrix
PaC = {}
weig = N
for i in range(0, 4):
    if i%2 == 1 :
    p1 = weig
    else :
        p1 = 1
    for j in range(0, 4):
        if j%2 ==1 :
            p2 = weig
        else :
            p2 = 1
        M[(i,j)] = 1.integrate(fe.N(i)*fe.N(j))*p1*p2
        B[(i,j)] = l.integrate(swig.diff(swig.diff(fe.N(i),x),x) *
            swig.diff(swig.diff(fe.N(j),x),x))*p1*p2
        A[(i,j)] = M[(i,j)]
        PaC[(i,j)]=B[(i,j)]+M[(i,j)]
#Lagrange mass function
M2 = {}
```

```
for \(i\) in range(0, lagOrder):
    for \(j\) in range ( 0 , lagOrder):
        \(\mathrm{M} 2[(\mathrm{i}, \mathrm{j})]=1\). integrate (feLa. \(\mathrm{N}(\mathrm{i}) *\) feLa. \(\mathrm{N}(\mathrm{j})\) )
\#mixed herm-lag matrix
\(\mathrm{AD}=\{ \}\)
for \(i\) in range \((0,4)\) :
    if \(\mathrm{i} \% 2=1\) :
        p1 = weig
    else :
            \(\mathrm{p} 1=1\)
    for j in range ( 0, lagOrder ):
        \(\operatorname{AD}[(\mathrm{i}, \mathrm{j})]=1 . \operatorname{integrate}(-\mathrm{feLa} . \mathrm{N}(\mathrm{j}) *\) swig.diff(swig.diff(fe.N(
            i) , x) , x) ) *p1
```


## Chapter 7

## Hermite elements

### 7.1 Hermite elements in 1D

The basis functions on the real line, where the element spans $[0,1]$ has the following basis functions

$$
\begin{aligned}
& \phi_{0}=2 x^{3}-3 x^{2}+1 \\
& \phi_{1}=-2 x^{3}+3 x^{2} \\
& \phi_{2}=x^{3}-2 x^{2}+x \\
& \phi_{3}=x^{3}-x^{2} .
\end{aligned}
$$

For a visual representation of the Hermite basis functions see figures: 7.1, 7.2, 7.3,7.4.


Figure 7.1: $\phi_{0}$, the basis function evaluation value at 0 .


Figure 7.2: $\phi_{1}$, the basis function evaluation value at 1.


Figure 7.3: $\phi_{2}$, the basis function evaluation derivative at 0 .


Figure 7.4: $\phi_{3}$, the basis function evaluation derivative at 1.

### 7.2 Scaling Hermite elements 1D

There are a few important issues to notice when working with implementing FEM on this problem. The first issue is that has to be worked out an issue
relating to scaling. When the mesh grid becomes more refined, the norm of different basis functions scales differently for Hermite elements. This issue caused a great deal of problems in the simulations before it was identified. If left unresolved, the matrices quickly becomes very unstable when the mesh is refined. Half of the diagonal values on the Hermite FEM matrix decreases $N^{2}$ times faster than the other diagonal values (as shown later in this chapter). This makes the FEM approximation unstable when confined to double precision.

The first attempt to solve this issue was to increase the mesh size, instead of refining it. The normal procedure for refinement is :

$$
\begin{aligned}
{[0,1] \rightarrow\left\{\left[0, \frac{1}{2}\right],\left[\frac{1}{2}, 1\right]\right\} } & \rightarrow\left\{\left[0, \frac{1}{4}\right],\left[\frac{1}{4}, \frac{2}{4}\right],\left[\frac{2}{4}, \frac{3}{4}\right],\left[\frac{3}{4}, 1\right]\right\} \\
& \rightarrow \cdots \rightarrow\left\{\left[0, \frac{1}{N}\right], \ldots,\left[\frac{N-1}{N}, 1\right]\right\}
\end{aligned}
$$

This was changed to

$$
[0,1] \rightarrow\{[0,1],[1,2]\} \rightarrow \cdots \rightarrow\{[0,1], \ldots,[N-1, N]\}
$$

This refinement mapping was implemented and tested. It solved the issue of the diagonal values but, was discarded for two reasons:

- The mass matrix (matrix representing $I$ ) dominated the eigenvalue analysis.
- The eigenvalue analysis in the literature is dependent on a fixed grid.

To mimic the positive results of the domain scaling idea, scaling of the basis functions is introduced. The basis functions associated with the rapidly decreasing diagonal values are the basis functions representing the derivative. These basis functions are multiplied with a number dependent on the mesh. This solves the issue. Note that after solving a scaled system, the values have to be rescaled to obtain the actual solution.

Next we calculate how to scale the basis functions. We compare how the diagonal values will scale when the mesh is refined. The first integrals are

$$
\begin{array}{r}
\int_{0}^{1} \phi_{0} \mathrm{~d} x=\int_{0}^{1} \phi_{1} \mathrm{~d} x=\frac{1}{2} \\
\int_{0}^{1} \phi_{2} \mathrm{~d} x=-\int_{0}^{1} \phi_{3} \mathrm{~d} x=\frac{1}{12}
\end{array}
$$

By calculating the basis functions for the Hermite element on the domain $\left[0, \frac{1}{N}\right]$
and sets $N=16$ we get:

$$
\begin{aligned}
& \phi_{0}^{N}=8192 x^{3}-768 x^{2}+1 \\
& \phi_{1}^{N}=-8192 x^{3}+768 x^{2} \\
& \phi_{2}^{N}=256 x^{3}-32 x^{2}+x \\
& \phi_{3}^{N}=256 x^{3}-16 x^{2} .
\end{aligned}
$$

The integrals are

$$
\begin{array}{r}
\int_{0}^{\frac{1}{16}} \phi_{0}^{N} \mathrm{~d} x=\int_{0}^{\frac{1}{16}} \phi_{1}^{N} \mathrm{~d} x=\frac{1}{32} \\
\int_{0}^{\frac{1}{16}} \phi_{2}^{N} \mathrm{~d} x=-\int_{0}^{\frac{1}{16}} \phi_{3}^{N} \mathrm{~d} x=\frac{1}{3072}
\end{array}
$$

Divide $\phi_{0}$ on $\phi_{0}^{N}$ and $\phi_{1}$ on $\phi_{1}^{N}$ :

$$
\begin{aligned}
& \frac{\int_{0}^{1} \phi_{0} \mathrm{~d} x}{\int_{0}^{\frac{1}{16}} \phi_{0}^{N} \mathrm{~d} x}=\frac{\frac{1}{2}}{\frac{1}{32}}=16=N \\
& \frac{\int_{0}^{1} \phi_{2} \mathrm{~d} x}{\int_{0}^{\frac{1}{16}} \phi_{2}^{N} \mathrm{~d} x}=\frac{\frac{1}{12}}{\frac{1}{3072}}=256=N^{2}
\end{aligned}
$$

Computed directly the condition number of a matrix produced with unscaled basis functions is bound to run into rounding error. This is illustrated by the table 7.2 (figures: $7.10,7.11,7.12,7.13,7.14$ ). To avoid this problem, new scaled basis functions are produced. The increased stability of the basis functions of equation (7.2) over the natural reference basis functions (7.1), is easily demonstrated by the table 7.1 (figures: $7.5,7.6,7.7,7.8$ and 7.9 ) as compared to table 7.2. As we can see in the table, the $M^{H}$ changes as $O(1)$. The mass matrix represents equations of the type $u=f$ and should therefore have stable condition numbers. $A^{H}$ changes as $O\left(2^{N}\right)$, this is as good as it should be, and a big improvement over the table 7.2. $A^{D}$, scales as $O\left(2^{N}\right)$, but it starts as a very large number. $B$ scales as $O\left(4^{N}\right)$, as it should.

$$
\begin{align*}
& \left\{\begin{array}{l}
\phi_{0}^{N}(0)=1 \\
\frac{\partial \phi_{1}^{N}(0)}{x^{x}}=1 \\
\phi_{2}^{N}(1)=1 \\
\frac{\partial \phi_{3}^{N}(1)}{\partial x}=1
\end{array}\right.  \tag{7.1}\\
& \left\{\begin{array}{l}
\phi_{0}^{N}(0)=N \\
\frac{\partial \phi_{1}^{N}(0)}{\partial x}=N \\
\phi_{2}^{N}(1)=N \\
\frac{\partial \phi_{3}^{N}(1)}{\partial x}=N .
\end{array}\right. \tag{7.2}
\end{align*}
$$

Table 7.1: Condition numbers $\left(\kappa_{1}\right)$ for scaled matrices from a one dimensional grid.

| number of <br> elements | $M^{H}$ <br> cond number | $A^{H}+M^{H}$ <br> cond number | $A^{D}+M^{H}$ <br> cond number | $B+M^{H}$ <br> cond number |
| :---: | :---: | :---: | :---: | :---: |
| $1=2^{0}$ | 1056.67 | 31.99 | $7.1703 \cdot 10^{4}$ | $9.1266 \cdot 10^{2}$ |
| $2=2^{1}$ | 878.62 | 37.05 | $1.9623 \cdot 10^{5}$ | $4.7009 \cdot 10^{3}$ |
| $4=2^{2}$ | 797.01 | 88.31 | $1.7188 \cdot 10^{6}$ | $3.2143 \cdot 10^{4}$ |
| $8=2^{3}$ | 807.56 | 336.34 | $2.8839 \cdot 10^{7}$ | $3.1096 \cdot 10^{5}$ |
| $16=2^{4}$ | 817.33 | 1295.55 | $6.3369 \cdot 10^{8}$ | $3.8925 \cdot 10^{6}$ |
| $32=2^{5}$ | 820.07 | 5058.34 | $1.4050 \cdot 10^{10}$ | $5.5665 \cdot 10^{7}$ |
| $64=2^{6}$ | 820.79 | 19957.34 | $3.1416 \cdot 10^{11}$ | $8.4549 \cdot 10^{8}$ |
| $128=2^{7}$ | 821.97 | 79246.83 | $7.0636 \cdot 10^{12}$ | $1.3197 \cdot 10^{10}$ |
| $256=2^{8}$ | 821.01 | 315790.78 | $1.5919 \cdot 10^{14}$ | $2.0861 \cdot 10^{11}$ |
| $512=2^{9}$ | 821.02 | 1260737.96 | $3.5711 \cdot 10^{15}$ | $3.3179 \cdot 10^{12}$ |
| $1024=2^{10}$ | 821.03 | 5038069.14 | $7.0389 \cdot 10^{16}$ | $5.2948 \cdot 10^{13}$ |

$$
\begin{aligned}
B_{i, j} & =\int \Delta H_{i} \Delta H_{j} \mathrm{~d} x \\
A_{i, j}^{H} & =\int D H_{i} \cdot D H_{j} \mathrm{~d} x \\
A_{i, j}^{D} & =\int-H_{i} \Delta H_{j} \mathrm{~d} x \\
S_{i, j}^{D} & =\int-L_{i} \Delta H_{j} \mathrm{~d} x \\
S_{i, j}^{S} & =\int \nabla L_{i} \cdot \nabla H_{j} \mathrm{~d} x \\
A_{i, j}^{L} & =\int \nabla L_{i} \cdot \nabla L_{j} \mathrm{~d} x \\
M_{i, j}^{S} & =\int H_{i} L_{j} \mathrm{~d} x \\
M_{i, j}^{L} & =\int L_{i} L_{j} \mathrm{~d} x \\
M_{i, j}^{H} & =\int H_{i} H_{j} \mathrm{~d} x
\end{aligned}
$$

Figure 7.5: Plot of the $2-\log$ condition number of scaled matrices in 1D as described in the table 7.1.


Figure 7.6: Plot of the condition number of Scaled mass matrix $\left(M^{H}\right)$ in 1D as described in the table 7.1.


Figure 7.7: Plot of the condition number of Scaled $A^{H}$ matrix in 1D as described in the table 7.1.


Figure 7.8: Plot of the condition number of Scaled $A^{D}$ matrix in 1D as described in the table 7.1.


Figure 7.9: Plot of the condition number of Scaled $B^{H}$ matrix in 1D as described in the table 7.1.


Table 7.2: Condition numbers $\left(\kappa_{1}\right)$ for unscaled matrices from a one dimensional grid.

| number of <br> elements | $M^{H}$ <br> cond number | $A^{H}+M^{H}$ <br> cond number | $A^{D}+M^{H}$ <br> cond number | $B+M^{H}$ <br> cond number |
| :---: | :---: | :---: | :---: | :---: |
| $1=2^{0}$ | $1.0567 \cdot 10^{3}$ | $3.1985 \cdot 10^{1}$ | $7.1703 \cdot 10^{4}$ | $9.1266 \cdot 10^{2}$ |
| $2=2^{1}$ | 3. $4735 \cdot 10^{3}$ | $1.4626 \cdot 10^{2}$ | $6.5010 \cdot 10^{5}$ | $1.2395 \cdot 10^{4}$ |
| $4=2^{2}$ | $1.2588 \cdot 10^{4}$ | $6.3619 \cdot 10^{2}$ | $1.6921 \cdot 10^{7}$ | $1.88138 \cdot 10^{5}$ |
| $8=2^{3}$ | 5. $1021 \cdot 10^{4}$ | 2. $6859 \cdot 10^{3}$ | $6.0728 \cdot 10^{8}$ | $2.8449 \cdot 10^{6}$ |
| $16=2^{4}$ | 2. $0660 \cdot 10^{5}$ | $1.0969 \cdot 10^{4}$ | 2. $4487 \cdot 10^{10}$ | $4.3518 \cdot 10^{7}$ |
| $32=2^{5}$ | 8. $2922 \cdot 10^{5}$ | $4.4140 \cdot 10^{4}$ | $1.0434 \cdot 10^{12}$ | $6.7675 \cdot 10^{8}$ |
| $64=2^{6}$ | 3. $3198 \cdot 10^{7}$ | $1.7684 \cdot 10^{5}$ | $4.5805 \cdot 10^{13}$ | $1.0656 \cdot 10^{10}$ |
| $128=2^{7}$ | 1. $3282 \cdot 10^{8}$ | $7.0769 \cdot 10^{5}$ | 2. $0494 \cdot 10^{15}$ | $1.6905 \cdot 10^{11}$ |
| $256=2^{8}$ | 5. $3132 \cdot 10^{8}$ | 2. $8311 \cdot 10^{6}$ | $1.0096 \cdot 10^{17}$ | 2. $6930 \cdot 10^{12}$ |
| $512=2^{9}$ | 2. $1253 \cdot 10^{9}$ | $1.1325 \cdot 10^{7}$ | 3. $1641 \cdot 10^{18}$ | $4.2990 \cdot 10^{13}$ |
| $1024=2^{10}$ | 8. $5013 \cdot 10^{9}$ | $4.5298 \cdot 10^{7}$ | 5. $9802 \cdot 10^{20}$ | 6. $8696 \cdot 10^{14}$ |

Figure 7.10: Plot of the $2-\log$ condition number of unscaled matrix in 1D as described in the table 7.2.


Figure 7.11: Plot of the condition number of unscaled mass matrix in 1D as described in the table 7.2.


### 7.3 Hermite elements 2D

For a two dimensional finite element mesh, the questions how to translate the scaling are not trivial. In this section, two different scalings are tested.

The set of basis functions for a two-dimensional Hermite element is a tensor product between the basis function sets of two 1D elements. These one dimensional elements are the Hermite element running along one of the edges, and of a Hermite element running along one of the other edges normal to the first edge, I will call them $L_{1}, L_{2}$. Each basis functions of a Hermite-2D element on a rectangle is a multiplication between one of basis functions $L_{1}$, and one of the basis functions of $L_{2}$. The set $\left\{\phi_{i}(x, y)\right\}_{2 D}=\left\{\phi_{i}(x) \phi_{j}(y)\right\}_{i, j=1,2,3,4}$. There is a total of 16 basis functions for the 2-D Hermite element.

Figure 7.12: Plot of the condition number of unscaled $A^{H}$ matrix in 1D as described in the table 7.2.


Figure 7.13: Plot of the condition number of unscaled $A^{D}$ matrix in 1D as described in the table 7.2.


Figure 7.14: Plot of the condition number of unscaled $B^{H}$ matrix in 1D as described in the table 7.2.


In 2D it is therefore natural to assume that the scaling also should be multiplied. For simplicity only the four basis functions evaluating the reference square in the point $(0,0)$ will be provided when describing the scaling of the basis functions. The unscaled basis functions that evaluate the reference element are calculated so that they are zero in every evaluation point away from $(0,0)$. In the point $(0,0)$ they will conform to the following conditions;

$$
\begin{aligned}
\frac{\partial \phi_{0}(0,0)}{\partial x} & =\frac{\partial \phi_{0}(0,0)}{\partial y}=\frac{\partial^{2} \phi_{0}(0,0)}{\partial x \partial y}=0 \\
\phi_{1}(0,0) & =\frac{\partial \phi_{1}(0,0)}{\partial y}=\frac{\partial^{2} \phi_{1}(0,0)}{\partial x \partial y}=0 \\
\phi_{2}(0,0) & =\frac{\partial \phi_{2}(0,0)}{\partial x}=\frac{\partial^{2} \phi_{2}(0,0)}{\partial x \partial y}=0 \\
\phi_{3}(0,0) & =\frac{\partial \phi_{3}(0,0)}{\partial x}=\frac{\partial \phi_{3}(0,0)}{\partial y}=0 \\
\phi_{0}(0,0) & =\frac{\partial \phi_{1}(0,0)}{\partial x}=\frac{\partial \phi_{2}(0,0)}{\partial y}=\frac{\partial^{2} \phi_{3}(0,0)}{\partial x \partial y}=1
\end{aligned}
$$

A first suggestion for scaling of these basis functions will look like;

$$
\begin{gathered}
\phi_{0}(0,0)=1 \\
\frac{\partial \phi_{1}^{N}(0,0)}{\partial x}=N \\
\frac{\partial \phi_{2}^{N}(0,0)}{\partial y}=N \\
\frac{\partial^{2} \phi_{3}^{N}(0,0)}{\partial x \partial y}=N^{2} .
\end{gathered}
$$

This is the most obvious choice, since this is the direct tensor product of the scaled one dimensional elements from previous sections. Running the simulation again but with these modifications gave the condition numbers of tables 7.3 and 7.4 (figures: $7.15,7.16,7.17,7.18,7.19$ and 7.20 ). $M^{H}$ behaves good. $A^{H}$ does not develop as it should, ideally it should scale as $O\left(4^{N}\right) . A^{D}$ scales as it should, but it starts out with a fairly big condition number. $B_{1}$ scales as it should, but it starts out with a fairly big condition number. $B_{2}$ scales as it should, and seems to be the best choice for a biharmonic matrix.

Table 7.3: Condition numbers $\left(\kappa_{1}\right)$ for matrices representing a 2 - dimensional grid with fully scaled basis functions.

| number of elements | $\begin{gathered} M^{H} \\ \text { cond number } \end{gathered}$ | $\begin{gathered} A^{H}+M^{H} \\ \text { cond number } \end{gathered}$ | $A^{D}+M^{H}$ cond number |
| :---: | :---: | :---: | :---: |
| $1=\left(2^{0}\right)^{2}$ | 1. $1165 \cdot 10^{6}$ | 1. $6961 \cdot 10^{4}$ | 1. $7368 \cdot 10^{9}$ |
| $4=\left(2^{1}\right)^{2}$ | 7. $7197 \cdot 10^{5}$ | 2. $1925 \cdot 10^{4}$ | 5. $3687 \cdot 10^{9}$ |
| $16=\left(2^{2}\right)^{2}$ | 6. $3522 \cdot 10^{5}$ | 1. $7823 \cdot 10^{4}$ | 8. $8709 \cdot 10^{10}$ |
| $64=\left(2^{3}\right)^{2}$ | 6. $5215 \cdot 10^{5}$ | 1. $7296 \cdot 10^{4}$ | 5. $7546 \cdot 10^{12}$ |
| $256=\left(2^{4}\right)^{2}$ | 6. $6803 \cdot 10^{5}$ | 1. $7616 \cdot 10^{4}$ | 6. $9134 \cdot 10^{14}$ |
| $1024=\left(2^{5}\right)^{2}$ | 6. $7252 \cdot 10^{5}$ | 1. $7722 \cdot 10^{4}$ | 9. $5156 \cdot 10^{15}$ |

Table 7.4: Condition numbers $\left(\kappa_{1}\right)$ for the two types biharmonic matrices representing a 2- dimensional grid with fully scaled basis functions.

| number of <br> elements | $B_{2}+M^{H}$ <br> cond number | $B_{1}+M^{H}$ <br> cond number |  |  |
| ---: | :---: | :---: | :---: | :---: |
| $1=\left(2^{0}\right)^{2}$ | 2. | $4324 \cdot 10^{3}$ | 2. | $7417 \cdot 10^{5}$ |
| $4=\left(2^{1}\right)^{2}$ | 1. | $1152 \cdot 10^{4}$ | 1. | $2095 \cdot 10^{6}$ |
| $16=\left(2^{2}\right)^{2}$ | 7. | $4394 \cdot 10^{4}$ | 1. | $6625 \cdot 10^{6}$ |
| $64=\left(2^{3}\right)^{2}$ | 6. | $6903 \cdot 10^{5}$ | 3. | $8788 \cdot 10^{6}$ |
| $256=\left(2^{4}\right)^{2}$ | 7. | $9697 \cdot 10^{6}$ | 2. | $0004 \cdot 10^{7}$ |
| $1024=\left(2^{5}\right)^{2}$ | 1. | $1081 \cdot 10^{7}$ | 1. | $9406 \cdot 10^{8}$ |

Figure 7.15: Plot of the $2-\log$ condition number of scaled matrices in 2D as described in the tables 7.3 and 7.4.


Figure 7.16: Plot of the condition number of Scaled mass matrix in 2D as described in the table 7.3.


Figure 7.17: Plot of the condition number of Scaled $A^{H}$ matrix in 2D as described in the table 7.3.


Figure 7.18: Plot of the condition number of Scaled $A^{D}$ matrix in 2D as described in the table 7.3.


Figure 7.19: Plot of the condition number of Scaled $B_{2}^{H}$ matrix in 2D as described in the table 7.4.


Figure 7.20: Plot of the condition number of Scaled $B_{1}^{H}$ matrix in 2D as described in the table 7.4.


This scaling is not the only scaling possible. To test the hypothesis that the full scaling model is the best option, the "Reduced scaling" will be introduced. The Reduced scaling, is a weaker form of scaling. It is different than the full scaling in how it treats the basis functions $\phi_{4 i}$. These are the basis functions that conforms the suggested solution with both the derivative in $x$ and $y$ direction. The reduced scaling is defined by the following conditions:

$$
\begin{aligned}
\phi_{0}(0,0) & =1 \\
\frac{\partial \phi_{1}^{N}(0,0)}{\partial x} & =N \\
\frac{\partial \phi_{2}^{N}(0,0)}{\partial y} & =N \\
\frac{\partial^{2} \phi_{3}^{N}(0,0)}{\partial x \partial y} & =N .
\end{aligned}
$$

Running simulations with the reduced scaling gives the condition numbers of the tables 7.5 and 7.6 (figures: 7.21, $7.22,7.23,7.24,7.25$ and 7.26 ). The $A^{H}$ is the only condition number that benefits from the reduced scaling over the full scaling.

Table 7.5: Condition numbers $\left(\kappa_{1}\right)$ for matrices representing a 2 - dimensional grid with reduced scaled basis functions.

| number of elements | $M^{H}$ cond number | $A^{H}+M^{H}$ cond number | $A^{D}+M^{H}$ cond number |
| :---: | :---: | :---: | :---: |
| $1=\left(2^{0}\right)^{2}$ | 1. $1165 \cdot 10^{6}$ | 1. $6961 \cdot 10^{4}$ | 1. $7368 \cdot 10^{9}$ |
| $4=\left(2^{1}\right)^{2}$ | 3. $0413 \cdot 10^{6}$ | 8. $6383 \cdot 10^{4}$ | 1. $5488 \cdot 10^{10}$ |
| $16=\left(2^{2}\right)^{2}$ | 9. $9365 \cdot 10^{6}$ | 2. $7881 \cdot 10^{5}$ | 5. $6817 \cdot 10^{11}$ |
| $64=\left(2^{3}\right)^{2}$ | 4. $0717 \cdot 10^{7}$ | 1. $0799 \cdot 10^{6}$ | 4. $7056 \cdot 10^{13}$ |
| $256=\left(2^{4}\right)^{2}$ | 1. $6678 \cdot 10^{8}$ | 4. $3982 \cdot 10^{6}$ | 9. $6000 \cdot 10^{15}$ |
| $1024=\left(2^{5}\right)^{2}$ | 6. $7154 \cdot 10^{8}$ | 1. $7698 \cdot 10^{7}$ | 3. $5710 \cdot 10^{16}$ |

Table 7.6: Condition numbers $\left(\kappa_{1}\right)$ for the two types biharmonic matrices representing a 2 - dimensional grid with reduced scaled basis functions.

| number of <br> elements | $B_{2}+M^{H}$ <br> cond number | $B_{1}+M^{H}$ <br> cond number |  |  |
| ---: | :---: | :---: | :---: | :---: |
| $1=\left(2^{0}\right)^{2}$ | 2. | $4324 \cdot 10^{3}$ | 2. | $7417 \cdot 10^{5}$ |
| $4=\left(2^{1}\right)^{2}$ | 1. | $3862 \cdot 10^{4}$ | 3. | $0276 \cdot 10^{6}$ |
| $16=\left(2^{2}\right)^{2}$ | 7. | $4437 \cdot 10^{4}$ | 8. | $1370 \cdot 10^{6}$ |
| $64=\left(2^{3}\right)^{2}$ | 6. | $6903 \cdot 10^{5}$ | 2. | $6019 \cdot 10^{7}$ |
| $256=\left(2^{4}\right)^{2}$ | 7. | $9697 \cdot 10^{6}$ | 2. | $8929 \cdot 10^{8}$ |
| $1024=\left(2^{5}\right)^{2}$ | 1. | $1081 \cdot 10^{7}$ | 3. | $4510 \cdot 10^{9}$ |

Figure 7.21: Plot of the $2-\log$ condition number of reduced scaled matrices in 2D.


Figure 7.22: Plot of the condition number of reduced scaled mass matrix in 2D as described in the tables 7.5 and 7.6.


Figure 7.23: Plot of the condition number of reduced scaled $A^{H}$ matrix in 2D as described in the table 7.5.


Figure 7.24: Plot of the condition number of reduced scaled $A^{D}$ matrix in 2D as described in the table 7.5.


Figure 7.25: Plot of the condition number of Scaled $B_{2}^{H}$ matrix in 2D as described in the table 7.6.


Figure 7.26: Plot of the condition number of reduced scaled $B_{1}^{H}$ matrix in 2D as described in the table 7.6.


## Chapter 8

## Solving in 1D

### 8.1 Solving as a linear PDE system

### 8.1.1 Representation from $H^{2}(\Omega) \times L^{2}(\Omega)$.

As described, there are two competing weak representation (4.4) and (4.5). This section will look at the $H^{2} \times L^{2}$ representation. We now select $v_{1}, \epsilon \in$ $H^{2}(\Omega)$ and $v_{2}, k \in L^{2}(\Omega)$. With this choice we use the equation (4.5) as our weak form (bilinear). The finite element to use will be the Hermite elements for $v_{1}, \epsilon$, and the Lagrange elements for $v_{2}, k$. In order to increase the accuracy and to keep the evaluations stable the choice to use the Lagrange-2 elements was made. The reasoning behind this choice is that if a line is divided into $N$ different elements, then Hermite elements will give $2 \cdot(N+1)$ evaluations, while a Lagrange- 2 will give $2 \cdot N+1$ evaluations. So combining this knowledge, the matrices that will be tested are:

$$
\begin{aligned}
\operatorname{Co}_{D} & =\left[\begin{array}{cc}
M^{H} & S^{D} \\
\left(S^{D}\right)^{T} & M^{L}
\end{array}\right] \\
\operatorname{PrCo} & =\left[\begin{array}{cc}
B+M^{H} & 0 \\
0 & M^{L}
\end{array}\right]^{-1}\left[\begin{array}{cc}
M^{H} & S^{D} \\
\left(S^{D}\right)^{T} & M^{L}
\end{array}\right] .
\end{aligned}
$$

The matrix $\operatorname{PrCo}$ is not symmetric and will have negative eigenvalues. The definition of the condition number used is $\kappa(\operatorname{PrCo})=\frac{\max _{i}\left\{\left|\lambda \lambda_{i}\right|\right\}}{\left.\min _{i}\left\{\mid \lambda_{i}\right\}\right\}}$. With this in mind, the key matrices of the system is in table 8.1. The preconditioned system in 8.1 does not scale as well as it should, and therefore the implementation does not work properly. There is some minor issues with the $S^{D}+M^{S}$ for small meshes, but it seems to work out.

Table 8.1: Condition numbers for scaled matrices from a one dimensional grid

| number of <br> elements | $\kappa_{3}\left(S^{D}+M^{S}\right)$ <br> condition number | $\kappa_{1}\left(C o o_{D}\right)$ <br> condition number | $\kappa(\operatorname{PrCo})$ <br> condition number |
| :---: | :---: | :---: | :---: |
| $1=2^{0}$ | 371.19 | $7.2404 \cdot 10^{1}$ | $1.000 \cdot 10^{0}$ |
| $2=2^{1}$ | 17.117 | $1.1163 \cdot 10^{2}$ | $1.3417 \cdot 10^{0}$ |
| $4=2^{2}$ | 17.662 | $4.4115 \cdot 10^{2}$ | $3.2533 \cdot 10^{0}$ |
| $8=2^{3}$ | 30.475 | $1.7138 \cdot 10^{3}$ | $1.5295 \cdot 10^{1}$ |
| $16=2^{4}$ | 57.266 | 6. $6693 \cdot 10^{3}$ | $6.2036 \cdot 10^{1}$ |
| $32=2^{5}$ | 106.13 | $2.6186 \cdot 10^{4}$ | $2.4880 \cdot 10^{2}$ |
| $64=2^{6}$ | 199.96 | $1.0362 \cdot 10^{5}$ | $9.959 \cdot 10^{2}$ |
| $128=2^{7}$ | 385.74 | $4.1209 \cdot 10^{5}$ | no data |
| $256=2^{8}$ | 756.62 | $1.6434 \cdot 10^{6}$ | no data |
| $512=2^{9}$ | 1498.2 | 6. $2635 \cdot 10^{6}$ | no data |
| $1024=2^{10}$ | 2981.2 | 2. $2233 \cdot 10^{7}$ | no data |

### 8.1.2 Representations from $H^{1}(\Omega) \times H^{1}(\Omega)$.

As described, there are two competing weak representation (4.4), (4.5). This section will look at the $H^{1} \times H^{1}$ representation.

$$
\int_{\Omega}\left[\begin{array}{cc}
v_{1} \epsilon & \nabla k \cdot \nabla v_{1} \\
\nabla v_{2} \cdot \nabla \epsilon & v_{2} k
\end{array}\right] d x=\int_{\Omega}\left[\begin{array}{c}
v_{1} f_{1} \\
v_{2} f_{2}
\end{array}\right] d x
$$

In this representation we choose $v_{1}, v_{2}, \epsilon, k \in H^{1}(\Omega)$, and solve the systems with the following system matrices.

$$
\begin{aligned}
\operatorname{Co}_{N} & =\left[\begin{array}{cc}
M^{L} & A^{L} \\
\left(A^{L}\right)^{T} & M^{L}
\end{array}\right] \\
\operatorname{PrCo}_{N} & =\left[\begin{array}{cc}
S^{L}+M^{L} & 0 \\
0 & S^{L}+M^{L}
\end{array}\right]^{-1}\left[\begin{array}{cc}
M^{L} & S^{L} \\
\left(S^{L}\right)^{T} & M^{L}
\end{array}\right]
\end{aligned}
$$

where $N$ is the degrees of freedom allowed in the Lagrange FEM approximation. The eigenvalues of this system is given in the tables 8.2 and 8.3 (figures: 8.1, 8.2, 8.3, 8.4). All the results in the tables 8.2 and 8.3 does behave as expected and we can conclude that $\mathrm{Co}_{2}$ and $\mathrm{Co}_{3}$ are successful implementations.

Table 8.2: Condition numbers $\left(\kappa_{1}\right)$ for Lagrange-type matrices from a one dimensional grid

| number of <br> elements | $\mathrm{Co}_{2}$ <br> condition number | $\mathrm{PrCo}_{2}$ <br> condition number |
| :---: | :---: | :---: |
| $1=2^{0}$ | 4.3333 | 1.1818 |
| $2=2^{1}$ | 18.640 | 2.0026 |
| $4=2^{2}$ | 73.041 | 1.8309 |
| $8=2^{3}$ | 279.99 | 1.5931 |
| $16=2^{4}$ | 1079.4 | 1.4500 |
| $32=2^{5}$ | 4215 | 1.3845 |
| $64=2^{6}$ | 16631 | 1.3374 |
| $128=2^{7}$ | 66039 | 1.3041 |
| $256=2^{8}$ | 263158 | 1.2807 |
| $512=2^{9}$ | 1050614 | 1.2644 |
| $1024=2^{10}$ | 4198390 | 1.2529 |

Table 8.3: Condition numbers $\left(\kappa_{1}\right)$ for Lagrange-type matrices from a one dimensional grid

| number of <br> elements | $\mathrm{Co}_{3}$ <br> condition number | $\mathrm{PrCo}_{3}$ <br> condition number |
| :---: | :---: | :---: |
| $1=2^{0}$ | 25.453 | 3.7338 |
| $2=2^{1}$ | 96.610 | 2.9384 |
| $4=2^{2}$ | 371.52 | 2.4709 |
| $8=2^{3}$ | 1436.78 | 2.2112 |
| $16=2^{4}$ | 5617.36 | 2.0722 |
| $32=2^{5}$ | 22172. | 1.9997 |
| $64=2^{6}$ | 88048 | 1.9627 |
| $128=2^{7}$ | 350875 | 1.9439 |
| $256=2^{8}$ | 1400816 | 1.9344 |
| $512=2^{9}$ | 5597851 | 1.9297 |
| $1024=2^{10}$ | 22380528 | 1.9273 |

Figure 8.1: Plot of the $2-\log$ condition number of Lagrange-2, Lagrange-3 and the preconditioned systems matrices in 1D as described in the tables 8.2 and 8.3


Figure 8.2: Plot of the condition number of Co 2 Lagrange-2 system 1D as described in the table 8.2


Figure 8.3: Plot of the condition number of Co3 Lagrange-3 system 1D as described in the table 8.3


Figure 8.4: Plot of the condition number the preconditioned Lagrange-2 and the preconditioned Lagrange-3 systems matrices in 1D as described in the tables 8.2 and 8.3

8.1.3 Representations from $H^{2}(\Omega) \times H^{1}(\Omega)$.

Another possible choice for the function spaces of the tests and trail functions is to chose $v_{2}, k \in H^{1}(\Omega), v_{1}, \epsilon \in H^{2}(\Omega)$. With this representation the following system is produced:

$$
\begin{aligned}
\operatorname{Co}_{S} & =\left[\begin{array}{cc}
M^{H} & S^{S}+M^{S} \\
\left(S^{S}+M^{S}\right)^{T} & M^{L}
\end{array}\right] \\
\operatorname{PrCo} & =\left[\begin{array}{cc}
B & 0 \\
0 & M^{L}
\end{array}\right]^{-1}\left[\begin{array}{cc}
M^{H} & S^{S}+M^{S} \\
\left(S^{S}+M^{S}\right)^{T} & M^{L}
\end{array}\right] .
\end{aligned}
$$

Before assembling the matrix $\operatorname{Co}_{S}$ it is important to note that choosing to use the Lagrange-1 representation will produce issues with the stiffness matrix $A^{S}$. Lets assume that we have the ordinary Hermite element. Then on a linear element $[a, b]$ the basis functions $\phi_{0}^{\tau}, \phi_{1}^{\tau}, \phi_{2}^{\tau}, \phi_{3}^{\tau}$ are as previously defined. The Lagrange functions will be $L_{0}=\frac{x}{a-b}+\frac{b}{b-a}$ and $L_{1}=\frac{x}{b-a}+\frac{a}{a-b}$. We then get

$$
\begin{aligned}
\int_{b}^{a} \phi_{0}^{\tau}(x)^{\prime} L_{0}(x)^{\prime} \mathrm{d} x & =\int_{b}^{a} \phi_{0}^{\tau}(x)^{\prime} \frac{1}{a-b} \mathrm{~d} x \\
& =\frac{1}{a-b}\left(\phi_{0}^{\tau}(a)-\phi_{0}^{\tau}(b)\right)=\frac{1}{a-b}(1-0)=\frac{1}{a-b} \\
\int_{b}^{a} \phi_{0}^{\tau}(x)^{\prime} L_{1}(x)^{\prime} \mathrm{d} x & =\int_{b}^{a} \phi_{0}^{\tau}(x)^{\prime} \frac{1}{b-a} \mathrm{~d} x \\
& =\frac{1}{b-a}\left(\phi_{0}^{\tau}(a)-\phi_{0}^{\tau}(b)\right)=\frac{1}{b-a}(1-0)=-\frac{1}{a-b} \\
\int_{b}^{a} \phi_{1}^{\tau}(x)^{\prime} L_{0}(x)^{\prime} \mathrm{d} x & =\int_{b}^{a} \phi_{1}^{\tau}(x)^{\prime} \frac{1}{a-b} \mathrm{~d} x \\
& =\frac{1}{a-b}\left(\phi_{1}^{\tau}(a)-\phi_{1}^{\tau}(b)\right)=\frac{1}{a-b}(0-0)=0 \\
\int_{b}^{a} \phi_{1}^{\tau}(x)^{\prime} L_{1}(x)^{\prime} \mathrm{d} x & =\int_{b}^{a} \phi_{1}^{\tau}(x)^{\prime} \frac{1}{b-a} \mathrm{~d} x \\
& =\frac{1}{b-a}\left(\phi_{1}^{\tau}(a)-\phi_{1}^{\tau}(b)\right)=\frac{1}{b-a}(0-0)=0 .
\end{aligned}
$$

With these kinds of calculations one can create the table 8.4 As showed in chapter 7 scaling the Hermite elements basis functions $\phi^{\tau}$ with a factor $C(\tau)=$ $\frac{1}{|a-b|}=\frac{1}{N}$ has proven to be beneficial with regards to the condition number of the FEM-matrix. The variant of 8.4 for scaled elements is the table 8.5 To avoid this problem the Lagrange- 2 scheme is used and it creates the table 8.6. The matrix $\mathrm{PrCo}_{s}$ in table 8.6 is very bad, we conclude that this implementation does not work.

Table 8.4: Values of inner products between Lagrange-1 and Hermite basis functions on a single linear element

| $(\cdot, \cdot)_{L^{2}(\tau)}$ | $\phi_{0}^{\tau}(x)^{\prime}$ | $\phi_{1}^{\tau}(x)^{\prime}$ | $\phi_{2}^{\tau}(x)^{\prime}$ | $\phi_{3}^{\tau}(x)^{\prime}$ |
| :---: | :---: | :---: | :---: | :---: |
| $L_{0}^{\prime}$ | $\frac{1}{a-b}$ | 0 | $-\frac{1}{a-b}$ | 0 |
| $L_{1}^{\prime}$ | $-\frac{1}{a-b}$ | 0 | $\frac{1}{a-b}$ | 0 |

Table 8.5: Values of inner products between Lagrange-1 and scaled Hermite basis functions on a single linear element

| $(\cdot, \cdot)_{L^{2}(\tau)}$ | $\phi_{0}^{\tau}(x)^{\prime}$ | $\phi_{1}^{\tau}(x)^{\prime}$ | $\phi_{2}^{\tau}(x)^{\prime}$ | $\phi_{3}^{\tau}(x)^{\prime}$ |
| :---: | :---: | :---: | :---: | :---: |
| $L_{0}^{\prime}$ | 1 | 0 | -1 | 0 |
| $L_{1}^{\prime}$ | -1 | 0 | 1 | 0 |

Table 8.6: Condition numbers $\left(\kappa_{3}\right)$ for combined matrices using Lagrange-2

| number of elements | $S^{S}+M^{S}$ <br> condition number | $\mathrm{Cos}_{S}$ condition number | $\mathrm{PrCos}_{S}$ <br> condition number |
| :---: | :---: | :---: | :---: |
| $1=2^{0}$ | 5. 8156 | 9. $1885 \cdot 10^{2}$ | 1. $5119 \cdot 10^{17}$ |
| $2=2^{1}$ | 11. 321 | 4. $8547 \cdot 10^{3}$ | 2. $9019 \cdot 10^{19}$ |
| $4=2^{2}$ | 13. 790 | 2. $0227 \cdot 10^{4}$ | 2. $5807 \cdot 10^{17}$ |
| $8=2^{3}$ | 22. 521 | 7. $3769 \cdot 10^{4}$ | 9. $8466 \cdot 10^{16}$ |
| $16=2^{4}$ | 42. 717 | 2. $7072 \cdot 10^{5}$ | 2. $5269 \cdot 10^{17}$ |
| $32=2^{5}$ | 83. 158 | 1. $0307 \cdot 10^{6}$ | 1. $2922 \cdot 10^{17}$ |
| $64=2^{6}$ | 163. 92 | 4. $0236 \cdot 10^{6}$ | 1. $6143 \cdot 10^{17}$ |
| $128=2^{7}$ | 325. 38 | 1. $5911 \cdot 10^{7}$ | 8. $4882 \cdot 10^{16}$ |
| $256=2^{8}$ | 648. 27 | 6. $3276 \cdot 10^{7}$ | 8. $6149 \cdot 10^{16}$ |
| $512=2^{9}$ | 1294. 0 | 2. $5238 \cdot 10^{8}$ | 9. $8993 \cdot 10^{16}$ |
| $1024=2^{10}$ | 2585. 5 | 1. $0081 \cdot 10^{9}$ | 2. $6523 \cdot 10^{19}$ |

Table 8.7: Condition numbers $\left(\kappa_{3}\right)$ for combined matrices using Lagrange-2 without mass matrix

| number of <br> elements | $S^{S}$ <br> condition number | $C o_{S H}$ <br> condition number |
| :---: | :---: | :---: |
| $1=2^{0}$ | $2.9565 \cdot 10^{16}$ | $9.1885 \cdot 10^{2}$ |
| $2=2^{1}$ | 17.220 | $3.8944 \cdot 10^{3}$ |
| $4=2^{2}$ | 14.997 | $1.5337 \cdot 10^{4}$ |
| $8=2^{3}$ | 22.813 | $5.8823 \cdot 10^{4}$ |
| $16=2^{4}$ | 42.840 | $2.2672 \cdot 10^{5}$ |
| $32=2^{5}$ | 83.217 | 8. $8523 \cdot 10^{5}$ |
| $64=2^{6}$ | 163.95 | $3.4926 \cdot 10^{6}$ |
| $128=2^{7}$ | 325.39 | $1.3868 \cdot 10^{7}$ |
| $256=2^{8}$ | 648.27 | $5.5263 \cdot 10^{7}$ |
| $512=2^{9}$ | 1294.0 | $2.2063 \cdot 10^{8}$ |
| $1024=2^{10}$ | 2585.5 | $88166 \cdot 10^{8}$ |

As one can clearly see this representation of the problem creates bad results, and should not be used. A modification of this representation can be:

$$
\begin{aligned}
& \operatorname{Co}_{S H}=\left[\begin{array}{cc}
M^{H} & S^{S} \\
\left(S^{S}\right)^{T} & M^{L}
\end{array}\right] \\
& \operatorname{PrCo} \\
& S H=\left[\begin{array}{cc}
B+M^{H} & 0 \\
0 & M^{L}
\end{array}\right]^{-1}\left[\begin{array}{cc}
M^{H} & S^{S} \\
\left(S^{S}\right)^{T} & M^{L}
\end{array}\right] \\
& \operatorname{PrCo}_{A H}=\left[\begin{array}{cc}
S^{H}+M^{H} & 0 \\
0 & S^{L}+M^{L}
\end{array}\right]^{-1}\left[\begin{array}{cc}
M^{H} & S^{S} \\
\left(S^{S}\right)^{T} & M^{L}
\end{array}\right],
\end{aligned}
$$

where

$$
S_{i, j}^{S}=\int \nabla L_{i} \cdot \nabla H_{j} \mathrm{~d} x
$$

It produces the tables 8.7 and 8.8 (figures: 8.5, 8.6, 8.7). The condition number of $S^{S}$ is very high, this can be expected. Derivatives without a mass matrix is usually badly conditioned when there is only one element in the mesh grid. The condition numbers of $\mathrm{Co}_{S H}$ are good, but none of the two tested preconditioner does not work. This might be because we used a $H^{2}$ conforming element to approximate a $H^{1}$ function.

Table 8.8: Condition numbers ( $\kappa_{3}$ ) for preconditioned combined matrices using Lagrange-2 without mass matrix
$\left.\begin{array}{|c||c|c|}\hline \begin{array}{c}\text { number of } \\ \text { elements }\end{array} & \begin{array}{c}\operatorname{PrCo} \\ S H\end{array} \\ \text { condition number }\end{array} \quad \begin{array}{c}\operatorname{PrCo} \\ \text { AH } \\ \text { condition number }\end{array}\right]$

Figure 8.5: Plot of the $2-\log$ condition number of the preconditioned Lagrange-3-Hermite systems matrices in 1D as described in table 8.8


Figure 8.6: Plot of the condition number of the preconditioned Lagrange-3Hermite systems matrices in 1D as described in table of $\operatorname{PrCo}_{S H}$ in 8.8


Figure 8.7: Plot of the condition number of the preconditioned Lagrange-3Hermite systems matrices in 1D as described in table of $\operatorname{PrCo}_{S H}$ in 8.8


### 8.2 Solving as a Biharmonic equation

Assuming $(\epsilon, k) \in H^{2}(\Omega) \times L^{2}(\Omega)$

$$
\left.\left.\begin{array}{c}
\left\{\begin{array}{c}
-\Delta \epsilon+k=f_{1} \mid \Delta \\
\epsilon-\Delta k=f_{2}
\end{array}\right. \\
\left\{\left.\begin{array}{c}
-\Delta \Delta \epsilon+\Delta k=\Delta f_{1} \\
\epsilon-\Delta k=f_{2}
\end{array} \right\rvert\,+(\mathrm{II})\right.
\end{array}\right\} \begin{array}{r}
-\Delta \Delta \epsilon+\Delta k+\epsilon-\Delta k=\Delta f_{1}+f_{2} \\
-\Delta \Delta \epsilon+\epsilon=\Delta f_{1}+f_{2}
\end{array}\right\} \begin{aligned}
& \int_{\Omega}-v \Delta \Delta \epsilon+v \epsilon \mathrm{~d} x=\int_{\Omega} v \Delta f_{1}+v f_{2} \mathrm{~d} x \\
& \int_{\Omega}-\Delta v \Delta \epsilon+v \epsilon \mathrm{~d} x=\int_{\Omega} v \Delta f_{1}+v f_{2} \mathrm{~d} x+\text { boundery terms }
\end{aligned}
$$

the discrete representation of this problem is

$$
\left[-B^{H}+M^{H}\right]
$$

The table for the condition numbers was described in the table 7.1. It is not preconditioned, but it scales as it should.

### 8.2.1 Representation from $H^{2}(\Omega) \times L^{2}(\Omega)$. Discontinuous Lagrange

In this section the Problem is implemented with discontinuous Galerkin. This section will look at the $H^{2} \times L^{2}$ representation. We now select $v_{1}, \epsilon \in H^{2}(\Omega)$ and $v_{2}, k \in L^{2}(\Omega)$. With this choice we use the equation (4.5) as our weak form (bilinear). The finite element to use will for $v_{1}, \epsilon$ be the Hermite elements, while for $v_{2}, k$ will be the Discontinuous Lagrange elements.

$$
\begin{aligned}
B_{i, j} & =\int \Delta H_{i} \Delta H_{j} \mathrm{~d} x \\
S_{i, j}^{d D} & =\int-d L_{i} \Delta H_{j} \mathrm{~d} x \\
M_{i, j}^{d L} & =\int d L_{i} d L_{j} \mathrm{~d} x \\
M_{i, j}^{H} & =\int H_{i} H_{j} \mathrm{~d} x
\end{aligned}
$$

Table 8.9: Condition numbers ( $\kappa_{2}$ ) for preconditioned scaled combined matrices using discontinuous Lagrange-2

| number of <br> elements | $\operatorname{PrCo}_{D}$ <br> condition number | $C o_{D H}$ <br> condition number |
| :---: | :---: | :---: |
| $1=2^{0}$ | 2.6249 | 72.404 |
| $2=2^{1}$ | 2.6279 | 78.838 |
| $4=2^{2}$ | 2.6279 | 123.82 |
| $8=2^{3}$ | 2.6280 | 290.23 |
| $16=2^{4}$ | 2.6280 | 899.37 |
| $32=2^{5}$ | 2.6280 | 3206.4 |
| $64=2^{6}$ | 2.6280 | $1.2166 e+04$ |
| $128=2^{7}$ | 2.6280 | $4.7465 \mathrm{e}+04$ |

$$
\begin{aligned}
C o_{D H} & =\left[\begin{array}{cc}
M^{H} & S^{d D} \\
\left(S^{d D}\right)^{T} & M^{L}
\end{array}\right] \\
\operatorname{PrCo}_{D} & =\left[\begin{array}{cc}
B+M^{H} & 0 \\
0 & M^{L}
\end{array}\right]^{-1}\left[\begin{array}{cc}
M^{H} & S^{d D} \\
\left(S^{d D}\right)^{T} & M^{L}
\end{array}\right] .
\end{aligned}
$$

The matrix $\mathrm{PrCo}_{D}$ is not symmetric and will have negative eigenvalues. The definition of the condition number used is $\kappa\left(\operatorname{PrCo}_{D}\right)=\frac{\max _{i}\left\{\backslash \lambda_{i} \mid\right\}}{\min _{i}\left\{\lambda \lambda_{i}\right\}}$. The "qz" algorithm in octave is used.

Table 8.10 and table 8.9 shows very good results. The condition numbers unpreconditioned system shows the importance of the scaling for numerical stability. The condition numbers preconditioned system shows condition numbers that conform with the expectations from chapter $5\left(\lim \left(\kappa\left(P^{-1} B\right)\right) \rightarrow\right.$ 2.66946). It appears we have found a good implementation for the simplified $k-\epsilon$ model with a $H^{2} \times L^{2}$ representation.

Table 8.10: Condition numbers ( $\kappa_{2}$ ) for preconditioned unscaled combined matrices using discontinuous Lagrange-2

| number of <br> elements | $\mathrm{PrCo}_{D}$ <br> condition number | $C o_{D H}$ <br> condition number |
| :---: | :---: | :---: |
| $1=2^{0}$ | 2.6249 | 72.404 |
| $2=2^{1}$ | 2.6279 | 213.85 |
| $4=2^{2}$ | 2.6279 | 731.68 |
| $8=2^{3}$ | 2.6280 | 2660.7 |
| $16=2^{4}$ | 2.6280 | $1.0055 \mathrm{e}+04$ |
| $32=2^{5}$ | 2.6280 | $3.8962 \mathrm{e}+04$ |
| $64=2^{6}$ | 2.6280 | $1.5324 \mathrm{e}+05$ |
| $128=2^{7}$ | 2.6280 | $6.0764 \mathrm{e}+05$ |

## Chapter 9

## System in 2 Dimensions

### 9.1 As a single equation in 2D

$$
\begin{align*}
\epsilon-\Delta k=f_{1}(x) & \text { on } \Omega  \tag{9.1}\\
-\Delta \epsilon+k=f_{2}(x) & \text { on } \Omega \tag{9.2}
\end{align*}
$$

By derivating one of the equations, we reduce this system into a system of one unknown.

$$
\begin{aligned}
-\Delta \epsilon(x)+k(x) & =f_{2}(x) \quad \text { on } \Omega \\
-\sum_{i=1}^{n} \frac{\partial^{2} \epsilon(x)}{\partial x_{i}}+k(x) & =f_{2}(x) \quad \text { on } \Omega \\
\sum_{j=1}^{n} \frac{\partial^{2}}{\partial x_{j}^{2}}\left(\sum_{i=1}^{n} \frac{\partial^{2} \epsilon(x)}{\partial x_{i}^{2}}+k(x)\right) & =\sum_{j=1}^{n} \frac{\partial^{2}}{\partial x_{j}^{2}} f_{2}(x) \quad \text { on } \Omega \\
\sum_{j=1}^{n} \sum_{i=1}^{n} \frac{\partial^{4} \epsilon(x)}{\partial x_{i}^{2} \partial x_{j}^{2}}+\sum_{i=1}^{n} \frac{\partial^{2}}{\partial x_{i}^{2}} k(x) & =\sum_{i=1}^{n} \frac{\partial^{2}}{\partial x_{i}^{2}} f_{2}(x) \quad \text { on } \Omega .
\end{aligned}
$$

This can be added to (9.2)

$$
\begin{aligned}
\sum_{j=1}^{n} \sum_{i=1}^{n} \frac{\partial^{4} \epsilon(x)}{\partial x_{i}^{2} \partial x_{j}^{2}}+\sum_{i=1}^{n} \frac{\partial^{2}}{\partial x_{i}^{2}} k(x)-\sum_{i=1}^{n} \frac{\partial^{2}}{\partial x_{i}^{2}} k(x)+\epsilon(x) & =\sum_{i=1}^{n} \frac{\partial^{2}}{\partial x_{i}^{2}} f_{2}(x)+f_{1}(x) \quad \text { on } \Omega \\
\sum_{j=1}^{n} \sum_{i=1}^{n} \frac{\partial^{4} \epsilon(x)}{\partial x_{i}^{2} \partial x_{j}^{2}}+\epsilon(x) & =\sum_{i=1}^{n} \frac{\partial^{2}}{\partial x_{i}^{2}} f_{2}(x)+f_{1}(x) \quad \text { on } \Omega
\end{aligned}
$$

We multiply with a test function

$$
\begin{array}{r}
\sum_{j=1}^{n} \sum_{i=1}^{n} \frac{\partial^{4} \epsilon(x)}{\partial x_{i}^{2} \partial x_{j}^{2}} v+\epsilon(x) v=\sum_{i=1}^{n} \frac{\partial^{2}}{\partial x_{i}^{2}} f_{2}(x) v+f_{1}(x) v \quad \text { on } \Omega \\
\int_{\Omega} \sum_{j=1}^{n} \sum_{i=1}^{n} \frac{\partial^{4} \epsilon(x)}{\partial x_{i}^{2} \partial x_{j}^{2}} v+\epsilon(x) v \mathrm{~d} x=\int_{\Omega} \sum_{i=1}^{n} \frac{\partial^{2}}{\partial x_{i}^{2}} f_{2}(x) v+f_{1}(x) v \mathrm{~d} x .
\end{array}
$$

Integrate by parts to obtain a weak solution,

$$
\begin{aligned}
& \int_{\partial \Omega} \sum_{j=1}^{n} \sum_{i=1}^{n} \frac{\partial^{3} \epsilon(x)}{\partial x_{i}^{2} \partial x_{j}} v n_{j}- \sum_{j=1}^{n} \sum_{i=1}^{n} \frac{\partial^{2} \epsilon(x)}{\partial x_{i}^{2}} \frac{\partial v}{\partial x_{j}} n_{j} \mathrm{~d} S(x)- \\
& \int_{\Omega} \sum_{j=1}^{n} \sum_{i=1}^{n} \frac{\partial^{2} \epsilon(x)}{\partial x_{i}^{2}} \frac{\partial^{2} v}{\partial x_{j}^{2}}+\epsilon(x) v \mathrm{~d} x=\int_{\Omega} \sum_{i=1}^{n} \frac{\partial^{2}}{\partial x_{i}^{2}} f_{2}(x) v+f_{1}(x) v \mathrm{~d} x .
\end{aligned}
$$

This can be written as
$\int_{\Omega} \Delta \epsilon(x) \Delta v+\epsilon(x) v \mathrm{~d} x=\int_{\Omega} \Delta f_{2}(x) v+f_{1}(x) v \mathrm{~d} x+$ boundery term.
A different representation is

$$
\begin{aligned}
& \int_{\Omega} \sum_{j=1}^{n} \sum_{i=1}^{n} \frac{\partial^{4} \epsilon(x)}{\partial x_{i}^{2} \partial x_{j}^{2}} v+\epsilon(x) v \mathrm{~d} x=\int_{\Omega} \sum_{i=1}^{n} \frac{\partial^{2}}{\partial x_{i}^{2}} f_{2}(x) v+f_{1}(x) v \mathrm{~d} x \\
& \int_{\partial \Omega} \sum_{j=1}^{n} \sum_{i=1}^{n} \frac{\partial^{3} \epsilon(x)}{\partial x_{i}^{2} \partial x_{j}} v n_{j}-\sum_{j=1}^{n} \sum_{i=1}^{n} \frac{\partial^{2} \epsilon(x)}{\partial x_{i} \partial x_{j}} \frac{\partial v}{\partial x_{j}} n_{i} \mathrm{~d} S(x)- \\
& \int_{\Omega} \sum_{j=1}^{n} \sum_{i=1}^{n} \frac{\partial^{2} \epsilon(x)}{\partial x_{i} \partial x_{j}} \frac{\partial^{2} v}{\partial x_{j} \partial x_{i}}+\epsilon(x) v \mathrm{~d} x=\int_{\Omega} \sum_{i=1}^{n} \frac{\partial^{2}}{\partial x_{i}^{2}} f_{2}(x) v+f_{1}(x) v \mathrm{~d} x .
\end{aligned}
$$

This can be written as

$$
\int_{\Omega} \mathbf{D}^{2}(\epsilon): \mathbf{D}^{2}(v)+\epsilon v \mathrm{~d} x=\int_{\Omega} \Delta f_{2} v+f_{1} v \mathrm{~d} x+\text { boundery term. }
$$

$\mathbf{D}^{2}(u)$ is the Hessian matrix. These weak formulations are equal in one dimension. By setting $n=2$ the left hand side of the equations looks like this:

$$
\begin{align*}
\int_{\Omega} u_{x x} v_{x x}+u_{x x} v_{y y}+u_{y y} v_{x x}+u_{y y} v_{y y}+u v \mathrm{~d} x & =\langle u, v\rangle_{(B)}  \tag{9.3}\\
\int_{\Omega} u_{x x} v_{x x}+2 u_{x y} v_{x y}+u_{y y} v_{y y}+u v \mathrm{~d} x & =\langle u, v\rangle_{(H)} \tag{9.4}
\end{align*}
$$

$u v$ is removed to obtain a $H_{0}^{2}$ inner product ((9.6), (9.5))

$$
\begin{array}{r}
\int_{\Omega} u_{x x} v_{x x}+u_{x x} v_{y y}+u_{y y} v_{x x}+u_{y y} v_{y y} \mathrm{~d} x=\left\langle u, v>_{(1)}\right. \\
\int_{\Omega} u_{x x} v_{x x}+2 u_{x y} v_{x y}+u_{y y} v_{y y} \mathrm{~d} x=\left\langle u, v>_{(2)}\right. \tag{9.6}
\end{array}
$$

Table 9.1: functions in the null space of the functionals $<\cdot, \cdot>_{(1)},<\cdot, \cdot>_{(2)}$

|  | $<u, u>_{(1)}$ | $\left\langle u, u>_{(2)}\right.$ |
| :---: | :---: | :---: |
| $u=C$ | 0 | 0 |
| $u=C x$ | 0 | 0 |
| $u=C y$ | 0 | 0 |
| $u=C x y$ | 0 | $2 C^{2}\|\Omega\|$ |

We identify which functions in $H^{2}$ that are part of the null space of the inner product. They are in table 9.1. The null spaces of these inner products are different. To simplify further investigation of the stability of these representations, it is prudent to remove the zero eigenvalues when determining the condition number.

### 9.1.1 Definition Point Spectrum

Let $T$ be a non-invertible bounded linear operator. Then the point spectrum $\sigma_{p}(T)$ consists of all the nonzero eigenvalues of $T$.

### 9.1.2 Definition non-zero discrete eigenvalues

Let $A \in M_{n \times n}$ be a symmetric matrix, with eigenvalues $0<\lambda_{1} \leq \lambda_{2} \leq \cdots \leq$ $\lambda_{n}$. Then an eigenvalue $\lambda_{i}$ is said to be non-zero if and only if

$$
\begin{align*}
& \lambda_{i+1} \quad \text { is non-zero }  \tag{9.7}\\
& \frac{\lambda_{i}}{\lambda_{i+1}}>10^{-7}  \tag{9.8}\\
& \lambda_{i}>10^{-4} h^{4} . \tag{9.9}
\end{align*}
$$

### 9.1.3 Definition non-zero discrete condition number

$\kappa_{p d}(A)=\frac{\max \left|\sigma_{p}(A)\right|}{\min \left|\sigma_{p}(A)\right|}$ with the discrete definition of non-zero eigenvalues.
For the two representations, this works out to removing the 3 and 4 smallest eigenvalues ((9.4) and (9.3) respectively) before computing the condition numbers. Eight different scaling of the basis functions is selected. The $\sqrt{3}$ scaling is added to make all the diagonal elements of a $B$ be exactly the same.

Table 9.2: Condition numbers ( $\kappa_{p d}$ ) of matrices arising from types of scaling for Hermite elements

| number of <br> elements | $B_{1,1,1}^{1}$ <br> cond number | $B_{1, \sqrt{3}, 1}^{1}$ <br> cond number | $B_{1, \sqrt{3}, 2}^{1}$ <br> cond number |
| :---: | :---: | :---: | :---: |
| $1=\left(2^{0}\right)^{2}$ | 2584.6 | 1116.5 | 500.35 |
| $4=\left(2^{1}\right)^{2}$ | 382880 | 130028 | 49000 |
| $16=\left(2^{2}\right)^{2}$ | 51853049 | 17331457 | 6468032 |
| $64=\left(2^{3}\right)^{2}$ | 8252398263 | 2755125937 | 1016105092 |
| $256=\left(2^{4}\right)^{2}$ | $1.5743 \mathrm{e}+12$ | 525437544994 | 192216629691 |
| $1024=\left(2^{5}\right)^{2}$ | $3.4127 \mathrm{e}+14$ | no result | no result |

### 9.1.4 The condition number of a Biharmonic matrix

$$
\begin{aligned}
\left(B_{2, w, p}^{N}\right)_{i, j} & =\int \mathbf{D}^{2}\left(H_{i}^{N}\right): \mathbf{D}^{2}\left(H_{j}^{N}\right) \mathrm{d} x \\
\left(B_{1, w, p}^{N}\right)_{i, j} & =\int \Delta\left(H_{i}^{N}\right) \Delta\left(H_{j}^{N}\right) \mathrm{d} x \\
\phi_{0}(0,0) & =1 \\
\frac{\partial \phi_{1}^{N}(0,0)}{\partial x} & =w N \\
\frac{\partial \phi_{2}^{N}(0,0)}{\partial y} & =w N \\
\frac{\partial^{2} \phi_{3}^{N}(0,0)}{\partial x \partial y} & =(w N)^{p}
\end{aligned}
$$

With this, the tables 9.2, 9.3, 9.4 and 9.5. As we can see in these tables full scaling with $N$ is better than reduced scaling with $N$. Reduced Scaling with $N$ is better than no scaling. Adding the $\sqrt{3}$ scaling improves the condition numbers further. The Hessian inner product ((9.4)) is better than the biharmonic inner product ((9.3)).

### 9.1.5 The condition number of the system as one equation

The $\epsilon-k$ equations is expressed in two different variational forms:

$$
\begin{align*}
\int_{\Omega} \Delta \epsilon(x) \Delta v+\epsilon(x) v \mathrm{~d} x & =\int_{\Omega} \Delta f_{2}(x) v+f_{1}(x) v \mathrm{~d} x+\text { boundery term }  \tag{9.10}\\
\int_{\Omega} \mathbf{D}^{2}(\epsilon): \mathbf{D}^{2}(v)+\epsilon v \mathrm{~d} x & =\int_{\Omega} \Delta f_{2} v+f_{1} v \mathrm{~d} x+\text { boundery term. } \tag{9.11}
\end{align*}
$$

Table 9.3: Condition numbers ( $\kappa_{p d}$ ) of matrices arising from types of scaling for Hermite elements

| number of <br> elements | $B_{2,1,1}^{1}$ <br> cond number | $B_{2, \sqrt{3}, 1}^{1}$ <br> cond number | $B_{2, \sqrt{3}, 2}^{1}$ <br> cond number |
| :---: | :---: | :---: | :---: |
| $1=\left(2^{0}\right)^{2}$ | 2427.2 | 999.36 | 512.07 |
| $4=\left(2^{1}\right)^{2}$ | 54294 | 18351 | 9257.0 |
| $16=\left(2^{2}\right)^{2}$ | 975975 | 325526 | 163189 |
| $64=\left(2^{3}\right)^{2}$ | 16833179 | 5611233 | 2807417 |
| $256=\left(2^{4}\right)^{2}$ | 275872937 | 91957807 | 45986184 |
| $1024=\left(2^{5}\right)^{2}$ | 4442364060 | 1480788243 | 740423344 |

Table 9.4: Condition numbers ( $\kappa_{p d}$ ) of matrices arising from types of scaling for Hermite elements

| number of <br> elements | $B_{1,1,1}^{N}$ <br> cond number | $B_{1, \sqrt{3}, 1}^{N}$ <br> cond, <br> number | $B_{1,1,2}^{N}$ <br> cond number | $B_{1, \sqrt{3}, 2}^{N}$ <br> cond number |
| :---: | :---: | :---: | :---: | :---: |
| $1=\left(2^{0}\right)^{2}$ | 2584.6 | 1116.53 | 2584.6 | 500.35 |
| $4=\left(2^{1}\right)^{2}$ | 98437 | 35381 | 29571 | 5620.4 |
| $16=\left(2^{2}\right)^{2}$ | 3307999 | 1153197 | 433374 | 141023 |
| $64=\left(2^{3}\right)^{2}$ | 135349903 | 49499784 | 14939717 | 8865144 |
| $256=\left(2^{4}\right)^{2}$ | 7154662894 | 3059553590 | 1324295802 | 1112003947 |
| $1024=\left(2^{5}\right)^{2}$ | no result | 308279719496 | 212530478242 | 202342715830 |

Table 9.5: Condition numbers ( $\kappa_{p d}$ ) of matrices arising from types of scaling for Hermite elements

| number of <br> elements | $B_{2,1,1}^{N}$ <br> cond number | $B_{2, \sqrt{3}, 1}^{N}$ <br> cond number | $B_{2,1,2}^{N}$ <br> cond number | $B_{2, \sqrt{3}, 2}^{N}$ <br> cond number |
| :---: | :---: | :---: | :---: | :---: |
| $1=\left(2^{0}\right)^{2}$ | 2427.2 | 999.36 | 2427.2 | 347.73 |
| $4=\left(2^{1}\right)^{2}$ | 13860 | 4890.1 | 3547.3 | 451.39 |
| $16=\left(2^{2}\right)^{2}$ | 61292 | 20670 | 3982.1 | 749.48 |
| $64=\left(2^{3}\right)^{2}$ | 263290 | 87986 | 6953.1 | 5801.09 |
| $256=\left(2^{4}\right)^{2}$ | 1077887 | 359511 | 68489 | 65081 |
| $1024=\left(2^{5}\right)^{2}$ | 4338501 | 1446384 | 883276 | 871224 |

Table 9.6: Condition numbers ( $\kappa_{1}$ ) of matrices arising from types of scaling for Hermite elements

| number of <br> elements | $H_{1,1,1}^{1}$ <br> cond number | $H_{1, \sqrt{3}, 1}^{1}$ <br> cond number | $H_{1, \sqrt{3}, 2}^{1}$ <br> cond number |
| :---: | :---: | :---: | :---: |
| $1=\left(2^{0}\right)^{2}$ | 274174 | 119207 | 99570 |
| $4=\left(2^{1}\right)^{2}$ | 11824630 | 4004690 | 1866975 |
| $16=\left(2^{2}\right)^{2}$ | 128049428 | 42776856 | 18323535 |
| $64=\left(2^{3}\right)^{2}$ | 1631605949 | 544222077 | 198267421 |
| $256=\left(2^{4}\right)^{2}$ | 72057123133 | 24024240804 | 8315282872 |
| $1024=\left(2^{5}\right)^{2}$ | $3.4379 \mathrm{e}+12$ | $1.1460 \mathrm{e}+12$ | 389233623024 |

Table 9.7: Condition numbers ( $\kappa_{1}$ ) of matrices arising from types of scaling for Hermite elements

| number of <br> elements | $H_{2,1,1}^{1}$ <br> cond number | $H_{2, \sqrt{3}, 1}^{1}$ <br> cond number | $H_{2, \sqrt{3}, 2}^{1}$ <br> cond number |
| :---: | :---: | :---: | :---: |
| $1=\left(2^{0}\right)^{2}$ | 2432.38 | 1001.0 | 921.30 |
| $4=\left(2^{1}\right)^{2}$ | 54306 | 18355 | 13779 |
| $16=\left(2^{2}\right)^{2}$ | 975967 | 325523 | 181955 |
| $64=\left(2^{3}\right)^{2}$ | 16833145 | 5611222 | 2807938 |
| $256=\left(2^{4}\right)^{2}$ | 275872874 | 91957787 | 45986398 |
| $1024=\left(2^{5}\right)^{2}$ | 4442364155 | 1480788307 | 740423427 |

The tables 9.6, $9.7,9.8$ and 9.9 provides the condition numbers for the left hand matrices as provided by FEM on these equations. $H_{1, .,}=B_{1, .,}+M_{:,,}$represents the equation (9.10) and $H_{2,,,}=B_{2,,, \prime}+M_{i,}$, represents the equation (9.11). They are the sums of the biharmonic and the mass matrix, with the same scaling of the basis functions.

As we can see in these tables full scaling with $N$ is better than reduced scaling with $N$. Reduced Scaling with $N$ is better than no scaling. Adding the $\sqrt{3}$ scaling improves the condition numbers further. The Hessian inner product ((9.4)) is better than the biharmonic inner product ((9.3)).

### 9.2 Representations as a Linear system in 2D

In this section the $k-\epsilon$ model is implemented with Lagrange- 3 and Hermite elements. The system is an $H^{2} \times L^{2}$ representation and is defined in (9.12).

Table 9.8: Condition numbers ( $\kappa_{1}$ ) of matrices arising from types of scaling for Hermite elements

| number of <br> elements | $H_{1,1,1}^{N}$ <br> cond number | $H_{1, \sqrt{3}, 1}^{N}$ <br> cond number | $H_{1, \sqrt{3}, 2}^{N}$ <br> cond number |
| :---: | :---: | :---: | :---: |
| $1=\left(2^{0}\right)^{2}$ | 274174 | 119207 | 99570 |
| $4=\left(2^{1}\right)^{2}$ | 3027583 | 1076377 | 286564 |
| $16=\left(2^{2}\right)^{2}$ | 8136984 | 2813752 | 530594 |
| $64=\left(2^{3}\right)^{2}$ | 26019491 | 9038317 | 1856347 |
| $256=\left(2^{4}\right)^{2}$ | 289291368 | 101728884 | 13539678 |
| $1024=\left(2^{5}\right)^{2}$ | 3451034726 | 1214028856 | 157858076 |

Table 9.9: Condition numbers $\left(\kappa_{1}\right)$ of matrices arising from types of scaling for Hermite elements

| number of <br> elements | $H_{2,1,1}^{N}$ <br> cond number | $H_{2, \sqrt{3}, 1}^{N}$ <br> cond number | $H_{2, \sqrt{3}, 2}^{N}$ <br> cond number |
| :---: | :---: | :---: | :---: |
| $1=\left(2^{0}\right)^{2}$ | 2432.4 | 1001.0 | 922.27 |
| $4=\left(2^{1}\right)^{2}$ | 13862 | 7303.4 | 7295.6 |
| $16=\left(2^{2}\right)^{2}$ | 74437 | 58516 | 58522 |
| $64=\left(2^{3}\right)^{2}$ | 669028 | 612352 | 612364 |
| $256=\left(2^{4}\right)^{2}$ | 7969676 | 7761649 | 7761663 |
| $1024=\left(2^{5}\right)^{2}$ | 110812171 | 110021915 | 110021930 |

Table 9.10: Condition numbers $\left(\kappa_{1}\right)$ of matrices arising from types of scaling for Hermite elements when solving the equations in 2-D

| number of <br> elements | $\mathrm{Co}_{1,1}^{1}$ <br> cond number | $\mathrm{Co}_{3,1}^{1}$ <br> cond number | $\mathrm{Co}_{3,2}^{1}$ <br> cond number |
| :---: | :---: | :---: | :---: |
| $1=\left(2^{0}\right)^{2}$ | $5.3102 \mathrm{e}+06$ | $2.1878 \mathrm{e}+06$ | $7.6073 \mathrm{e}+05$ |
| $4=\left(2^{1}\right)^{2}$ | $9.5665 \mathrm{e}+05$ | $3.2232 \mathrm{e}+05$ | $1.1365 \mathrm{e}+05$ |
| $16=\left(2^{2}\right)^{2}$ | $5.2881 \mathrm{e}+07$ | $1.7635 \mathrm{e}+07$ | $5.9285 \mathrm{e}+06$ |
| $64=\left(2^{3}\right)^{2}$ | $1.3235 \mathrm{e}+09$ | $4.4118 \mathrm{e}+08$ | $1.4781 \mathrm{e}+08$ |

Table 9.11: Condition numbers ( $\kappa_{1}$ ) of matrices arising from types of scaling for Hermite elements when solving the equations in 2-D

| number of <br> elements | $C o_{1,1}^{N}$ <br> cond number | $C o_{3,1}^{N}$ <br> cond number | $C o_{1,2}^{N}$ <br> cond number | $C o_{3,2}^{N}$ <br> cond number |
| :---: | :---: | :---: | :---: | :---: |
| $1=\left(2^{0}\right)^{2}$ | $5.3102 \mathrm{e}+06$ | $2.1878 \mathrm{e}+06$ | $5.3102 \mathrm{e}+06$ | $7.6073 \mathrm{e}+05$ |
| $4=\left(2^{1}\right)^{2}$ | $2.4307 \mathrm{e}+05$ | $8.4899 \mathrm{e}+04$ | $6.6080 \mathrm{e}+04$ | 9516.7 |
| $16=\left(2^{2}\right)^{2}$ | $3.3166 \mathrm{e}+06$ | $1.1155 \mathrm{e}+06$ | $2.2289 \mathrm{e}+05$ | $3.4470 \mathrm{e}+04$ |
| $64=\left(2^{3}\right)^{2}$ | $2.0696 \mathrm{e}+07$ | $6.9126 \mathrm{e}+06$ | $3.8493 \mathrm{e}+05$ | $6.8329 \mathrm{e}+04$ |

The preconditioned system is (9.13).

$$
\begin{align*}
\operatorname{Co}_{k, n} & =\left[\begin{array}{cc}
M^{H} & S^{D} \\
\left(S^{D}\right)^{T} & M^{L}
\end{array}\right]  \tag{9.12}\\
\operatorname{PrCo}_{v, k, n} & =\left[\begin{array}{cc}
B_{v, k, n}+M^{H} & 0 \\
0 & M^{L}
\end{array}\right]^{-1}\left[\begin{array}{cc}
M^{H} & S^{D} \\
\left(S^{D}\right)^{T} & M^{L}
\end{array}\right] \tag{9.13}
\end{align*}
$$

where $v$ represents the type of biharmonic matrix used (as in the previous section), $k$ is the weighting of the Hermite basis functions and $n$ is the power of the weighting of the double derivative evaluation. If $k=3$ and $n=2$ all the values along the diagonal of the Hermite stiffness matrix (without the mass matrix) will be of the same size. The tables for the scaled matrices of (9.12) are given in the tables 9.10 and 9.11 . They confirm the importance of scaling the Hermite elements.

The tables for the scaled matrices of (9.13) are given in the tables 9.12 and 9.13. The condition numbers in the preconditioned tables increases. This implies that this implementation does not work. As is evident in 9.13 and 9.12, the scaling does not change the condition number, this is to be expected. Let $T$ be a scaling matrix. $T$ will be a diagonal matrix, with diagonal values corre-

Table 9.12: Condition numbers $\left(\kappa_{2}\right)$ of matrices arising from types of scaling for Hermite elements when solving the equations in 2-D

| number of <br> elements | $\operatorname{Pr} C o_{1,1,1}^{N}$ <br> cond number | $\operatorname{PrCo}$ <br> cond number | $\operatorname{Pr} C o_{1,3,2}^{N}$ <br> cond number | $\operatorname{Pr} C o_{1,3,2}^{N}$ <br> cond number |
| :---: | :---: | :---: | :---: | :---: |
| $1=\left(2^{0}\right)^{2}$ | $2.7184 \mathrm{e}+04$ | $2.7184 \mathrm{e}+04$ | $2.7184 \mathrm{e}+04$ | $2.7184 \mathrm{e}+04$ |
| $4=\left(2^{1}\right)^{2}$ | 4513.0 | 4513.0 | 4513.0 | 4513.0 |
| $16=\left(2^{2}\right)^{2}$ | $2.5914 \mathrm{e}+05$ | $2.5914 \mathrm{e}+05$ | $2.5914 \mathrm{e}+05$ | $2.5914 \mathrm{e}+05$ |
| $64=\left(2^{3}\right)^{2}$ | $1.3422 \mathrm{e}+06$ | $1.3422 \mathrm{e}+06$ | $1.3422 \mathrm{e}+06$ | $1.3422 \mathrm{e}+06$ |

Table 9.13: Condition numbers $\left(\kappa_{2}\right)$ of matrices arising from types of scaling for Hermite elements when solving the equations in 2-D

| number of <br> elements | $\operatorname{PrCo}$ <br> cond number <br> con | $\operatorname{PrCo}_{2,3,1}^{N}$ <br> cond number | $\operatorname{PrCo}_{2,1,2}^{N}$ <br> cond number | $\operatorname{PrCo}$ <br> cond number |
| ---: | :---: | :---: | :---: | :---: |
| $1=\left(2^{0}\right)^{2}$ | $1.4648 \mathrm{e}+05$ | $1.4648 \mathrm{e}+05$ | $1.4648 \mathrm{e}+05$ | $1.4648 \mathrm{e}+05$ |
| $4=\left(2^{1}\right)^{2}$ | $9.2140 \mathrm{e}+04$ | $9.2140 \mathrm{e}+04$ | $9.2140 \mathrm{e}+04$ | $9.2140 \mathrm{e}+04$ |
| $16=\left(2^{2}\right)^{2}$ | $1.0307 \mathrm{e}+06$ | $1.0307 \mathrm{e}+06$ | $1.0307 \mathrm{e}+06$ | $1.0307 \mathrm{e}+06$ |
| $64=\left(2^{3}\right)^{2}$ | $1.0034 \mathrm{e}+07$ | $1.0034 \mathrm{e}+07$ | $3.8493 \mathrm{e}+05$ | $1.0034 \mathrm{e}+07$ |

sponding to the square root of the scaling of the basis function corresponding with that row and column. The elements on $T^{-1}$ will be one divided by the corresponding element on $T$. Thus the preconditioned problem will be;

$$
\begin{aligned}
(T \operatorname{Pr} T)^{-1}(T C o T) & =T^{-1} \operatorname{Pr}^{-1} T^{-1} T \operatorname{CoT} \\
& =T^{-1} \operatorname{Pr}^{-1} \operatorname{CoT}
\end{aligned}
$$

This gives no benefit when it comes to the condition number. The diagonal elements of the unscaled and the scaled preconditioned problems are the same, the only thing that changes is the off diagonal elements. Since the matrices are diagonally dominant, the change between different scalings is dramatic. However including the scaling increases the stability when calculating the preconditioner.

## Chapter 10

## Conclusions

In this thesis a system of elliptical partial differential equations called the simplified $k-\epsilon$ model has been examined. Its weak form is well defined analytically. The boundary conditions we can apply are dependent on which solution space we choose. For all $(\epsilon, k) \in H^{1}(\Omega) \times H^{1}(\Omega)$ the trace $\|(\epsilon, k)\|_{L^{2}(\partial \Omega) \times L^{2}(\partial \Omega)}$ is bounded by $\|(\epsilon, k)\|_{H^{1}(\Omega) \times H^{1}(\Omega)}$. For all $(\epsilon, k) \in H^{2}(\Omega) \times L^{2}(\Omega)$ the trace of $\epsilon\left(\left\|\frac{\partial \epsilon}{\partial n}\right\|_{L^{2}(\partial \Omega)}+\|\epsilon\|_{L^{2}(\partial \Omega)}\right)$ is bounded by $\|\epsilon\|_{H^{2}(\Omega)}$. Therefore these formulations were used with their respective boundary conditions. $(\epsilon, k) \in H^{1}(\Omega) \times$ $H^{1}(\Omega)$ is used with one boundary condition on each variable, while $(\epsilon, k) \in$ $H^{2}(\Omega) \times L^{2}(\Omega)$ is used for two boundary conditions on $\epsilon$.

The formulations are well defined analytically and Riesz mappings has been used as preconditioners for each formulation.

In this master thesis the $H^{2}$ conforming Hermite elements has been successfully implemented. The identified issues pertaining to mesh refinement has been fixed by scaling. To scale a Hermite element we multiply the basis functions representative of the derivates based on the size of the element and how many derivatives it includes. The scaling of the Hermite elements allowed for numerical experiments. Various combinations of elements, scaling and weak formulations were tested. Many of the experiments were implemented both in one and two dimensions.

The results of the $H^{2} \times L^{2}$ formulations, when continuous Lagrange (CG) is used, are not stable. The continuous Lagrange paired with Hermite elements is $H^{2} \times H^{1}$ conforming. Because $H^{1} \subset L^{2}$ the combination of Hermite and CG is also $H^{2} \times L^{2}$ conforming. The idea was therefore that Hermite-CG formulation would work. This thesis have showed by numerical experiments that this idea is flawed, and the Hermite-CG formulation is not stable.

The results of the $H^{2} \times L^{2}$ formulations, when discontinuous Lagrange (dCG) is used, are very uplifting. The condition numbers are very close to the condition numbers of the analytical calculations in (5.28). This implies that
the preconditioned simplified $k-\epsilon$ system is successfully implemented and that

$$
\begin{equation*}
B \sim S^{d D} \cdot\left(S^{d D}\right)^{T} \tag{10.1}
\end{equation*}
$$

The scaling of the basis functions for Hermite with dCG simulations shows its usefulness when comparing unpreconditioned tables. It is reasonable to suspect that scaling allows higher degree of mesh refinement for a fixed accuracy demand, even when a system is preconditioned.

As expected from [1] the $H^{1} \times H^{1}$ formulation works. It gave quite good results as the mathematical formulation suggest. The condition numbers are stable as the tables of section 8.1 .2 shows. However boundary conditions for both $\epsilon$ and $k$ are required. It does not work very well if Hermite combined with CG elements are used.

In 2-dimensions with Hermite and CG elements the results is similar to the 1-dimensional case. Here, scaling shows its usefulness. We also learned that (9.6) was more stable than (9.5).

It is reasonable to suspect that the lessons learned about CG vs dCG in the 1-dimensional case also applies to the 2-dimensional cases. We can suspect this because the Hermite-CG system gave similar types of results in 1-dimensions as it did in 2-dimensions, and because the systems that did not work in 1 dimensions used elements that conformed with higher order Sobolev spaces than was strictly needed (see section 8.1.1 and 8.1.3).

Since the implementation of the preconditioned simplified $k-\epsilon$ model was successful the next logical step is to implement the full $k-\epsilon$ model with this new set of boundary conditions.

It is important to note that the $H^{2} \times L^{2}$ formulations are symmetric. Therefore two new possible sets of boundary conditions can be used

$$
\begin{gather*}
B A_{k}=\left\{\begin{array}{cc}
k=g_{1} & \text { on } \partial \Omega \\
\frac{\partial k}{\partial n}=g_{2} & \text { on } \partial \Omega
\end{array}\right.  \tag{10.2}\\
B A_{\epsilon}=\left\{\begin{array}{cc}
\epsilon=g_{1} & \text { on } \partial \Omega \\
\frac{\partial \epsilon}{\partial n}=g_{2} & \text { on } \partial \Omega
\end{array}\right. \tag{10.3}
\end{gather*}
$$

It is useful if the full $k-\epsilon$ model can be successfully implemented with the new set of boundary conditions. The results of this thesis open up for numerical approximations of the $k-\epsilon$ model in situations where it is possible to obtain accurate information about one of the variables, but not the other. This can be very useful in engineering where the $k-\epsilon$ models are among of the most commonly used turbulence models.

So from this thesis we should be able to conclude that the $H^{2} \times L^{2}$ formulations work. It should be implemented with Hermite and dCG elements. The $H^{1} \times H^{1}$ formulation worked, but only if two CG elements are used. It is important to not chose elements that conform with Sobolev spaces of higher order than the particular one we are interested in. The Hermite elements should be scaled to produce a more stable approximation.

## Chapter 11

## Appendix

### 11.1 Definitions

Definition of $L^{2}(\Omega)$ space $[10] 1 L^{2}(\Omega)$ is the functions space of square integrable functions on $\Omega$. $u$ is said to be in $L^{2}(\Omega)$ if

$$
\begin{equation*}
\|u\|_{L^{2}(\Omega)}=\left(\int_{\Omega}|u|^{2} \mathrm{~d} x\right)^{\frac{1}{2}}<\infty . \tag{11.1}
\end{equation*}
$$

Definition of $H^{n}(\Omega)$ space [10] $2 H^{n}(\Omega)$ is the Sobolev space containing derivatives of power $n$ and and $L^{2}$ type norm. The norm is defined as

$$
\begin{equation*}
\|u\|_{H^{n}(\Omega)}=\left(\int_{\Omega} \sum_{|\alpha| \leq n}\left|D^{\alpha} u\right|^{2} \mathrm{~d} x\right)^{\frac{1}{2}}<\infty \tag{11.2}
\end{equation*}
$$

It has the following inner product:

$$
\begin{equation*}
(u, v)_{H^{n}(\Omega)}=\int_{\Omega} \sum_{|\alpha| \leq n} D^{\alpha} u D^{\alpha} v \mathrm{~d} x . \tag{11.3}
\end{equation*}
$$

Where $\alpha=\left(\alpha_{1}, \alpha_{2}, \ldots, \alpha_{d}\right)$ is a multi index, and $|\alpha|=\alpha_{1}+\alpha_{2}+\cdots+\alpha_{d}$. While $d$ is the number of dimensions and

$$
D^{\alpha} u=\frac{\partial^{\alpha_{1}} u}{\partial x_{i}^{\text {alpha }}}+\frac{\partial^{\alpha_{2}} u}{\partial x_{2}^{a l p h a_{2}}}+\cdots+\frac{\partial^{\alpha_{d}} u}{\partial x_{d}^{\text {alpha }}}
$$

$u \in H^{n} \leftrightarrow\|u\|_{H^{n}}<\infty$.

Definition of $H_{0}^{1}(\Omega)$ space [10] $3 H_{0}^{1}(\Omega)$ is the Sobolev space containing derivatives of power $n$ and and $L^{2}$ type norm. the norm is defined as:

$$
\begin{equation*}
\|u\|_{H_{0}^{1}(\Omega)}=\left(\int_{\Omega}|\nabla u|^{2} \mathrm{~d} x\right)^{\frac{1}{2}}<\infty . \tag{11.4}
\end{equation*}
$$

It has the following inner product:

$$
\begin{equation*}
(u, v)_{H_{0}^{1}(\Omega)}=\int_{\Omega} \nabla u \cdot \nabla v \mathrm{~d} x . \tag{11.5}
\end{equation*}
$$

Definition of $H^{1}(\Omega)$ space [10] $4 H^{1}(\Omega)$ is the Sobolev space containing derivatives of power $n$ and and $L^{2}(\Omega)$ type norm. The norm is defined as:

$$
\begin{equation*}
\|u\|_{H^{1}(\Omega)}=\left(\int_{\Omega}|\nabla u|^{2}+|u|^{2} \mathrm{~d} x\right)^{\frac{1}{2}}<\infty . \tag{11.6}
\end{equation*}
$$

It has the following inner product:

$$
\begin{equation*}
(u, v)_{H^{1}(\Omega)}=\int_{\Omega} \nabla u \cdot \nabla v+u v \mathrm{~d} x \tag{11.7}
\end{equation*}
$$

Note that $L^{2}(\Omega)=H^{0}(\Omega)$.
$C_{c}^{\infty}(\Omega)$ is the space of all infinitely continuous differentiable functions with compact support.
$H_{0}^{n}(\Omega)=H^{n}(\Omega) \cap C_{c}^{n-1}(\bar{\Omega})$.

### 11.2 Theorems and Functional analysis

Youngs inequality [10] 5

$$
\begin{equation*}
a b \leq \frac{a^{2}}{2}+\frac{b^{2}}{2} \tag{11.8}
\end{equation*}
$$

Minkowskis inequality [10] 6

$$
\begin{equation*}
\|f+g\|_{L^{2}(\Omega)} \leq\|f\|_{L^{2}(\Omega)}+\|g\|_{L^{2}(\Omega)} \tag{11.9}
\end{equation*}
$$

Definition of a Bounded Linear Functional [10] 7 Let $X$ be a Banach Space. A bounded linear operator $u^{*}: X \rightarrow \mathbb{R}$ is called a bounded linear functional on $X$.

We write $X^{*}$ to denote the collection of all bounded linear functionals on $X . X^{*}$ is the dual space of $X$.

Let $H$ be a Hilbert space with inner product $(\cdot, \cdot)$ and $u \in H, u^{*} \in H^{*}$ (so $\left.u^{*}(u) \in \mathbb{R}\right)$.

Riesz Representation Theorem [10] $8 H^{*}$ can canonically be identified with $H$. For each $u^{*} \in H^{*}$ there exists a unique element $u \in H$ such that $u^{*}(v)=(u, v)$ for all $v \in H . u^{*} \mapsto u$ is a linear isomorphism from $H^{*}$ onto $H$.

Let $\mathcal{L}\left(H^{*}, H\right)$ be the set of all linear functionals from $H^{*}$ to $H$
Definition of Riesz map $9 R \in \mathcal{L}\left(H^{*}, H\right)$ is the Riesz map if: Given $u^{*} \in H^{*}$ then $u^{*}(v)=\left(R\left(u^{*}\right), v\right)_{H}$.

Independent Eigenvector Theorem 1 If $A$ is an real $N \times N$ matrix with $\{\lambda\}_{i=1}^{N}$ distinct real nonzero eigenvalues. Then any set of $\left\{e_{i} \neq 0\right\}_{i=1}^{N}$ eigenvectors with $A e_{i}=\lambda_{i} e_{i}$ form a basis for $\mathbb{R}^{N}$

## Proof

It is sufficient to prove that $\left\{e_{i}\right\}_{i=1}^{N}$ is a linearly independent set. If $\left\{e_{i}\right\}_{i=1}^{N}$ is dependent, then there must exist at least one minimal subset of $2 \leq j \leq N$ linearly dependent vectors $\left\{v_{i}\right\}_{i=1}^{j}$, with the eigenvalues $\left\{\rho_{i}\right\}_{i=1}^{j}$. Since the eigenvectors are dependent there must exist as set $\left\{a_{i}\right\}_{i=1}^{j}$ with at least two non-zero constants such that,

$$
a_{1} v_{1}+a_{2} v_{2}+\cdots+a_{j} v_{j}=0
$$

We can assume that $a_{j} \neq 0$ and define $b_{k}=-a_{k} / a_{j}$ and get

$$
\begin{equation*}
v_{j}=b_{1} v_{1}+v_{2} v_{2}+\cdots+b_{j-1} v_{j-1} . \tag{11.10}
\end{equation*}
$$

Multiply (11.10) with $A$ and get

$$
\begin{equation*}
\rho_{j} v_{j}=\rho_{1} b_{1} v_{1}+\rho_{2} v_{2} v_{2}+\cdots+\rho_{j-1} b_{j-1} v_{j-1} \tag{11.11}
\end{equation*}
$$

Multiply (11.10) with $\rho_{j}$ and get

$$
\begin{equation*}
\rho_{j} v_{j}=\rho_{j} b_{1} v_{1}+\rho_{j} v_{2} v_{2}+\cdots+\rho_{j} b_{j-1} v_{j-1} . \tag{11.12}
\end{equation*}
$$

Subtract (11.11) from (11.12) and get

$$
\begin{equation*}
0=\left(\rho_{j}-\rho_{1}\right) b_{1} v_{1}+\left(\rho_{j}-\rho_{2}\right) v_{2} v_{2}+\cdots+\left(\rho_{j}-\rho_{j-1} b_{j-1} v_{j-1}\right. \tag{11.13}
\end{equation*}
$$

Since at least one of the $b_{k}$ is non-zero, the set $\left\{v_{i}\right\}_{i=1}^{j-1}$ is linearly dependent, and we have a contradiction.

## Matrixes for 1D

$$
\begin{aligned}
B_{i, j} & =\int \Delta H_{i} \Delta H_{j} \mathrm{~d} x \\
A_{i, j}^{H} & =\int D H_{i} \cdot D H_{j} \mathrm{~d} x \\
A_{i, j}^{D} & =\int-H_{i} \Delta H_{j} \mathrm{~d} x \\
S_{i, j}^{D} & =\int-L_{i} \Delta H_{j} \mathrm{~d} x \\
S_{i, j}^{S} & =\int \nabla L_{i} \cdot \nabla H_{j} \mathrm{~d} x \\
A_{i, j}^{L} & =\int \nabla L_{i} \cdot \nabla L_{j} \mathrm{~d} x \\
M_{i, j}^{S} & =\int H_{i} L_{j} \mathrm{~d} x \\
M_{i, j}^{L} & =\int L_{i} L_{j} \mathrm{~d} x \\
M_{i, j}^{H} & =\int H_{i} H_{j} \mathrm{~d} x
\end{aligned}
$$

## Matrixes for 2D

$$
\begin{aligned}
& \left(B_{2}^{N}\right)_{i, j}=\int \mathbf{D}^{2}\left(H_{i}^{N}\right): \mathbf{D}^{2}\left(H_{j}^{N}\right) \mathrm{d} x \\
& \left(B_{1}^{N}\right)_{i, j}=\int \Delta\left(H_{i}^{N}\right) \Delta\left(H_{j}^{N}\right) \mathrm{d} x \\
& A_{i, j}^{H}=\int D H_{i} \cdot D H_{j} \mathrm{~d} x \\
& A_{i, j}^{D}=\int-H_{i} \Delta H_{j} \mathrm{~d} x \\
& S_{i, j}^{D}=\int-L_{i} \Delta H_{j} \mathrm{~d} x \\
& S_{i, j}^{S}=\int \nabla L_{i} \cdot \nabla H_{j} \mathrm{~d} x \\
& A_{i, j}^{L}=\int \nabla L_{i} \cdot \nabla L_{j} \mathrm{~d} x \\
& M_{i, j}^{S}=\int H_{i} L_{j} \mathrm{~d} x \\
& M_{i, j}^{L}=\int L_{i} L_{j} \mathrm{~d} x \\
& M_{i, j}^{H}=\int H_{i} H_{j} \mathrm{~d} x
\end{aligned}
$$

Where

$$
\begin{aligned}
\mathbf{D}^{2} u & =\sum_{\substack{i=1 \\
j=i}}^{d} \frac{\partial^{2} u}{\partial x_{i} \partial x_{j}} \\
\Delta u & =\sum_{i=1}^{d} \frac{\partial^{2} u}{\left(\partial x_{i}\right)^{2}}
\end{aligned}
$$

and $d$ is the number of dimensions.

### 11.3 Boundary condition table

Table 11.1: Table of different boundary conditions

| $(\epsilon, k) \in H^{2}(\Omega) \times L^{2}(\Omega)$ |  | $(\epsilon, k) \in H^{1}(\Omega) \times H^{1}(\Omega)$ |  |  |
| :---: | :---: | :---: | :---: | :---: |
| $g_{1}=$ | $g_{2}=$ | $g_{1}=$ | $g_{2}=$ |  |
| $\frac{\partial \epsilon}{\partial n}$ | $\epsilon$ | $\epsilon$ | $k$ |  |
| $(\epsilon, k) \in H^{2}(\bar{\Omega}) \times L^{2}(\Omega)$ | $(\epsilon, k) \in H^{1}(\bar{\Omega}) \times H^{1}(\Omega)$ |  |  |  |
| $\frac{\partial^{2} \epsilon}{(\partial n)^{2}}$ |  | $\epsilon$ | $\frac{\partial \epsilon}{\partial n}$ | $\epsilon$ |
| $\frac{\partial^{2} \epsilon}{\partial n)^{2}}$ | $\frac{\partial \epsilon}{\partial n}$ | $\frac{\partial \epsilon}{\partial n}$ | $k$ |  |
| $(\epsilon, k) \in H^{2}(\Omega) \times L^{2}(\bar{\Omega})$ | $(\epsilon, k) \in H^{1}(\Omega) \times H^{1}(\bar{\Omega})$ |  |  |  |
| $\epsilon$ | $k$ | $\frac{\partial k}{\partial n}$ | $k$ |  |
| $\frac{\partial \epsilon}{\partial n}$ |  | $k$ | $\epsilon$ | $\frac{\partial k}{\partial n}$ |
| $(\epsilon, k) \in H^{2}(\bar{\Omega}) \times L^{2}(\bar{\Omega})$ | $(\epsilon, k) \in H^{1}(\bar{\Omega}) \times H^{1}(\bar{\Omega})$ |  |  |  |
| $\frac{\partial^{2} \epsilon}{(\partial n)^{2}}$ | $k$ | $\frac{\partial \epsilon}{\partial n}$ |  | $\frac{\partial k}{\partial n}$ |

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[^0]:    ${ }^{1}$ Notation $\bar{\Omega}=\Omega \cup \partial \Omega$ is used.

[^1]:    ${ }^{1}$ see appendix

[^2]:    ${ }^{1} \overline{f(v)}$ is the complex conjugate of $f(v)$

