# CLIFFORD ANALYSIS AND ITS APLICATIONS IN MATHEMATICAL PHYSICS 

Wolfgang Sprössig<br>Freiberg University of Mining and Technology<br>Faculty of Mathematics and Informatics<br>Agricola-Strasse 1<br>09596 Freiberg<br>Germany

## 1 On hypercomplex numbers

The really great success of the introduction of complex numbers makes it natural to look for generalizations to higher dimensions. The true reason is founded in the possibility of its geometric interpretation. So the consideration of so-called "geometric algebras" with their elements the "hypercomplex" numbers shifts the geometric-algebraic observation techniques to higher dimensions.

### 1.1 Complex numbers

We consider in the vector space $\mathbb{R}^{2}$ pairs of real numbers $z=(x, y)$. Any such pair characterize a free vector in this plane. Two vectors $z_{1}$ and $z_{2}$ are equal if and only if $x_{1}=x_{2}$ and $y_{1}=y_{2}$ where $z_{i}=\left(x_{i}, y_{i}\right)(i=1,2)$. Vectors which denote points lying symmetric to the real axis $x$ are called conjugated, i.e. $\bar{z}:=(x,-y)$ is conjugated to $z$. The set of all such pairs is denoted by $\mathbb{C}$. Further, we induce in $\mathbb{C}$ the addition and multiplication with real numbers from the vector space $R^{2}$. Now any element from $\mathbb{C}$ can be given in the so-called cartesian form

$$
z=x \mathbf{1}+y \mathbf{i}=: x+\mathbf{i} y
$$

where $\mathbf{1}=(1,0)$ and $\mathbf{i}=(0,1)$. These two special pairs are called unit vectors of the corresponding axis $x$ and $y$. Let $z_{i}=x_{i}+\mathbf{i} y_{i} \in \mathbb{C}(i=1,2)$. We define a scalar product and a vector product as follows:

$$
\begin{aligned}
& z_{1} \cdot z_{2}:=x_{1} x_{2}+y_{1} y_{2} \\
& {\left[\dot{z}_{1}, z_{2}\right]:=x_{1} y_{2}-x_{2} y_{1}}
\end{aligned}
$$

By definition we obtain $\mathbf{i} \cdot \mathbf{i}=-1$ and $[\mathbf{i}, \mathbf{i}]=0$. Each of these products is not suited to generate a commutative field. In the following way we can define a product which transforms $\mathbb{C}$ in a commutative field. We have only to set

$$
z_{1} z_{2}=x_{1} x_{2}-y_{1} y_{2}+\mathbf{i}\left(x_{1} y_{2}+y_{1} x_{2}\right)=\overline{z_{1}} \cdot z_{2}+\left[\overline{z_{1}}, z_{2}\right] .
$$

$\mathbb{C}$ is now called field of complex numbers and the elements $z=x+\mathbf{i} y$ is called complex number. Furthermore, it is called $x=\operatorname{Re} z$ real part and $y=\operatorname{Im} z$ imaginary part of the complex number $z$. Complex numbers with zero real part are called imaginary numbers. From complex numbers with zero imaginary part we reobtain our traditionally real numbers. Because of multiplication and division of such complex numbers leads to very complicated expressions it is necessary to represent complex numbers in the polar form :

$$
z=r(\cos \phi+\mathbf{i} \sin \phi)
$$

Each complex number $z \neq 0$ is uniquely representable by the modulus $r=|z|$ of the complex number $z$ and its argument of $\phi:=\operatorname{Arg} z(\bmod 2 \pi)$. It is useful to introduce the abbreviation:

$$
\cos \phi+\mathbf{i} \sin \phi=: e^{\mathbf{i} \phi}
$$

Thus the complex number $z$ has the representation

$$
z=r e^{\mathrm{i} \phi} .
$$

By the help of trigonometric relations and the definition of the complex multiplication we obtain the following rule:

$$
r_{1} e^{\mathrm{i} \phi_{1}} r_{2} e^{\mathrm{i} \phi_{2}}=r_{1} r_{2} e^{\mathrm{i}\left(\phi_{1}+\phi_{2}\right)}
$$

where $z_{k}=r_{k} e^{\mathrm{i} \phi_{k}}(k=1,2)$. It is simply to show by induction that

$$
(\cos \phi+\mathrm{i} \sin \phi)^{n}=\cos n \phi+\mathrm{i} \sin n \phi \quad \text { (A. De Moivre). }
$$

In the history of complex numbers the so-called circle division equation

$$
z^{n}=1
$$

plays an important role. We have

$$
\begin{gathered}
z^{n}-1=\left(z-\zeta_{0}\right)\left(z-\zeta_{1}\right) \ldots\left(z-\zeta_{n-1}\right) \\
\zeta_{k}=\mathbf{e}^{\frac{2 \pi k}{n}} \quad 0 \leq k \leq(n-1)
\end{gathered}
$$

The complex numbers $\zeta_{k}$ are just the cornerpoints lying on the unit circle of a equilateral $n$-angle. Moreover, they are the zeros of the polynomial $p(z):=z^{n}-1$, which is a special case of the so-called fundamental theorem of the algebra:

Theorem 1 A non-constant polynomial of degree $n$ posesses exactly $n$ zeros. Here the number of zeros is to compute in accordance with its multiplicity.

Corollary 2 The quadratic equation

$$
z^{2}+\alpha z+\beta=0 \quad(\alpha, \beta \in \mathbb{C})
$$

has at most two solutions $z_{1}$ and $z_{2}$ in $\mathbb{C}$.
Remark 3 An alternative definition can be given in the following way: Let

$$
M_{\mathbf{C}}:=\left\{\left(\begin{array}{cc}
x & -y \\
y & x
\end{array}\right): x, y \in \mathbb{R}\right\} \subset \mathbb{R}(2)
$$

furnished with the operation matrix addition and (real) multiplication. $M_{\mathrm{C}}$ has the structure of a field, which is isomorphic to the field of complex numbers $\mathbb{C}$.

### 1.2 Hamilton's creation - real quaternions

There is a little story on the discovery of quaternions. Briefly we will describe this: It was on Monday, October 16 in 1843 and it happened that in this morning W.R. Hamilton had to preside at a meeting in the Royal Irish Academy. He was walking with his wife along the Royal Canal in Dublin when the answer of this 10 years old problem came to his mind. With his knife he then and there carved on a stone on Broome Bridge the formulae:

$$
\begin{equation*}
\mathbf{i}^{2}=\mathrm{j}^{2}=\mathbf{k}^{2}=\mathbf{i j k}=-1 \tag{1}
\end{equation*}
$$

By the way in the above mentioned letter of Hamilton named this bridge errounously Brougham Bridge and up to now this bridge bears this name given by W.R. Hamilton.

Hamilton called these numbers

$$
q=a+b \mathbf{i}+c \mathbf{j}+d \mathbf{k}
$$

quaternions (cf. [18]). In that way the notion of a quaternion was introduced. The first paper on quaternions appeared 14. November 1843 in the Council Books of the Royal Academy at the First General Meeting of the Session (cf. [1]).

Consider the real four-dimensional vector space $\mathbb{R}^{4}$ with its standard basis $e_{0}, \mathbf{e}_{1}, \mathbf{e}_{\mathbf{2}}$ and $\mathbf{e}_{3}$ where the element $e_{k}(k=0,1,2,3)$ is to identify with that 4 -tuple which has at the ( $k+1$ )-th component the number one and has zeros otherwise in a cyclic denotation. The element $e_{0}$ will be identified in the following by 1 and will mostly be omitted. Notice that $\mathbf{e}_{\mathbf{k}}(k=1,2,3)$ are so-called "axial" vectors in $\mathbb{R}^{4}$. In case of cancellation of the first component one can identify these "axial" vectors in $\mathbb{R}^{4}$ with ordinary vectors or "polar" vectors in $\mathbb{R}^{3}$ (cf. [21]). It is necessary to use sometimes these identifications. An arbitrary element $x \in \mathbb{R}^{4}$ can now be represented by

$$
x=x_{0}+\mathbf{x} \quad \text { with } \quad \mathbf{x}=x_{1} \mathbf{e}_{1}+x_{2} \mathbf{e}_{\mathbf{2}}+x_{3} \mathbf{e}_{\mathbf{3}} \quad\left(x_{k} \in \mathbb{R}, k=1,2,3\right)
$$

The part $x_{0}$ is called scalar part of $x$ written $x_{0} e_{0}=: \mathrm{Sc}(x)$ and the part $\mathbf{x}$ is called vector part and written $\mathbf{x}=\operatorname{Vec}(x)$.

Recalling the computation rules in $\mathbb{R}^{4}$ we have for $x, y \in \mathbb{R}^{4}$ :
(i) Identity: $x=y$ if and only if $x_{0}=y_{0}$ and $\mathbf{x}=\mathbf{y}$
(ii) Scalar multiplication: $\mathbb{R} \times \mathbb{R}^{4} \ni(\lambda, x) \rightarrow \lambda x \in \mathbb{R}^{4}$ with

$$
\lambda x=\lambda x_{0}+\lambda x_{1} \mathbf{e}_{1}+\lambda x_{2} \mathbf{e}_{\mathbf{2}}+\lambda x_{3} \mathbf{e}_{\mathbf{3}}=: \lambda x_{0}+(\lambda \mathbf{x}) \quad(\lambda \in \mathbb{R}) .
$$

(iii) Addition: $\mathbb{R}^{4} \times \mathbb{R}^{4} \ni(x, y) \rightarrow x+y \in \mathbb{R}^{4}$ with

$$
\begin{aligned}
x+y & =\left(x_{0}+y_{0}\right)+\left(x_{1}+y_{1}\right) \mathbf{e}_{\mathbf{1}}+\left(x_{2}+y_{2}\right) \mathbf{e}_{\mathbf{2}}+\left(x_{3}+y_{3}\right) \mathbf{e}_{\mathbf{3}} \\
& =:\left(x_{0}+y_{0}\right)+\mathbf{x}+\mathbf{y}
\end{aligned}
$$

and $x_{k}, y_{k} \in \mathbb{R}$.
Now we define a multiplication in the following way:
$\mathbb{R}^{4} \times \mathbb{R}^{4} \ni(x, y) \rightarrow x y \in \mathbb{R}^{4}$ with

$$
x y=x_{0} y_{0}-\mathbf{x} \cdot \mathbf{y}+x_{0} \mathbf{y}+y_{0} \mathbf{x}+\mathbf{x} \times \mathbf{y}
$$

where

$$
\begin{aligned}
\mathbf{x} \cdot \mathbf{y} & :=x_{1} y_{1}+x_{2} y_{2}+x_{3} y_{3} \\
\mathbf{x} \times \mathbf{y} & :=\left|\begin{array}{lll}
\mathbf{e}_{\mathbf{1}} & \mathbf{e}_{\mathbf{2}} & \mathbf{e}_{\mathbf{3}} \\
x_{1} & x_{2} & x_{3} \\
y_{1} & y_{2} & y_{3}
\end{array}\right| \\
& :=\left(x_{2} y_{3}-x_{3} y_{2}\right) \mathbf{e}_{\mathbf{1}}+\left(x_{3} y_{1}-x_{1} y_{3}\right) \mathbf{e}_{\mathbf{2}}+\left(x_{1} y_{2}-x_{2} y_{1}\right) \mathbf{e}_{\mathbf{3}} .
\end{aligned}
$$

Definition 4 The space $\mathbb{R}^{4}$ furnished with the above defined multiplication rule has the structure of an algebra and is called algebra of real quaternions. Its elements are called simply quaternions. In honour of W.R. Hamilton we will denote this algebra by $\mathbb{H}$.

If $x=\mathbf{x}$ then $x$ is called pure quaternion or simply vector. The set of all pure quaternions is denoted by $\operatorname{Vec} \mathbb{H} \subset \mathbb{H}$, while the set of all scalars will be denoted by $\mathbf{S c} \mathbb{H} \subset \mathbb{H}$. Vec $\mathbb{H}$ and $\mathbf{S c} \mathbb{H}$ are real linear subspaces of $\mathbb{H}$ which are not closed relatively to the multiplication introduced above.

Corollary 5 The map $\rho_{y}: \mathbb{R}^{3} \ni \underline{x} \rightarrow-\underline{y} \underline{x} \underline{y}^{-1} \in \mathbb{R}^{3}$ is a reflection in the plane which lies orthogonal to the vector $\underline{y}$.

Proof. The map $\rho_{y}$ is linear and $\rho_{y}(\underline{y})=-\underline{y}$, while for any vector $r \perp \underline{y}$ it follows that $\rho_{y}(r)=-\underline{y}^{r} \underline{y}^{-1}=r \underline{y} \underline{y}^{-1}=r$.

Corollary 6 Each rotation in $\mathbb{R}^{3}$ has for some non-zero $\underline{y}$ the form $-\rho_{y}$. Conversely, any such map can be seen as rotation in $\mathbb{R}^{3}$.

Proof. We already know that the product of two plane reflections is just a rotation and vice versa.

Proposition 7 Let $y$ be a quaternion. Then there exists a vector $\underline{a} \neq 0$, such that ya is also a vector.

Proof. Let $\underline{a}$ be a vector in $\mathbb{R}^{3}$ orthogonal to the vector part of $y$. Then

$$
y \underline{a}=y_{0} \underline{a}+\underline{y} \times \underline{a} \in \mathbb{R}^{3} .
$$

Corollary 8 Each quaternion can be described as a product of two vectors.

Proposition 9 An arbitrary unit quaternion can be represented as the product $y x y^{-1} x^{-1}$, where $x=\underline{x}$ and $y=\underline{y}$ are non-zero vectors.

Proof. We know from proposition 7 that for any unit quaternion $e$ a nonzero vector $\underline{x}$ exists such that $e \underline{x}$ is a vector. Because of $|e|=1$ we have $|e \underline{x}|=|\underline{x}|$. In this way $e \underline{x}$ has to be a rotation. Then there exists a vector $\underline{y} \neq 0$ with $e \underline{x}=\underline{y} \underline{x}^{-1}$ and so the statement follows.

### 1.2.1 Representation of real quaternions

We will group here some of the most important properties of the representation of real quaternions.

Theorem 10 An arbitrary quaternion $x \in \mathbb{H}, \underline{x} \neq 0$ permits the representation

$$
x=|x|(\cos \phi+\omega(x) \sin \phi),
$$

where $\phi=\operatorname{arccot}\left(x_{0} /|\underline{x}|\right)$ and $\omega(x)=\underline{x} /|\underline{x}| \in S^{3}$.

Proof. It is well known that

$$
\sin \phi=\frac{1}{\sqrt{\left(1+\cot ^{2} \phi\right)}} \text { and } \cos \phi=\frac{\cot \phi}{\sqrt{\left(1+\cot ^{2} \phi\right)}}
$$

We obtain under our assumption

$$
\sin \operatorname{arccot} \frac{x_{0}}{|\underline{x}|}=\frac{1}{\sqrt{1+\left(\frac{x_{0}}{|\underline{x}|}\right)^{2}}} \text { and } \cos \operatorname{arccot} \frac{x_{0}}{|\underline{x}|}=\frac{\frac{x_{0}}{|\underline{x}|}}{\sqrt{1+\left(\frac{x_{0}}{\mid \underline{x}}\right)^{2}}}
$$

Then we get by a straightforward calculation $\cos \phi+(\underline{x} /|\underline{x}|) \sin \phi=x /|x|$ which verifies our theorem.

Example 11 Let $x=3+2 e_{1}+2 e_{2}+e_{3}$ then $|x|=3 \sqrt{2},|\underline{x}|=3, \phi=45^{\circ}$. Thus we obtain the representation

$$
x=3 \sqrt{2}\left[\cos 45^{\circ}+\frac{\left(2 e_{1}+2 e_{2}+e_{3}\right)}{3} \sin 45^{\circ}\right] .
$$

Corollary 12 (Morvre's formula). Let $x \in \mathbb{H}, \underline{x} \neq 0, n \in \mathbb{N}$, then the following formula is valid:

$$
(\cos \phi+\omega(x) \sin \phi)^{n}=\cos n \phi+\omega(x) \sin n \phi
$$

Proposition 13 The $\mathbb{R}$-linear hull of the set $\{1, x\}$, where $x$ is not real, forms a subalgebra which is isomorphic to $\mathbb{C}$.

Theorem 14 The algebra of real quaternions $\mathbb{H}$ can be represented by the matrices

$$
\left(\begin{array}{ll}
1 & 0 \\
0 & 1
\end{array}\right), \quad\left(\begin{array}{cc}
-i & 0 \\
0 & i
\end{array}\right), \quad\left(\begin{array}{cc}
0 & -1 \\
1 & 0
\end{array}\right), \quad\left(\begin{array}{cc}
0 & -i \\
-i & 0
\end{array}\right)
$$

Therefore each quaternion $x$ permits the representation

$$
x=\left(\begin{array}{cc}
\bar{z}_{1} & -\bar{z}_{2} \\
z_{2} & z_{1}
\end{array}\right)
$$

with $z_{1}:=x_{0}+i x_{1}$ and $z_{2}:=x_{2}+i x_{3}$.

Corollary 15 The algebra of real quaternions $\mathbb{H}$ can be generated by the real matrices

$$
\begin{aligned}
& \left(\begin{array}{llll}
1 & 0 & 0 & 0 \\
0 & 1 & 0 & 0 \\
0 & 0 & 1 & 0 \\
0 & 0 & 0 & 1
\end{array}\right), \quad\left(\begin{array}{cccc}
0 & -1 & 0 & 0 \\
1 & 0 & 0 & 0 \\
0 & 0 & 0 & 1 \\
0 & 0 & -1 & 0
\end{array}\right), \\
& \left(\begin{array}{cccc}
0 & 0 & -1 & 0 \\
0 & 0 & 0 & -1 \\
1 & 0 & 0 & 0 \\
0 & 1 & 0 & 0
\end{array}\right), \quad\left(\begin{array}{cccc}
0 & 0 & 0 & 1 \\
0 & 0 & -1 & 0 \\
0 & 1 & 0 & 0 \\
-1 & 0 & 0 & 0
\end{array}\right) .
\end{aligned}
$$

### 1.3 Pauli algebra - a realization of complex quaternions

More than real quaternions, complex quaternions take an important part in theoretical physics. Let us now discuss the fundamental properties of quaternions with complex-valued coefficients. We will use the so-called Pauli matrices:

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

Pauli matrices form an algebra $P$. Then we get the representation of an arbitrary element $x \in P$ in the form

$$
x=x_{0} \sigma_{0}+x_{1} \sigma_{1}+x_{2} \sigma_{2}+x_{3} \sigma_{3}+x_{4} \sigma_{2} \sigma_{3}+x_{5} \sigma_{3} \sigma_{1}+x_{6} \sigma_{1} \sigma_{2}+x_{7} \sigma_{1} \sigma_{2} \sigma_{3}
$$

We have to distinguish four classes of elements, namely complex linear combinations of scalars $\sigma_{0}$, vectors $\sigma_{1}, \sigma_{2}, \sigma_{3}$, bivectors $\sigma_{2} \sigma_{3}, \sigma_{3} \sigma_{1}, \sigma_{1} \sigma_{2}$ and pseudoscalars $\sigma_{1} \sigma_{2} \sigma_{3}$. Pauli matrices satisfy the conditions $\sigma_{i} \sigma_{j}+\sigma_{j} \sigma_{i}=2 \delta_{i j}$, where $\delta_{i j}$ denotes Kronecker's symbol. On account of $\left(\sigma_{1} \sigma_{2} \sigma_{3}\right)^{2}=-\sigma_{0}$ we obtain that the linear space generated by $\left\{\sigma_{0}, \sigma_{1} \sigma_{2} \sigma_{3}\right\}$ is isomorphic to the field of complex numbers $\mathbb{C}$.

Remark 16 The centre of $P$ is $\mathbb{C}$.
Proposition 17 With $\varepsilon:=i \sigma_{0}$ (Hodge star map) we get

$$
\begin{array}{r}
-\varepsilon \sigma_{0}=\sigma_{1} \sigma_{2} \sigma_{3},-\varepsilon \sigma_{1}=\sigma_{2} \sigma_{3},-\varepsilon \sigma_{2}=\sigma_{3} \sigma_{1},-\varepsilon \sigma_{3}=\sigma_{1} \sigma_{2}, \varepsilon \sigma_{1} \sigma_{2}=+\sigma_{3} \\
\varepsilon \sigma_{2} \sigma_{3}=+\sigma_{1}, \varepsilon \sigma_{3} \sigma_{1}=+\sigma_{2}, \varepsilon \sigma_{1} \sigma_{2} \sigma_{3}=+\sigma_{0}
\end{array}
$$

i.e. the multiplication by $\varepsilon$ transforms scalars into pseudoscalars, vectors into bivectors, bivectors into vectors, and pseudoscalars into scalars.

## 2 Clifford algebras

We will consider a class of algebras which form the frame for our further considerations. This structure shall incorporate number systems like complex numbers and quaternions as well as vector and multivector systems. We will be able to realize in this algebra a large number of fundamental geometric and analytic ideas. Let us start with the description of this algebra in expressing the main ideas.

Definition 1 (Real Clifford algebra). An associative algebra over $\mathbb{R}$ with unit $e_{0}:=1$ which is freely generated by $n$ basis elements $e_{1}, \ldots, e_{n}$ together with the defining relations

$$
\begin{equation*}
e_{i} e_{j}+e_{j} e_{i}=-2 \delta_{i j} \quad(i, j=1, \ldots, n) \tag{2}
\end{equation*}
$$

is called real Clifford algebra (denoted by: $C \ell_{0, n}$ ). Here $\delta_{i j}$ denotes the Kronecker symbol. This means

$$
e_{i} e_{j}+e_{j} e_{i}=0 \quad \text { iff } \quad i \neq j, \quad e_{i}^{2}=-1 \quad(i, j=1, \ldots, n)
$$

Usually the elements $e_{1}, e_{2}, \ldots, e_{n}$ belong to the vector space $\mathbb{R}^{n}$ and form there a basis. After the canonical embedding we have identified the "initial" vectors $e_{1}, e_{2}, \ldots, e_{n}$ in $\mathbb{R}^{n}$ with corresponding elements in $C \ell_{0, n}$. Furthermore, we identify the unit element in $C \ell_{0, n}$ with $1 \in \mathbb{R}$.

Definition 2 (Clifford algebra of the (p,q)-type). Let $e_{1}, \ldots, e_{q}, \ldots, e_{p+q}$ form a basis of $\mathbb{R}^{p+q}$. We introduce the denotation

$$
\dot{e}_{j}:=\mathbf{i} e_{q+j} .
$$

The Clifford algebra of the ( $p, q$ )-type (denoted by: $C \ell_{p, q}$ ) is defined following an idea of $F$. Sommen by the real associative algebra with unit $e_{0}:=1$ freely generated by elements of the form

$$
\dot{e}_{1}, \ldots, \dot{e}_{p}, e_{1}, \ldots, e_{q}
$$

which fulfil the defining relations

$$
\begin{aligned}
& \dot{e}_{j} \dot{e}_{k}+\dot{e}_{k} \dot{e}_{j}=2 \delta_{j k}, \\
& e_{j} e_{k}+e_{k} e_{j}=-2 \delta_{j k}, \\
& e_{j} \dot{e}_{k}+\dot{e}_{k} e_{j}=0
\end{aligned}
$$

Definition 3 The Clifford algebra $C \ell_{n, 0}$ is called complex Clifford algebra can be seen as complexification of the real Clifford algebra $C \ell_{0, n}$ i.e. $C \ell_{n, 0}=$ $\mathbb{C} \otimes C \ell_{0, n}$.

### 2.1 Involutions and Bott's classification

### 2.1.1 Involutions

In a Clifford algebra there are at least three involutions. Depending on the degree $k$ of $x \in C \ell_{p, q}$ we define the following:
(i) grade involution (or main involution) by the formula $\hat{x}:=(-1)^{k} x$,
(ii) reverse involution (or inversion) by the relation $\tilde{x}:=(-1)^{\frac{k(k-1)}{2}} x$,
(iii) conjugation by $\bar{x}:=(-1)^{\frac{k(k+1)}{2}} x$.

Proposition 4 These involutions satisfy the following isomorphic and antiisomorphic conditions with respect to the multiplication of two elements

$$
(\widehat{x y})=\hat{x} \hat{y}, \quad(\widetilde{x y})=\tilde{y} \tilde{x}, \quad(\overline{x y})=\bar{y} \bar{x} .
$$

Reversion and conjugation can be extended complex-linear to complexified Clifford algebras. In addition it is usual to introduce a complex conjugation by $x^{*}=x_{1}-i x_{2}$ for $x=x_{1}+i x_{2}\left(x_{k} \in C \ell_{p, q}\right)$.

Example 5 Let be $x \in \mathbb{C} \otimes C \ell_{3,0}$ the element

$$
x=1+3 e_{1}+2 e_{2}+e_{21}+i e_{23}+e_{123},
$$

where $e_{k l}:=e_{k} e_{l}$ and $e_{k l m}:=e_{k} e_{l} e_{m}$, then we have for the above-mentioned involutions:

$$
\begin{aligned}
\hat{x} & =1-3 e_{1}-2 e_{2}+e_{21}+i e_{23}-e_{123}, \\
\tilde{x} & =1+3 e_{1}+2 e_{2}-e_{21}-i e_{23}-e_{123}, \\
\bar{x} & =1-3 e_{1}-2 e_{2}-e_{21}-i e_{23}+e_{123}, \\
x^{*} & =1+3 e_{1}+2 e_{2}+e_{21}-i e_{23}+e_{123} .
\end{aligned}
$$

Definition 6 Let $x \in C \ell_{n, 0} \cup C \ell_{0, n}$ then $\sqrt{S c(\tilde{x} x)}$ is called absolute value of $x$. It is denoted by $|x|$.

Remark 7 For $x \in C \ell_{p, q}$ the product $\tilde{x} x$ is not necessarily real. P. LounesTO gave a counterexample: Let $p=3, q=1, x=\left(1+e_{1}\right)\left(1+e_{234}\right) \in C \ell_{3,1}$ and $x \tilde{x}=0 \in \mathbb{R}$ but $\tilde{x} x=4 e_{234}+4 e_{1234}$ obviously does not belong to $\mathbb{R}$. The so-called square norm which can be defined by $\bar{x} x=\|x\|$ is often used.

### 2.1.2 Table of matrix representations

Clifford algebras are isomorphic as associative algebras with unit to corresponding matrix algebras. Setting now $s:=(p+q) / 2$ one can show that only eight different rings of matrices occur in the description of the structure of Clifford algebras. The following table gives an overview on the possible structures. For this reason let $M(d, F)$ denote the ring of $d \times d$ matrices over the field $F$.

| $(p-q) \bmod 8$ | matrix ring |
| :---: | :---: |
| 0 | $M\left(2^{[s]}, \mathbb{R}\right)$ |
| 1 | $M\left(2^{[s]}, \mathbb{R}\right) \oplus M\left(2^{[s]}, \mathbb{R}\right)$ |
| 2 | $M\left(2^{[s]}, \mathbb{R}\right)$ |
| 3 | $M\left(2^{\left[\frac{-1}{2}\right]}, \mathbb{C}\right)$ |
| 4 | $M\left(2^{[s-1]}, \mathbb{H}\right)$ |
| 5 | $M\left(2^{\left[\frac{-3}{2}\right]}, \mathbb{H}\right) \oplus M\left(2^{\left[\frac{\rho-3}{2}\right]}, \mathbb{H}\right)$ |
| 6 | $M\left(2^{[s-1]}, \mathbb{H}\right)$ |
| 7 | $M\left(2^{\left.\frac{[-1}{2}\right]}, \mathbb{C}\right)$ |

We mention that $s$ depends only on the dimension of the algebra and not on its special structure. The denotation $[s]$ means the entire part of $s$.

### 2.2 Examples of Clifford algebras

An excellent possibility of representing Clifford algebras, especially those of low dimensions, is to do it with the help of the Pauli matrices $\sigma_{0}, \sigma_{1}, \sigma_{2}, \sigma_{3}$ introduced in Section 1.3. In this way we will interchange the basic elements $e_{i}(i=0,1,2,3)$ by the PaUli matrices $\sigma_{i}(i=0,1,2,3)$. It is easy to see that $\sigma_{i}^{2}=\sigma_{0}$ and that the elements $\sigma_{i}$ also fulfil the anti-commutativity property. We will agree upon the use of the symbol " = " for the identification of algebras. Let us now discuss the following examples:
$\mathbf{C} \ell_{0,0}$ : In this case the basis consists only of the identity element $\sigma_{0}$. Therefore $C \ell_{0,0}=\mathbb{R}$. An isomorphism is explicitly given by $x=x_{0} \sigma_{0}$.
$\mathbf{C} \ell_{0,1}$ : This algebra is generated by $\sigma_{0}$ and $i \sigma_{2}$ with $\sigma_{2}^{2}=-\sigma_{0}$, therefore $C \ell_{0,1}=\mathbb{C}$. This is clear by the representation $x=x_{0} \sigma_{0}+x_{1} i \sigma_{2}$.
$\mathbf{C} \ell_{0,2}$ : The matrices $\sigma_{0,-}, i \sigma_{1},-i \sigma_{2}, \sigma_{1} \sigma_{2}\left(=-i \sigma_{3}\right)$ form a basis in $C \ell_{0,2}$. For any element $x$ of this algebra we obtain the representation

$$
x=\left(\begin{array}{cc}
x_{0}-i x_{1} & -x_{2}+i x_{3} \\
x_{2}+i x_{3} & x_{0}+i x_{1}
\end{array}\right)
$$

with $x_{i} \in \mathbb{R} \quad(i=0,1,2,3)$.
This describes a matrix realization of the algebra of real quaternions $\mathbb{H}$. Hence $C \ell_{0,2}=\mathbb{H}$.
$\mathbf{C} \ell_{\mathbf{3}, \mathbf{0}}$ : The matrices $\sigma_{0}, \sigma_{1}, \sigma_{2}, \sigma_{3}, \sigma_{1} \sigma_{2}, \sigma_{2} \sigma_{3}, \sigma_{3} \sigma_{1}, \sigma_{1} \sigma_{2} \sigma_{3}$ form a basis in the algebra $C \ell_{3,0}$. We find that $\sigma_{1} \sigma_{2}=-i \sigma_{3}, \sigma_{2} \sigma_{3}=-i \sigma_{1}, \sigma_{3} \sigma_{1}=-i \sigma_{2}$ and $\sigma_{1} \sigma_{2} \sigma_{3}=-i \sigma_{0}$. An arbitrary element $x$ of this algebra can be represented by

$$
x=\left(x_{0}-x_{7} i\right) \sigma_{0}+\left(x_{1}-x_{5} i\right) \sigma_{1}+\left(x_{2}-x_{6} i\right) \sigma_{2}+\left(x_{3}-x_{4} i\right) \sigma_{3}
$$

Then it is easy to get $C \ell_{3,0}=\mathbb{C} \otimes \mathbb{H}$. This algebra coincides with the Pauli algebra. The denotation by $P$ is made in honour of Wolfgang Pauli.
$\mathbf{C} \ell_{1,3}$ : In this algebra a basis can be constructed by the so-called $d$ numbers $\gamma_{0}, \gamma_{1}, \gamma_{2}, \gamma_{3}$ of DIRAC which are formal $2 \times 2$ matrices in which entries are the Pauli matrices. So we have

$$
\gamma_{0}=\left(\begin{array}{cc}
\sigma_{0} & 0 \\
0 & -\sigma_{0}
\end{array}\right), \gamma_{i}=\left(\begin{array}{cc}
0 & -\sigma_{i} \\
\sigma_{i} & 0
\end{array}\right)
$$

for $i=1,2,3$. Clearly, we have $\gamma_{0}^{2}=\gamma_{0}$ and $\gamma_{i}^{2}=-\gamma_{0}$. The basis in $C \ell_{1,3}$ is now given by

$$
\begin{aligned}
& 1, \\
& \gamma_{0}, \gamma_{1}, \gamma_{2}, \gamma_{3}, \\
& \gamma_{1} \gamma_{0}, \gamma_{2} \gamma_{0}, \gamma_{3} \gamma_{0}, \gamma_{1} \gamma_{2}, \gamma_{2} \gamma_{3}, \gamma_{3} \gamma_{1}, \\
& \gamma_{0} \gamma_{1} \gamma_{2}, \gamma_{0} \gamma_{2} \gamma_{3}, \gamma_{0} \gamma_{3} \gamma_{1}, \gamma_{1} \gamma_{2} \gamma_{3}, \\
& \gamma_{0} \gamma_{1} \gamma_{2} \gamma_{3},
\end{aligned}
$$

where $1: \mathbb{R}^{1,3} \longmapsto \mathbb{R}^{1,3}$ is the unit matrix. This algebra is called space-time algebra. One can show that this algebra is isomorphic to $M(4, \mathbb{R})$.

## 3 Important operators in the theory

### 3.1 Classes of monogenic functions

Let $G \subset \mathbb{R}^{n}$ be an open (nonempty) domain and $\Gamma=\partial G$ a piecewise smooth surface. We denote by $\mathcal{B}(\Omega)$ one of the following BanACH spaces of real or complex valued functions on $\Omega$ where $\Omega$ can be $G, \partial G=\Gamma$, or any suitable subset of $\bar{G}$ :
(i) $C^{(k)}(\Omega)(k=1,2, \ldots)$ is the space of $k$-times continuously $\mathbb{R}$-differentiable functions in $\Omega$.
(ii) $C^{(k, \alpha)}(\Omega)(k=0,1,2, \ldots ; \alpha \in(0,1])$ is the space of $k$-times $\mathbb{R}$-differentiable functions, whose $k$-th derivative is Hölder continuous with the exponent $\alpha$.
(iii) $L_{p}(\Omega),(1 \leq p<\infty)$, is just the space of all functions, whose $p$-th power is LEBESGUE integrable in $\Omega$.
(iv) $W_{p}^{k}(\Omega)$ is the space of $k$-times $\mathbb{R}$-differentiable functions in SoboLEv's sense, whose $k$-th derivative belongs to $L_{p}(\Omega)$.
(v) $\stackrel{\circ}{W}_{p}^{k}(G)(1 \leq p<\infty ; k=1,2, \ldots)$, is the space of $k$-times $\mathbb{R}$ differentiable functions in Sobolev's sense which $k$-th derivative belongs to $L_{p}(G)$ and vanish on the boundary $\Gamma$.

Definition 1 Let $\mathcal{E}$ be either $\mathbb{R}$ or $\mathbb{C}$. Then the space $\mathcal{B}(\Omega) \otimes_{\mathcal{E}} \mathcal{A}$, where $\mathcal{A}:=C \ell_{p, q}$ is called $\mathcal{E}$-linear space $\mathcal{B}(\Omega, \mathcal{A})$ of all $\mathcal{A}$-valued functions.

Remark 2 For both $\mathcal{A}=C \ell_{0, n}$ or $\mathcal{A}=C \ell_{n, 0}$ the space $\mathcal{B}(\Omega, \mathcal{A})$ is a BA NACH space. This means that

$$
\mathcal{B}(\Omega, \mathcal{A})=\left\{u=\sum_{i=0}^{\operatorname{dim} \mathcal{A}} u_{i} \varepsilon_{i}: u_{i} \in \mathcal{B}\right\}
$$

is furnished with the norm $\|u\|=\left\{\sum_{i=0}^{\operatorname{dim} \mathcal{A}}\left\|u_{i}\right\|^{2}\right\}^{\frac{1}{2}}$. Here $\varepsilon_{k}$ is a $k$-vector such that $\left\{\varepsilon_{0}, \ldots, \varepsilon_{\text {dim } \mathcal{A}}\right\}$ form a basis in $\mathcal{A}$.

Definition 3 Let $u$ be given in a basis representation $u=\sum_{i=0}^{\operatorname{dim} A} u_{i}(x) \varepsilon_{i}$, $x \in \Omega \subset \mathbb{R}^{n}$, where $\left\{\varepsilon_{0}, \ldots, \varepsilon_{\operatorname{dimA}}\right\}$ is a basis in $\mathcal{A}$, then we say that $u \in$ $\mathcal{B}(\Omega, \mathcal{A})$ if and only if $u_{i} \in \mathcal{B}(\Omega)$. Often we will briefly write instead of $\mathcal{B}(\Omega, \mathcal{A}) \mathcal{B}(\Omega)$.

Corollary 4 Topological properties as continuity, differentiability, and integrability of the coefficients $u_{i}(x), x \in \Omega$, in a suitable basis representation transfer to the $\mathcal{A}$-valued function $u=\sum_{i=0}^{\operatorname{dim} A} u_{i}(x) \varepsilon_{i}$.

### 3.1.1 Clifford regular functions

In general regular functions are those which belong to the kernel of a proper differential operator.

Definition 5 The operator

$$
D:=\sum_{i=1}^{n} e_{i} \partial_{i}
$$

which acts on the space $C^{1}\left(G, C \ell_{p, q}\right)$, where $G \subset \mathbb{R}^{p, q}$ is called Dirac operator and $\partial_{i}:=\partial / \partial x_{i}$ is the $i$-th partial derivative and $x=\sum_{i=1}^{n} x_{i} e_{i} \in \mathbb{R}^{p, q}$. The operator

$$
\partial:=\partial_{0}+D
$$

which acts on $C^{1}\left(G, C \ell_{p, q}\right)$ with $G \subset \mathbb{R}^{p, q}$ is called CAUCHY-FUETER operator where $\partial_{0}:=\partial / \partial x_{0}$ and $x=x_{0}+\sum_{i=1}^{n} x_{i} e_{i} \in \mathbb{R} \oplus \mathbb{R}^{p, q}$.

Definition 6 Let $G \subset \mathbb{R}^{p, q}, p+q=n$, be an $n$-dimensional vector space with signature $p, q$. A function $u \in C^{1}\left(G, C l_{p, q}\right)$ is called left Clifford regular (right Clifford regular) if and only if

$$
D u=0((u D)=0)
$$

A function which is both left and right Clifford regular is called two-sided Clifford regular. Functions which fulfil the condition

$$
\partial u=0 \quad((u \partial)=0)
$$

are called left(right) Clifford holomorphic. We agree that the denotations $C \ell_{p, q^{-}}$regular or $C \ell_{p, q^{-}}$holomorphic will sometimes be used if it is necessary to emphasize this special algebra.

Remark 7 In $C \ell_{0, n}$ left(right) Clifford regular functions are often called left (right) monogenic functions. In such algebras the DIRAC operator is sometimes called standard Euclidean Dirac operator. The set of all monogenic functions is denoted by $\mathcal{M}\left(G, C \ell_{0, n}\right)$.

Through the whole paper we will view the action of $D$ and $\partial$ as an action on the left. It is sufficiently known that there are corresponding results for action of these operators on the right. Sometimes we will simply say regular functions instead of left regular functions.

Remark 8 The CaUchy-Fueter operator $\partial$ which acts on $C^{1}\left(G, C \ell_{p, q}\right)$ with $G \in \mathbb{R}^{p, q}$ can be considered as the DIRAC operator on $C^{1}\left(G, C \ell_{p+1, q}\right)$ with $G \in \mathbb{R}^{p+1, q}$.

Example 9 Let $G \subset \mathbb{R}^{3}, e_{0}=(1,0,0,0)^{T}, e_{1}=(0,1,0,0)^{T}, e_{2}=(0,0,1,0)^{T}$, $e_{3}=(0,0,0,1)^{T}$ form a basis in $\mathbb{R}^{4}$. Under the governing conditions
(i) $e_{i} e_{j}+e_{j} e_{i}=-2 \delta_{i j}(i, j=1,2,3 ; i<j)$,
(ii) $e_{0}^{2}=1, e_{i} e_{0}=e_{0} e_{i}=e_{i}$,
(iii) $e_{1} e_{2}=e_{3}$,
$\left\{e_{0}, e_{1}, e_{2}, e_{3}\right\}$ also form a basis in the algebra of real quaternions $\mathbb{H}$. The DIRAC operator has the form

$$
D:=e_{1} \partial_{1}+e_{2} \partial_{2}+e_{3} \partial_{3} .
$$

Together with $u=u_{0} e_{0}+\underline{u}, \underline{u}=\sum_{i=1}^{3} u_{i} e_{i}$, it follows that

$$
D u=-\operatorname{div} \underline{u}+\operatorname{rot} \underline{u}+\operatorname{grad} u_{0}
$$

where

$$
\operatorname{div} \underline{u}:=\sum_{\mathrm{i}=1}^{3} \partial_{\mathrm{i}} u_{\mathrm{i}}, \quad \operatorname{grad} u_{0}:=\sum_{\mathrm{i}=1}^{3} \mathrm{e}_{\mathrm{i}} \partial_{\mathrm{i}} u_{0}, \quad \operatorname{rot} \underline{u}:=\left|\begin{array}{lll}
e_{1} & e_{2} & e_{3} \\
\partial_{1} & \partial_{2} & \partial_{3} \\
u_{1} & u_{2} & u_{3}
\end{array}\right| .
$$

For $(u D)$ we get in an analogous way

$$
(u D)=-\operatorname{div} \underline{u}-\operatorname{rot} \underline{u}+\operatorname{grad} u_{0} .
$$

Proposition 10 Quaternionic regular functions can be also characterized by the following properties:
(i) quaternionic-regular functions are sourceless and in the case of $u_{0}=$ const. also rotationless vector fields.
(ii) $\mathrm{Cl}_{3,0}$-regular functions are rotationless and sourceless vector fields with $u_{0}=$ const.

Definition 11 Let be $\bar{\partial}$ defined by

$$
\bar{\partial}:=\partial_{0}-D
$$

This operator is called adjungated CAUCHY-Fueter operator. In $C l_{p, q}, p+q=n$, the operator $\bar{D}$ defined by

$$
\bar{D}=\sum_{i=1}^{n} \overline{e_{i}} \partial_{i}
$$

is called adjoint DIRAC operator.
Assume now that $G$ is a domain $\mathbb{R}^{3}$, or sometimes in $\mathbb{R}^{4}$.
Proposition 12 Let be $u \in C^{2}\left(G, C \ell_{0, n}\right)$. Then we have

$$
\bar{D} D u=D \bar{D} u=\Delta u
$$

where $\Delta u=\sum_{i=1}^{n} \partial_{i}^{2} u$.
Corollary 13 Let $\left.u \in \operatorname{ker} D \cap C^{2}\left(G, C \ell_{0, n}\right)\right)$ then $u \in \operatorname{ker} \Delta \cap C^{2}\left(G, C \ell_{0, n}\right)$, e.g., Clifford regular functions of this type are harmonic.

Proposition 14 (Generalized Leibniz rule). Let be $u, v \in C^{1}(G, \mathbb{H})$ then

$$
D(u v)=(D u) v+\bar{u} D v+2 S c(u D) v .
$$

Proof. In [11] is contained the proof for the algebra of real quaternions.

Corollary 15 (Product rules in vector analysis). Let $u, v \in C^{1}(G, \mathbb{H})$. Then in accordance with the definition of the vector field operations grad, div, rot we have the following relations:
(i) $\operatorname{grad}\left(u_{0} v_{0}\right)=\left(\operatorname{grad} u_{0}\right) v_{0}+u_{0} \operatorname{grad} v_{0}$,
(ii) $\operatorname{div}\left(u_{0} \underline{v}\right)=\left(\operatorname{grad} u_{0}\right) \underline{v}+u_{0} \operatorname{div} \underline{v}$,
(iii) $\operatorname{rot}\left(u_{0} \underline{v}\right)=\operatorname{grad} u_{0} \times \underline{v}+u_{0} \operatorname{rot} \underline{v}$,
(iv) $\operatorname{grad}(\underline{u}, \underline{v})=\underline{u} \times \operatorname{rot} \underline{v}+\underline{v} \times \operatorname{rot} \underline{u}+(\underline{u}, \operatorname{grad}) \underline{v}+(\underline{v}, \operatorname{grad}) \underline{u}$,
(v) $\operatorname{rot}(\underline{u} \times \underline{v})=\underline{u} \operatorname{div} \underline{v}-\underline{v} \operatorname{div} \underline{u}+(\underline{v}, \operatorname{grad}) \underline{u}-(\underline{u}, \operatorname{grad}) \underline{v}$.

Proposition 16 (Multiplier problem). Let $u \in C^{1}(G, \mathbb{H}) \cap \operatorname{ker} D$ be an arbitrarily given function. If for all $v \in \operatorname{ker} D \cap C^{1}(G, \mathbb{H})$ also $v u \in \operatorname{ker} D \cap$ $C^{1}(G, \mathbb{H})$ then $u=$ const.

Proof. A straightforward calculation leads to the result.

Proposition 17 The class of $\mathbb{H}$-regular functions does not contain the squares of each of its elements.

Proof. Let $x=x_{1} e_{1}-x_{2} e_{2} \in \operatorname{ker} D \cap C^{1}\left(G, C l_{3,0}\right)$. Then we have

$$
x^{2}=\left(x_{1} e_{1}-x_{2} e_{2}\right)\left(x_{1} e_{1}-x_{2} e_{2}\right)=x_{1}^{2}+x_{2}^{2}
$$

Hence $D x^{2}=2\left(x_{1} \sigma_{1}+x_{2} \sigma_{2}\right) \neq 0$.

Remark 18 For a so-called totally analytic variable $z$ with $z(x) \in \operatorname{ker} D \cap$ $C^{1}\left(\mathbb{R}^{4}, \mathbb{H}\right)$ follows with necessity $z^{k} \in \operatorname{ker} D \cap C^{1}\left(R^{4}, \mathbb{H}\right)$.
We studied in [11] a generalization of an interpolation formula of LAGRANGE's type with quaternionic regular functions. The result is as follows: let $z(x):=\sum_{i=1}^{3} x_{i} d_{i}, z\left(a_{i}{ }^{(j)}\right):=\sum_{i=1}^{3} a_{i}{ }^{(j)} d_{i}, a_{i}^{(j)} \in \mathbb{H}$. We write in abbreviated form:

$$
z_{j}(x):=z(x)-z\left(a^{(j)}\right), z_{k j}:=z\left(a^{(k)}\right)-z\left(a^{(j)}\right)
$$

and demand $z_{k j} \neq 0$ for $k \neq j$. A quaternionic regular interpolation function is then given by the following formula:

$$
\left(L_{n} u\right)(x)=\sum_{k=1}^{n}\left[\frac{z_{1}(x)}{z_{k 1}} \cdot \ldots \cdot \frac{z_{k-1}(x)}{z_{k(k-1)}} \frac{z_{k+1}(x)}{z_{k(k+1)}} \cdot \ldots \cdot \frac{z_{n}(x)}{z_{k n}}\right] u_{k}
$$

with
(i) $\left(L_{n} u\right)^{k} \in \operatorname{ker} D \cap C^{1}\left(\mathbb{R}^{4}, \mathbb{H}\right)(k=1, \ldots, n)$,
(ii) $\left(L_{n} u\right)\left(a^{(j)}\right)=u_{j}(j=1, \ldots, n)$.

Proposition 19 Let $u \in \operatorname{ker} D \cap C^{1}(G, \mathbb{H}), u(x) \neq 0$ then $u^{-1} \in \operatorname{ker} D \cap$ $C^{1}(G, \mathbb{H})$ not in general.

Proof. It is sufficient to give a counterexample. Taking $u(x)=x_{1} \mathbf{e}_{1}-x_{2} e_{2} \in$ ker $D \cap C^{1}(G, \mathbb{H})$ then for $0 \notin G$ we obtain

$$
u^{-1}=\frac{x_{2} e_{2}-x_{1} e_{1}}{x_{1}^{2}+x_{2}^{2}}
$$

and one can easily verify that

$$
D u^{-1}=\frac{2\left(x_{1}^{2}-x_{2}^{2}\right)}{\left(x_{1}^{2}+x_{2}^{2}\right)^{2}} \sigma_{0}+i \frac{4 x_{1} x_{2}}{\left(x_{1}^{2}+x_{2}^{2}\right)^{2}} \sigma_{3} \neq 0 .
$$

Omission of a Leibniz rule in the classical sense leads to considerable difficulties in the construction of systems of regular functions. We will formulate some easy principles to produce quaternionic regular functions.

### 3.1.2 Quaternionic regular polynomials

Let $\left\{e_{0}, e_{1}, e_{2}, e_{3}\right\}$ be a basis in $\mathbb{H}$ and be $x=\sum_{i=0}^{3} x_{i} e_{i}$. We assume as before that $e_{0}^{2}=e_{0}, e_{i} e_{j}+e_{j} e_{i}=-2 \delta_{i j}$. We consider the new variables

$$
z_{k}:=x_{0} e_{k}-x_{k} e_{0} \quad(k=1,2,3)
$$

For $\partial=\sum_{i=0}^{3} e_{i} \partial_{i}$ one has $\partial z_{k}=0$. Unfortunately, the product of two of such variables is not quaternionic regular, e.g. $\partial z_{k} z_{l} \neq 0$ for any $k$ and $l$. Nevertheless, one can find homogeneous quaternionic regular polynomials of any degree by symmetrization. These polynomials are given by the formula

$$
P_{m}(x)=\frac{1}{m!} \sum_{\left(\mu_{1}, \ldots, \mu_{m}\right)} z_{\mu_{1}} \cdot \ldots \cdot z_{\mu_{m}},
$$

where $\left(\mu_{1}, \ldots, \mu_{m}\right)$ covers all distinguishable permutations of $\{1, \ldots, m\}$ and $\left(\mu_{1}, \ldots, \mu_{m}\right)$ is an arbitrary combination of $m$ elements out of repetitions of the set $\{1,2,3\}$. For $m=0$ we fix $P_{0}(x)=e_{0}$. This construction was first made by R. Delanghe and is also valid in real Clifford algebras.
Comprehensive results on Clifford regular polynoms can be found in [6] and also in [7].

### 3.1.3 Regular singularity functions

Moreover, we will describe an important class of quaternionic regular functions which can be used in collocation methods for boundary value problems of partial differential equations. Let $\underline{x}^{(k)}$ be a system of points outside a given domain $G \subset \mathbb{R}^{3}$. Then there exists a system of quaternionic regular functions $\left\{\phi_{k}\right\}$ which are defined by

$$
\phi_{k}(\underline{x})=\frac{\underline{x}-\underline{x}^{(k)}}{\left|\underline{x}-\underline{x}^{(k)}\right|^{3}} .
$$

Under some conditions completeness and advantageously numerical properties are possible to obtain. This is a system of singularity functions which are used in calculations.

### 3.1.4 Regular functions from harmonic functions

Let $G$ be an open set in $\mathbb{R}^{n}$ and star-shaped with respect to the origin and $u_{0}: G \longmapsto \mathbb{R}$ is harmonic in $G$ then

$$
u(x)=u_{0}(x)-\left\{\int_{0}^{1} t^{n-2}\left(D u_{0}\right)(t x) x d t-\left(\int_{0}^{1} t^{n-2}\left(D u_{0}\right)(t x) x d t\right)_{0}\right\}
$$

is $C \ell_{0, n^{-}}$regular in $G$.

### 3.2 On the spherical Dirac operator

In this part we will put together useful properties of operators which result of a suitable decomposition of the DIRAC operator. Let $x$ be the paravector $x=x_{0}+\underline{x}$ with $\underline{x}=\sum_{i=1}^{n} x_{i} e_{i}$. Setting $\omega_{i}=x_{i} /|x| \quad(i=1, \ldots, n)$ and $\omega=\sum_{i=1}^{n} w_{i} e_{i}$. Then $\underline{x}=|\underline{x}| \omega$, where $\omega^{2}=-1$. We consider the so-called CaUCHY-Fueter operator $\partial=\partial_{0}+D$. Here denotes $\partial_{0}:=\partial / \partial x_{0}$ and $D=\sum_{i=1}^{n} e_{i} \partial_{i}$ with $\partial_{i}:=\partial / \partial x_{i}$. Introducing the denotations

$$
L:=\sum_{i=1}^{n} e_{i} L_{i}(x) \quad \text { with } \quad L_{i}(x)=|\underline{x}| \partial_{i}-x_{i} \ell_{\omega}
$$

and $\ell_{\omega}:=\sum_{i=1}^{n} \omega_{i} \partial_{i}$ we obtain $D=\frac{1}{|x|} L+\omega \ell_{\omega}$.
Proposition 20 We have for $j=1, \ldots, n$
(i) $\quad \ell_{\omega} \underline{x}=\omega, \quad$ (projection onto the unit sphere)
(ii) $\ell_{\omega} \omega=0$
(iii) Let $f=f(|\underline{x}|)$, then $\quad \ell_{\omega} f=d / d_{|\underline{\mid x}|} f=: f^{\prime}$
(iv) $\quad \partial_{j} \omega_{k}=\frac{1}{|\underline{x}|}\left[\delta_{j k}-\omega_{k} \omega_{j}\right]$,
(v) $\quad L_{j} \omega_{k}=\delta_{j k}-\omega_{k} \omega_{j}=|\underline{x}| \partial_{j} \omega_{k}$.

Proof. The following straightforward calculations lead to the result:

$$
\begin{equation*}
\ell_{\omega} x_{j}=\sum_{i=1}^{n} \omega_{i} \partial_{i}\left(x_{j}\right)=\sum_{i=1}^{n} \omega_{i} \delta_{i j}=\omega_{j} \quad(j=1, \ldots, n) \tag{i}
\end{equation*}
$$

(ii) $\quad \ell_{\omega} \omega_{j}=\ell_{\omega}\left(\frac{x_{j}}{|\underline{x}|}\right)=\frac{\left(\ell_{\omega} x_{j}\right)|\underline{x}|-\ell_{\omega}|\underline{x}| x_{j}}{|\underline{x}|^{2}}=\frac{\omega_{j}|\underline{x}|-x_{j}}{|\underline{x}|^{2}}=0$

$$
(j=1, \ldots, n) .
$$

$$
\begin{equation*}
\ell_{\omega} f=\sum_{i=1}^{n} \omega \partial_{i} f=\sum_{i=1}^{n} \omega_{i} f^{\prime} \partial_{i}|\underline{x}|=\sum_{i=1}^{n} \omega_{i}^{2} f^{\prime}=f^{\prime} \tag{iii}
\end{equation*}
$$

Due to (iv) it is easy to see, that
$\partial_{i} \omega_{k}=\partial_{j} \frac{x_{k}}{|\underline{x}|}=\frac{\left(\partial_{j} x_{k}\right)|x|-\partial_{j}|\underline{x}| x_{k}}{|\underline{x}|^{2}}=\frac{\delta_{j k}|x|-\omega_{j} x_{k}}{|\underline{x}|^{2}}=\frac{1}{|\underline{x}|}\left(\delta_{j k}-\omega_{j} \omega_{k}\right)$.
Relation (v) we get from

$$
L_{j} \omega_{k}=|\underline{x}| \partial_{j} \omega_{k}-x_{j} \ell_{\omega} \omega_{k}=\delta_{j k}-\omega_{k} \omega_{j}
$$

Corollary 21 Let $f \in C^{1}(\mathbb{R})$ and $f=f(|\underline{x}|)$, then $L f=0$.
Proof. By definition we get for $j=1, \ldots, n$

$$
L_{j}(|\underline{x}|)=|\underline{x}| \partial_{j}|\underline{\mid x}|-\omega_{j} \ell_{x}|\underline{x}|=|\underline{x}| \omega_{j}-\omega_{j}|\underline{x}| \ell_{\omega}|\underline{x}|=|\underline{x}| \omega_{j}-\omega_{j}|\underline{x}|=0 .
$$

Hence, it follows

$$
L|\underline{x}|=\sum_{i=1}^{n} e_{i} L_{i}|\underline{x}|=0 .
$$

Furthermore, we obtain

$$
\begin{aligned}
L_{j} f(|\underline{x}|) & =|\underline{x}| \partial_{j} f-\omega_{j} \sum_{i=1}^{n} x_{i} \partial_{i} f=|\underline{x}| \frac{d f}{d|\underline{x}|} \partial_{j}|\underline{x}|-\omega_{j} \sum_{i=1}^{n} x_{i} \frac{d f}{d|\underline{x}|} \partial_{i}|\underline{x}| \\
& =\frac{d f}{d|\underline{x}|}\left(|\underline{x}| \partial_{j}|\underline{x}|-x_{j} \ell_{\omega}|\underline{x}|\right)=\frac{d f}{d|\underline{x}|} L_{j}|\underline{x}|=0
\end{aligned}
$$

and therefore $L f=0$.
Proposition 22 It holds
(i) $L \omega=\omega L+(1-n)$
(ii) $\omega L=-\underline{x} \wedge D=\sum_{i<j} e_{i} e_{j}\left(x_{i} \partial_{j}-x_{j} \partial_{i}\right)$.
where the operator $\omega L=: \Gamma$ is just the spherical DIRAC operator.
Proof. By definition we have

$$
\begin{aligned}
\Gamma= & \underline{x} D+|\underline{x}| e_{\omega}=-\sum_{i=1}^{n} x_{i} \partial_{i}+\sum_{i \neq j} x_{i} \partial_{j} e_{i} e_{j}+\sum_{i=1}^{n} x_{i} \partial_{i} \\
& \sum_{i<j} x_{i} \partial_{j} e_{i} e_{j}+\sum_{i>j} x_{i} \partial_{j} e_{i} e_{j}=\sum_{i<j} x_{i} \partial_{j} e_{i} e_{j}-\sum_{i>j} x_{i} \partial_{j} e_{j} e_{i} \\
= & \sum_{i<j} x_{i} \partial_{j} e_{i} e_{j}-\sum_{j>i} x_{j} \partial_{i} e_{i} e_{j}=\sum_{i<j}\left(x_{i} \partial_{j}-x_{j} \partial_{i}\right) e_{i} e_{j} .
\end{aligned}
$$

Theorem 23 [5] The following formula is valid:

$$
\Gamma \omega=(n-1) \omega
$$

Proof. By definition

$$
\begin{aligned}
\Gamma \omega & =(\omega L) \omega=\sum_{i, j, k} \omega_{i} e_{i} e_{j} L_{j}\left(\omega_{k} e_{k}\right)==\sum_{i, j, k} \omega_{i} e_{i} e_{j}\left\{\delta_{k j}-\omega_{j} \omega_{k}\right\} e_{k} \\
& =\sum_{k=j=1}^{n} \sum_{i=1}^{n} \omega_{i} e_{i}-\sum_{i, j, k} \omega_{i} \omega_{j} \omega_{k} e_{i} e_{j} e_{k}=+n \omega-\sum_{i, j, k} \omega_{i} \omega_{j} \omega_{k} e_{i} e_{j} e_{k} .
\end{aligned}
$$

Finally, we find

$$
\begin{aligned}
\sum_{i, j, k} \omega_{i} \omega_{j} \omega_{k} e_{i} e_{j} e_{k} & =\sum_{\substack{k=1 \\
i<j}}^{n}\left(\omega_{i} \omega_{j}-\omega_{j} \omega_{i}\right) e_{k} e_{i} e_{j}-\sum_{k=1}^{n} \sum_{i=j=1}^{n} \omega_{i}^{2} \omega_{k} e_{i}^{2} e_{k} \\
& =\sum_{k=1}^{n}\left(\sum_{i=1}^{n} \omega_{i}^{2} \omega_{k} e_{k}\right)=\omega .
\end{aligned}
$$

Corollary 24 (5] The DIRAC operator permits the representation

$$
D=\omega\left(\ell_{\omega}+\frac{1}{|\underline{x}|} \Gamma\right)
$$

Proof. We know that $\omega L=\underline{x} D+|\underline{x}| \ell_{\omega}$. Then

$$
\underline{x} \omega L=-|\underline{x}|^{2} D+\underline{x}|\underline{x}| \ell_{\omega}
$$

and

$$
\begin{aligned}
D & =-\frac{\underline{x}}{|\underline{x}|^{2}} \omega L+\frac{x}{|\underline{x}|} \ell_{\omega}=\frac{1}{|\underline{x}|}\left(-L+x \ell_{\omega}\right) \\
& =\omega \ell_{\omega}+\frac{w}{|x|} \omega L=\omega\left(\ell_{\omega}+\frac{1}{|\underline{x}|} \Gamma\right)
\end{aligned}
$$

### 3.3 Teodorescu transform

This part is dedicated to the consideration of a very special weakly singular integral operator, which will play a key role in the whole theory.

Let $G$ be a domain in $\mathbb{R}^{n}$ with a piecewise smooth boundary $\partial G=\Gamma$. We will use here the following norms:

$$
\|u\|_{\infty}:=\|u\|_{L_{\infty}(G)}=\operatorname{vrai} \max _{x \in G}|u(x)|,
$$

where $|\cdot|$ is just the Clifford operator norm . Furthermore, for $1 \leq p<\infty$ we define

$$
\begin{aligned}
& \|u\|_{p}:=\|u\|_{L_{p}(G)}=\left(\int_{G}|u|^{p} d x\right)^{1 / p} \\
& \|u\|_{p, \alpha}:=\left(\int_{G}\left[|u \| x|^{\alpha}\right]^{p} d x\right)^{1 / p}
\end{aligned}
$$

We abbreviate

$$
\|u\|_{W_{p}^{1}(G)}=:\|u\|_{p, 1} .
$$

By $|G|$ we denote the volume of $G$.

Definition 25 Let $u \in C(G))$. Then the linear integral operator

$$
\begin{equation*}
\left(T_{G} u\right)(x)=-\int_{G} e(x-y) u(y) d y \tag{3}
\end{equation*}
$$

with

$$
e(x)=\frac{1}{\sigma_{n}} \frac{\bar{\omega}(x)}{|x|^{n-1}}, \omega(x)=\frac{x}{|x|}, x=\underline{x}=\sum_{i=1}^{n} x_{i} e_{i}
$$

is called Teodorescu transform over $G$.
We remark that $T_{G}$ corresponds to the known $T$-operator from the one variable complex analysis. In the case $n=2$ it coincides with the complex $T$-operator up to a constant factor. First in this section we consider the existence of this integral and some elementary properties.

Proposition 26 Let $u \in C(\bar{G})$. Then the integral (3) exists for all $x \in \mathbb{R}^{n}$ and we have

$$
\left\|T_{G} u\right\|_{C} \leq \frac{1}{\sigma_{n}} \max _{x \in \bar{G}} \int_{G} \frac{1}{|x-y|^{n-1}} d y\|u\|_{C}
$$

Proof. Because of the the uniform convergence of $\int_{G} 1 /\left(|x-y|^{n-1}\right) d y$ it is continuous according to $y$ and, therefore, we can take the maximum over $\bar{G}$.

This means that $T_{G}$ is a continuous operator in $C(\bar{G})$.
Corollary 27 For $|x| \rightarrow \infty$ it follows that $\left|\left(T_{G} u\right)(x)\right| \rightarrow 0$. Furthermore, we have $\left(T_{G} u\right)(x) \in C^{\infty}\left(\mathbb{R}^{n} \backslash \bar{G}\right)$.

This property will be essential for the investigation of boundary value problems.

Proposition 28 Let $u \in L_{1}(G)$. Then $\left(T_{G} u\right)(x)$ exists almost everywhere on $\mathbb{R}^{n}$ and belongs to $L_{q}\left(G^{\prime}\right)$ for $q$ with $1<q<n /(n-1)$ and any domain $G^{\prime} \subset \mathbb{R}^{n}$.

Proof. Let $v \in L_{p}(G), p>n$. Then we get by HöLDER's inequality

$$
V(x)=\int_{G} \frac{|v(y)|}{|x-y|^{n-1}} d y \leq \int_{G} \frac{1}{|x-y|^{(n-1) q}} d y\|v\|_{p}
$$

Clearly, $(n-1) q<n$. Thus $V(x)$ is defined by a uniformly convergent integral and therefore continuously. Then $V(x) u(x) \in L_{1}(G)$. Using Fubini's theorem we obtain the identity

$$
\begin{equation*}
\int_{G}|u(y)| V(y) d y=\int_{G}|v(y)| U(y) d y \tag{4}
\end{equation*}
$$

with

$$
U(x)=\int_{G} \frac{|u(y)|}{|x-y|^{n-1}} d y
$$

Hence we conclude $U(x) \in L_{q}(G)$, where $1 / p+1 / q=1$. In this way we get $\left(T_{G} u\right)(x) \in L_{q}(G)$.

Now, we will prove continuity results of $T_{G}$ in the scale of Sobolev spaces.
Proposition 29 Let $u \in C_{0}\left(\mathbb{R}^{n}\right)$ then

$$
\partial_{k}\left(T_{G} u\right)(x)=-\frac{1}{\sigma_{n}} \int_{G} \partial_{k, x} e(x-y) u(y) d y+\bar{e}_{k} \frac{u(x)}{n} .
$$

Proof. For the proof we will refer to our book ([11]).

Theorem 30 The operator

$$
\partial_{k} T: L_{p}\left(\mathbb{R}^{n}\right) \rightarrow L_{p}\left(R^{n}\right)
$$

is continuous and

$$
\left\|\partial_{k} T_{G}\right\|_{L\left(L_{p}\left(R^{n}\right)\right)} \leq\left(C \sigma_{n}^{-\frac{1}{p}}+\frac{1}{n}\right) \quad(1<p<\infty)
$$

and

$$
T_{G}: L_{p}(G) \rightarrow W_{p}^{1}(G)
$$

is continuous.
By the help of Sobolev embedding theorems it is easy to obtain the following results on the continuity of $T_{G}$ between other important spaces.

## Corollary 31

(i) $T_{G}:\left[L_{p}(G), C^{0, \frac{p-n}{p}}(G)\right]$ is continuous for $p>n$.
(ii) Let $u \in L_{p}(G), 1<p<n$. Then for all $r<\frac{n p}{n-q}$ we have that $T_{G}: L_{p}(G) \rightarrow L_{r}(G)$ is compact. For all $\varepsilon>0$ there exists $\delta>0$ with

$$
\left\|T_{G} u(x)-T_{G} u\left(x^{\prime}\right)\right\|_{r} \leq \varepsilon\|u\|_{P}
$$

for $\left|x-x^{\prime}\right|<\delta$. Especially, we have that $T_{G}$ is compact in $L_{2}(G)$.
Remark 32 Also in the case of unbounded domains, continuity properties of the operator $T_{G}$ can be proved. Unfortunately, the proofs require additional expertise. We will only formulate some results and refer to the literature. The Teodorescu transform $T_{G}$ is in the case of $p<n$ a continuous mapping from $L_{p}(G)$ into $L_{p}^{-1}(G)$. For

$$
1-\frac{n}{p}<\alpha<n\left(1-\frac{1}{p}\right)
$$

the transform $T_{G}$ is continuous from $L_{p}^{\alpha}(G)$ into $L_{p}^{\alpha-1}(G)$. Corresponding considerations for different kinds of unbounded domains can be found in [14].

### 3.4 Borel-Pompeiu's formula and its consequences

In this part we will concern ourselves with $\mathrm{C} \ell_{0, \mathrm{n}}$-versions of classical theorems of the function theory of one complex variable. In the centre of the considerations will be the generalization of BOREL-POMPEIU's formula.

Proposition 33 Let $u \in C^{1}(G) \cap C(\bar{G}$,$) . Then$

$$
\int_{G}(D u)(y) d y=\int_{\Gamma} n(y) u(y) d \Gamma_{y}
$$

where $n(y)=\sum_{i=1}^{n} n_{i}(y) e_{i}$ is the unit vector of the outward pointing normal at $y$.

Proof. Using Gauss' formula we obtain

$$
\begin{aligned}
\int_{G}(D u)(y) d y & =\sum_{i=1}^{n} \sum_{A} e_{i} e_{A} \int_{G} \partial_{i} u_{A}(y) d y \\
& =\sum_{i=1}^{n} \sum_{A} e_{i} e_{A} \int_{\Gamma} n_{i}(y) u_{A}(y) d \Gamma_{y} \\
& =\int_{\Gamma} n(y) u(y) d \Gamma_{y}
\end{aligned}
$$

Corollary 34 (CAUCHY's integral theorem.)
Let $u \in C^{1}(G) \cap C(\bar{G})$ and $u \in \operatorname{ker} D(D u=0!)$ then

$$
\int_{\gamma} n(y) u(y) d \Gamma_{y}=0
$$

where $\gamma$ is an arbitrary surface with $\gamma \subset G$.
Let

$$
E(x)=\frac{1}{\sigma_{n}} \frac{1}{(2-n)}|x|^{-(n-2)} \quad(n>2) \quad \text { with } \quad \sigma_{n}=\frac{2 \pi^{n / 2}}{\Gamma\left(\frac{n}{2}\right)}
$$

then we obtain $\Delta E(x)=0(x \neq 0)\left(\sigma_{n}\right.$ is again the surface area of the unit sphere in $\mathbb{R}^{n}$ ). It can be easily seen that the DIRAC operator $D=\sum_{i=1}^{n} e_{i} \partial_{i}$ yields

$$
e(x)=\bar{D} E(x)=\frac{-x}{\sigma_{n}|x|^{n}}
$$

Proposition 35 Let $u \in C^{2}(G) \cap C(\bar{G})$. Then

$$
\left(D T_{G} u\right)(x)=\left\{\begin{array}{cl}
u(x) & \text { in } G \\
0 & \text { in } \mathbb{R}^{n} \backslash \bar{G}
\end{array}\right.
$$

Proof.Firstly, we note that in case of $\partial_{i j}=\partial_{j i}$

$$
\begin{aligned}
D \bar{D} & =\sum_{i=1}^{n} \sum_{j=1}^{n} e_{i} \bar{e}_{j} \partial_{i j}=-\sum_{i=1}^{n} \sum_{j=1}^{n} e_{i} e_{j} \partial_{i j} \\
& =-\sum_{i=j} e_{i} e_{j} \partial_{i j}-\sum_{i>j}\left(e_{i} e_{j}-e_{j} e_{i}\right) \partial_{i j}=\Delta .
\end{aligned}
$$

Furthermore, we have for $x \in G$

$$
\begin{aligned}
\left(D T_{G} u\right)(x) & =D\left(-\int_{G} e(x-y) u(y) d y\right)=D\left(-\int_{G} \bar{D} E(x-y) u(y) d y\right) \\
& =\Delta\left(-\int_{G} \frac{1}{\sigma_{n}(2-n)} \frac{1}{|x-y|^{n-2}} u(y) d y\right)
\end{aligned}
$$

The latter integral is just the volume potential which solves the Poisson equation. Hence, $\left(D T_{G} u\right)(x)=u(x)$.
For $\Delta E(x)=0$ in $\mathbb{R}^{n} \backslash \bar{G}$ it follows our assertion.
Theorem 36 Let $G \subset \mathbb{R}^{n}$ be a domain which is bounded by a piecewise Liapunov surface $\Gamma$. Then for each $u \in C^{1}(G) \cap C(\bar{G})$

$$
\int_{\Gamma} e(x-y) n(y) u(y) d \Gamma_{y}-\int_{G} e(x-y)(D u)(y) d y=\left\{\begin{array}{cl}
u(x) & , \quad x \in \frac{G}{G}  \tag{5}\\
0, & x \notin, ~
\end{array}\right.
$$

where $n(y)$ denotes again the unit vector of the outward pointing normal at $\Gamma$ in point $y$.

Proof. Let $x$ be an interior point of $G, B_{\varepsilon}(x)$ a ball centered at $x$ with the radius $\varepsilon(\varepsilon>0)$. Further, let $G_{\varepsilon}:=G \backslash B_{\varepsilon}(x)$. Applying Proposition 35 and choosing for $u=e(x)$ then it follows that

$$
\begin{align*}
& \int_{G_{\epsilon}} e(x-y)(D v)(y) d y= \\
& =\int_{\Gamma} e(x-y) n(y) v(y) d \Gamma_{y}-\int_{S_{\varepsilon}} e(x-y) n(y) v(y) d S_{\varepsilon_{y}} . \tag{6}
\end{align*}
$$

On $S_{\varepsilon}\left(y \in S_{\varepsilon}\right)$ holds

$$
e\left(x-y n(y)=\frac{1}{\sigma_{n}} \frac{\bar{x}-\bar{y}}{|x-y|^{n}} \frac{y-x}{|x-y|}=\frac{1}{\sigma_{n}} \frac{1}{|x-y|^{n-1}} .\right.
$$

It remains to show that the latter integral converges to $v(x)$ if $\varepsilon$ tends to zero. We have

$$
\begin{equation*}
\lim _{\varepsilon \rightarrow 0} \int_{S_{\varepsilon}} e(x-y) n(y) v(y) d y=\lim _{\varepsilon \rightarrow 0}\left(\frac{1}{\sigma_{n} \varepsilon^{n-1}} \int_{S_{\epsilon}(x)} v(y) d y\right)=v(x) . \tag{7}
\end{equation*}
$$

For $\varepsilon \rightarrow 0$ the first integral in (8) exists and Borel-Pompeiu's formula for interior points is shown. For the second part we have to repeat the proof without excluding a neighbourhood of $x$.

Remark 37 Using Stokes' Theorem 36 can be proved on compact $n$ dimensional, oriented $C^{\infty}$-manifolds which are lying in a domain of $\mathbb{R}^{n}$ (cf. [8], [5]).

Definition 38 Let $u \in C^{1}(G) \cap C(\bar{G})$. The operator $F_{\Gamma}$ defined by

$$
\left(F_{\Gamma} u\right)(x):=\int_{\Gamma} e(x-y) n(y) u(y) d \Gamma_{y}
$$

is called Cauchy-Bizadse operator.
Remark 39 In this new notation formula (5) has the representation

$$
\left(F_{\Gamma} u\right)(x)+\left(T_{G} D u\right)(x)=\left\{\begin{array}{cl}
u(x) & , x \in G \\
0 & , x \in \mathbb{R}^{n} \backslash \bar{G}
\end{array}\right\}
$$

Proof. (Theorem 36) Let $x$ be an interior point of $G, B_{\varepsilon}(x)$ a ball centered at $x$ with the radius $\varepsilon(\varepsilon>0)$. Further, let $G_{\varepsilon}:=G \backslash B_{\varepsilon}(x)$. Applying Proposition 35 and choosing for $u=e(x)$ then it follows that

$$
\begin{align*}
& \int_{G_{\varepsilon}} e(x-y)(D v)(y) d y= \\
& =\int_{\Gamma} e(x-y) n(y) v(y) d \Gamma_{y}-\int_{S_{\varepsilon}} e(x-y) n(y) v(y) d S_{\varepsilon_{v}} \tag{8}
\end{align*}
$$

On $S_{\varepsilon}\left(y \in S_{\varepsilon}\right)$ holds

$$
e(x-y) n(y)=\frac{1}{\sigma_{n}} \frac{\bar{x}-\bar{y}}{|x-y|^{n}} \frac{y-x}{|x-y|}=\frac{1}{\sigma_{n}} \frac{1}{|x-y|^{n-1}}
$$

It remains to show that the latter integral converges to $v(x)$ if $\varepsilon$ tends to zero. We have

$$
\lim _{\varepsilon \rightarrow 0} \int_{S_{\varepsilon}} e(x-y) n(y) v(y) d y=\lim _{\varepsilon \rightarrow 0}\left(\frac{1}{\sigma_{n} \varepsilon^{n-1}} \int_{S_{\varepsilon}(x)} v(y) d y\right)=v(x)
$$

For $\varepsilon \rightarrow 0$ the first integral in (8) exists and Borel-Pompeiu's formula for interior points is shown. For the second part we have to repeat the proof without excluding a neighbourhood of $x$.

Remark 40 Borel-Pompeiu's formula is also valid in the case of multiply connected domains. Suppose $G_{i}(i=0,1, \ldots, n)$ be simply connected bounded domains in $\mathbb{R}^{n}$ with smooth boundaries $\Gamma_{i}(i=0,1, \ldots, n)$ which fulfil the relations
(i) $G_{0} \supset \bigcup_{i=1}^{n} \bar{G}_{i}=K$ and (ii) $\bar{G}_{i} \cap \bar{G}_{j}=\emptyset \quad(i \neq j)$. Then

$$
\left(F_{\Gamma_{0}} u\right)(x)-\sum_{i=1}^{n}\left(F_{\Gamma_{i}} u\right)(x)+\left(T_{G_{0} \backslash K} D u\right)(x)=\left\{\begin{array}{cll}
u(x) & \text { in } & G_{0} \backslash K \\
0 & \text { in } & \mathbb{R}^{n} \backslash\left[G_{0} \backslash K\right]
\end{array}\right\}
$$

for every function in $C^{1}\left(G_{0} \backslash K\right) \cap C(\overline{G \backslash K})$.
Now it is easy to transfer many of the results of the theory of one complex variable which are connected with CaUCHY's integral theorem.

Theorem 41 (CAUCHY's integral formula). Let $G \subset \mathbb{R}^{n}$ be a domain with a piecewise LiApunov boundary. Furthermore, let $u \in \operatorname{ker} D$, then

$$
\left(F_{\Gamma} u\right)(x)=\left\{\begin{array}{cll}
u(x) & \text { in } & G  \tag{9}\\
0 & \text { in } & \mathbb{R}^{n} \backslash \bar{G}
\end{array}\right\}
$$

holds.
Proof. Because of $D u=0$ formula (9) holds.
Corollary 42 Let $u \in \operatorname{ker} D$. Then $u$ has partial derivatives $\partial_{i_{1} i_{2} \ldots i_{k}}$ ( $i_{l} \in\{1, \ldots, n\} ; l \in\{1, \ldots, k\}$ ) of any order.

Proof. The kernel function $e(x-y)$ has for $x \neq y$ partial derivatives of any order.

Corollary 43 (Mean-value theorem). Let $B_{r}(x) \subset \mathbb{R}^{n}$ be a ball centred at $x$ of radius $r$. For any $u \in \operatorname{ker} D$

$$
u(x)=\frac{n}{\sigma_{n} r^{n}} \int_{B_{r}(x)} u(y) d y
$$

Note that $\sigma_{n} r^{n} / n$ is just the volume of the ball $B_{r}(x)$.
Proof. The proof is an immediate consequence of Cauchy's formula.
Corollary 44 (Mean-value formula). Let $B_{r}(x)$ again be a ball centred at $x$ of radius $r$. Then

$$
u(x)=\frac{1}{\sigma_{n} r^{n-1}} \int_{S_{r}(x)} u(y) d S_{r_{v}}
$$

where $u \in(\operatorname{ker} D)\left(B_{r}(x)\right) \cap C\left(\bar{B}_{r}(x)\right)$.
Proof. Applying CAUCHY's integral formula then

$$
u(x)=\frac{1}{\sigma_{n}} \int_{S_{r}(x)} \frac{\bar{y}-\bar{x}}{|y-x|^{n}} \frac{y-x}{|y-x|} u(y) d S_{r_{y}}=\frac{1}{\sigma_{n} r^{n-1}} \int_{S_{r}(x)} u(y) d S_{r_{v}} .
$$

Theorem 45 (Maximum modulus theorem). Let $G \subset \mathbb{R}^{n}$ be a connected bounded domain and $u \in(\operatorname{ker} D)(G) \cap C(\bar{G})$ with $\bar{G}=G \cup \Gamma, \Gamma=\partial G$. If there exists some $z \in G$ with

$$
|u(x)| \leq|u(z)| \quad(\forall x \in G)
$$

it is necessary and sufficient that $u$ is constant.
Proof. For $z \in G$ there exists a ball $B_{\varepsilon}(z) \subset G$. Applying the mean value formula we obtain

$$
\begin{equation*}
u(z)=\frac{1}{\sigma_{n} \varepsilon^{n-1}} \int_{S_{\varepsilon}} u(x) d S_{\varepsilon, x} \tag{10}
\end{equation*}
$$

Assume now that $|u(x)| \leq|u(z)|$ and $u$ is not constant in $B_{\varepsilon}(z)$. Then there exists a decomposition of $S_{\varepsilon}$ into two sets

$$
S_{\varepsilon}^{\prime}=\left\{x \in S_{\varepsilon}:|u(x)|=|u(z)|\right\}, \quad S_{\varepsilon}^{\prime \prime}=\left\{x \in S_{\varepsilon}:|u(x)|<|u(z)|\right\} .
$$

Formula (10) yields

$$
|u(z)|<\frac{1}{\sigma_{n} \varepsilon^{n-1}}\left[\int_{S_{\varepsilon}^{\prime}}|u(x)| d S_{\varepsilon, x}^{\prime}+\int_{S_{\varepsilon}^{\prime \prime}}|u(z)| d S_{\varepsilon, x}^{\prime \prime}\right]=\frac{\left|S_{\varepsilon}^{\prime}\right|+\left|S_{\varepsilon}^{\prime \prime}\right|}{\sigma_{n} \varepsilon^{n-1}}|u(z)|=|u(z)| .
$$

This is a contradiction. Hence on $S_{\varepsilon}$ we have $|u(x)|=|u(z)|$.
Choosing now smaller balls $B_{\varepsilon^{\prime}}(z)\left(\varepsilon^{\prime}<\varepsilon\right)$ we obtain the same. Therefore, $|u(x)|=|u(z)|$ in the whole ball $B_{\varepsilon}(z)$. Since $G$ is connected we get immediately $|u(x)|=$ const on $G$ and for the continuity of $|u(x)|$ such also on $\bar{G}$. It remains to show that $u(x)$ is also a constant. From $|u|^{2}=$ const. we conclude

$$
0=D|u|^{2}=D \Delta\left(\sum_{A} u_{A}^{2}\right)=2 \sum_{A} \sum_{j=1}^{n}\left(\partial_{j} u_{A}\right) u_{A} e_{j}
$$

and also

$$
0=\Delta|u|^{2}=\bar{D} D|u|^{2}=2 \sum_{A} \sum_{i=1}^{n} \sum_{j=1}^{n}\left[\left(\partial_{i j} u_{A}\right) u_{A} e_{i} e_{j}+\left(\partial_{j} u_{A}\right)\left(\partial_{i} u_{A}\right) e_{i} e_{j}\right] .
$$

Since the components of a CLIFFORD left regular function are harmonic one gets

$$
0=\sum_{A} \sum_{i=1}^{n}\left(\partial_{i} u_{A}\right)^{2}
$$

and consequently $\partial_{i} u_{A}=0(i=1, \ldots, n ; A)$.
Corollary 46 Let $G \subset \mathbb{R}^{n}$ be a connected and bounded domain and $u \in$ $(\operatorname{ker} D)(G) \cap C(\bar{G})$ then

$$
\sup _{x \in \bar{G}}|u(x)|=\sup _{x \in \partial G}|u(x)| .
$$

Theorem 47 (Liouville's theorem). Let $u \in(\operatorname{ker} D)\left(\mathbb{R}^{n}\right)$. If $|u(x)| \leq$ $M\left(x \in \mathbb{R}^{n}\right)$ then $u$ is a constant.

Proof. The proof is left to the reader.
Theorem 48 (MORERA's theorem). Let $u \in C^{1}(G), D u \in L_{r}(G), r>n$. If for all balls $B_{r}(x)(r>0, x \in G)$

$$
\int_{B_{r}(x)} n(y) u(y) d \gamma(y)=0
$$

then $D u=0$ in $G$.
Proof. Let $x \in G$ be an arbitrary point $\left(B_{r_{k}}\right) k \in N$ a regular sequence of balls which is contracting to $x \in G$. We have $x \in \bigcap_{k} B_{r_{k}}$. For the validity of LEBESGUE's theorem we obtain for each $v \in L_{r}(G)(r>n)$,

$$
\lim _{k \rightarrow \infty} \frac{1}{\left|B_{r_{k}}\right|} \int_{B_{r_{k}}} v(y) d B_{r_{k}}(y)=: \bar{v}(x)
$$

and $v=\tilde{v}$ in $L_{\mathrm{r}}(G)$. Substituting $v=D u$ it follows by Proposition 33 that

$$
\int_{B_{r_{k}}}(D u)(y) d y=\int_{S_{r_{k}}} n(y) u(y) d S_{r_{k, v}}
$$

and also

$$
\frac{1}{\left|B_{r_{k}}\right|} \int_{B_{r_{k}}}(D u)(y) d y=0 \quad(k=0,1,2, \ldots)
$$

and, therefore, $(D u)(x)=0$ almost everywhere on $G$.
Theorem 49 Let $u \in(\operatorname{ker} D)\left(\mathbb{R}^{n} \backslash \bar{G}\right) \cap C\left(\mathbb{R}^{n} \backslash \bar{G}\right)$. Suppose $\lim _{|x| \rightarrow \infty} u(x)=u(\infty)$. Then

$$
\left(F_{\Gamma} u\right)(x)= \begin{cases}u(\infty)-u(x) & , x \in \mathbb{R}^{n} \backslash \bar{G} \\ u(\infty) & , x \in G\end{cases}
$$

Proof. Let $B_{R}(x)$ be such a ball that $B_{R}(x) \supset G$. Write briefly $s=\left(\omega_{x}, \omega_{y}\right)=(x /|x|, y /|y|)$. On account of

$$
\frac{1}{|x-y|^{n-1}}=\left(\sum_{k=0}^{\infty} C_{k}^{\frac{n-3}{2}}(s)\left|\frac{x}{y}\right|^{k}\right) \frac{1}{|y|^{n-1}} \quad(|x|<|y|)
$$

and CaUCHY's integral formula for multiply connected domains we get

$$
\begin{aligned}
& \frac{1}{\sigma_{n}} \int_{S_{R}} \frac{x-y}{|x-y|^{n}} n(y) u(y) d S_{R, y}-\frac{1}{\sigma_{n}} \int_{\Gamma} \frac{x-y}{|x-y|^{n}} n(y) u(y) d \Gamma_{y} \\
&= \begin{cases}0 & , x \in \mathbb{R}^{n} \backslash \overline{\left(B_{R} \backslash \bar{G}\right)} \\
u(x) & , x \in B_{R} \backslash \bar{G}\end{cases}
\end{aligned}
$$

The first integral transforms into

$$
\begin{aligned}
& \frac{1}{\sigma_{n}} \int_{S_{R}} \frac{x-y}{|x-y|^{n}} \frac{y-x}{|y-x|} u(y) d S_{R_{y}}=\frac{1}{\sigma_{n}} \int_{S_{R}} \frac{1}{|x-y|^{n-1}} u(y) d S_{R_{y}}= \\
& =\sum_{k=0}^{\infty} \frac{1}{\sigma_{n}} \int_{S^{n}} \frac{C_{k}^{\frac{n-3}{2}}(s)}{R^{k}} u(y) d S_{y}^{n}|x|^{k} .
\end{aligned}
$$

For $R \rightarrow \infty$ the right-hand side converges to

$$
\frac{C_{0}^{\frac{n-3}{2}}(s)}{\sigma_{n}} \int_{S^{n}} u(\infty) d S_{y}^{n}=u(\infty) \quad\left(C_{0}^{\mu}(s) \equiv 1\right)
$$

which provides the result.

### 3.5 Plemelj-Sokhotzkij-type projections

The operator $P_{\Gamma}:=\frac{1}{2}\left(I+S_{\Gamma}^{\delta}\right)$ denotes the Plemelj-Sokhotzkij-type projection onto the space of all $\mathbb{C} \ell_{0, n}$-valued functions which may be $C \ell_{0, n}$-regular extended into the domain $G . Q_{\Gamma}:=\frac{1}{2}\left(I-S_{\Gamma}^{\delta}\right)$ denotes the Plemelj-Sokhotzkitype projection onto the space of all $C \ell_{0, n}$-valued functions which can be $\mathbb{C} \ell_{0, n}$-regular extended into the domain $R^{n} \backslash \bar{G}$ and vanish at infinity. The operator

$$
\left.\left(S_{\Gamma}^{\delta} u\right):=\frac{4 \pi-2 \delta(x)}{4 \pi} u(x)+\left(S_{\Gamma} u\right)(x) \quad 0<\delta \leq 4 \pi\right)
$$

where $\delta$ is the space angle taken from outside at the point $x$. It is easy to see that

$$
\left(S_{\Gamma}^{\delta}\right)^{2}=I
$$

Note that Plemelj-Sokhotzkij-type projections coincide in case of the unit ball with the well-known Szegö projections. We have the following statements :

Proposition 50 [11] Let $u \in C^{1}(G) \cap C(\bar{G})$. Then we have the formulas
(i) $\left(F_{\mathrm{r}} u\right)(x)+T_{G} D u(x)=\left\{\begin{array}{cl}u(x) & , x \in G \text { (Borel-Pompeiu formula) } \\ 0 & , x \in R^{n} \backslash \bar{G}\end{array}\right\}$
(ii) $\left(D T_{G} u\right)(x)=\left\{\begin{array}{ll}u(x) & , x \in G \\ 0 & , x \in R^{n} \backslash \bar{G}\end{array}\right\}$
(iii) $\left(D F_{\mathrm{\Gamma}}\right) u(x)=0$ in $G \cup\left(R^{n} \backslash \bar{G}\right)$

Theorem 51 [11](Plemelj-Sokhotzkij's formulas). Let $u \in C^{0, \alpha}(G), 0<$ $\alpha<1$. Then we have

$$
\text { (i) } \left.\lim _{\substack{-\epsilon \in \in \mathrm{F} \\ z \in C}} F_{\Gamma} u\right)(x)=\left(P_{\Gamma} u\right)(\xi)
$$

(ii) $\left.\lim _{\substack{x \rightarrow \in \mathbb{R} \\ x \in \mathbb{R}^{-} \mid G}} F_{\Gamma} u\right)(x)=\left(-Q_{\Gamma} u\right)(\xi)$
for any $\xi \in \Gamma$.

Corollary 52 [11] Let $u \in C^{0, \alpha}(\Gamma)$. Then the relations
(i) $\left(S_{\Gamma}^{2} u\right)(\xi)=u(\xi)$
(ii) $\quad\left(F_{\Gamma} P_{\Gamma} u\right)(\xi)=F_{\Gamma} u(\xi)$
(iii) $\quad\left(P_{\Gamma}^{2} u\right)(\xi)=\left(P_{\Gamma} u\right)(\xi)$
(iv) $\left(Q_{\Gamma}^{2} u\right)(\xi)=\left(Q_{\Gamma} u\right)(\xi)$
are valid for any $\xi \in \Gamma$.

### 3.6 Hilbert space decomposition

Let us now consider the Hilbert space $L_{2}(G)$ with an inner product $(u, v)=$ $\int_{G} \bar{u} v d x \in C l_{0, n}$.

Theorem 53 The Hilbert space $L_{2}(G)$ allows an orthogonal decomposition

$$
L_{2}(G)=\left[\operatorname{ker} D \cap L_{2}(G)\right] \oplus D\left[{\left.\stackrel{\circ}{W_{2}^{1}}(G)\right]}^{\circ}\right.
$$

Proof. The right-linear sets $X_{1}=L_{2}(G) \cap \operatorname{ker} D$ and $X_{2}=L_{2}(G) \backslash X_{1}$ are subspaces of $L_{2}(G)$. Let $u \in X_{2}$. Then it follows $v=T_{G} u \in W_{2}^{1}(G)$ and $u$ has a representation $u=D v, v \in W_{2}^{1}(G)$. From $u \in X_{2}$ we have $\int_{G} \overline{D v} g d x=0$ for all $g \in X_{1}$ and in particular $\int_{G} \overline{D v} e\left(x-y_{l}\right) d x=0$ for all numbers $l \in N$, where $y_{l} \in \Gamma_{A}, \Gamma_{A}=\partial G_{A}$, and $\operatorname{clos} G \subset G_{A}$. We assume now that the set $\left\{u_{l}\right\} \in \Gamma_{A}$, is dense in $\Gamma_{A}$. Integration by parts leads to $\int_{\Gamma} e\left(y_{l}-x\right) n(x) v(x) d \Gamma_{x}=\left(F_{\Gamma} t r_{\Gamma} v\right)\left(y_{l}\right)=0$ for all $l \in N$, hence, $F_{\Gamma} t r_{\Gamma} v=0$ in $G^{-}$and $\lim _{\substack{v \rightarrow G^{-} \\ v \in G^{-}}}\left(F_{\Gamma} t r_{\Gamma} v\right)(y)=0$. Consequently, we have $t r_{\Gamma} v \in \operatorname{im} P_{\Gamma} \cap$ $W_{2}^{1 / 2}(G)$. Using theorems about traces we obtain the existence of a function $V \in \operatorname{ker} D \cap W_{2}^{1}(G): \operatorname{tr}_{\Gamma} V=t r_{\Gamma} v$. If we consider now $v-V \in \dot{W}_{2}^{1}(G)$, then the application of $D$ shows that $D(v-V)=D v=u$ and we have $u \in D\left[\stackrel{\circ}{W}_{2}^{1}(G)\right]$. This result means that $\left(\left(\operatorname{ker} D \cap L_{2}(G)\right)^{\perp} \subseteq D\left[\stackrel{\circ}{W}_{2}^{1}(G)\right]\right.$. If we suppose now that $w \in D\left[\dot{W}_{2}^{1}(G)\right]$ we conclude as follows:

$$
w \in D\left[\stackrel{\circ}{W}_{2}^{1}(G)\right] \Rightarrow w=D z, z \in \stackrel{\circ}{W}_{2}^{1}(G) \Rightarrow
$$

$$
\int_{G} \bar{u} w d x=\int_{G} \bar{u} D z d x=\int_{G} \overline{D u} z d x=0
$$

for all $u \in \operatorname{ker} D$.
Then we have $D\left[\stackrel{\circ}{W}_{2}^{1}(G)\right] \subseteq\left[\operatorname{ker} D \cap L_{2}(G)\right]^{\perp}$.

Corollary 54 There exist two orthoprojections $\mathbb{P}_{G}$ and $Q_{G}$ with

$$
\begin{aligned}
& \mathbb{P}_{G}: L_{2}(G) \rightarrow \operatorname{ker} D \cap L_{2}(G) \\
& Q_{G}: L_{2}(G) \rightarrow D\left[\dot{\circ}_{2}^{1}(G)\right], Q_{G}=I-\mathbb{P}_{G}
\end{aligned}
$$

Remark 55 If we use the operator $R u=\sum_{A} r_{A} u_{A} e_{A}$ with positive real numbers $r_{A}$ and the inner product $[u, v]_{R}=\int_{G} \overline{R^{-1} u} R^{-1} v d G$ then Theorem 53 can be generalized
in the following way:

$$
L_{2}(G)=\left[R \operatorname{ker} D \cap L_{2}(G)\right] \oplus_{R} D\left[\stackrel{\circ}{W_{2}^{1}}(G)\right]
$$

We can prove this decomposition using the same methods as in the proof of Theorem 53.

Corollary 56 Let $f \in L_{2}(G),\left(T_{G} f\right)(x)=0$ for $x \in \mathbb{R}^{n} \backslash \operatorname{clos} G \Rightarrow f \in$ $i m Q_{G}$.

Proof. We use the representation $f=D g$ with $g=T_{G} f$. From the assumption it follows that $\operatorname{tr}_{\Gamma} g=0$ and hence, $f=D g \in D\left[W_{2}^{1}(G)\right]$.

The last corollary enables us to formulate a theorem concerning the completeness of $\left\{e\left(x-y_{l}\right)\right\}$ and to prove the theorem without using HahnBanach's Theorem.

Theorem 57 Under the above mentioned conditions for $\left\{y_{l}\right\}_{l \in \boldsymbol{N}} \subset \Gamma_{A}$ the system $\left\{e\left(x-y_{l}\right)\right\}_{l \in \boldsymbol{N}}$ is complete in $L_{2}(G) \cap$ ker $D$.

Proof.As usual, we assume the existence of $u \in \operatorname{ker} D \cap L_{2}(G)=i m \mathbb{P}_{G}$ with property $\left(u, e\left(\cdot-y_{l}\right)\right)=0 \quad \forall l \in \mathbb{N}$. Note that $\operatorname{im} \mathbb{P}_{G}$ is the image of the operator $\mathbb{P}_{G}$. Then we have $\left(T_{G} u\right)(x)=0$ for $\forall x \in \Gamma_{A}$, because $T_{G} u \in C^{\infty}\left(\mathbb{R}^{n} \backslash \operatorname{clos} G\right)$. Furthermore, using $T_{G} u \in \operatorname{ker} \Delta\left(\mathbb{R}^{n} \backslash \operatorname{clos} G\right)$, $\left(T_{G} u\right)(x) \underset{|x| \rightarrow \infty}{\longrightarrow} 0$ we arrive at $\left(T_{G} u\right)(x)=0$ for $x \in \mathbb{R}^{n} \backslash$ clos $G$. The previous corollary implies that $u \in \operatorname{im} Q_{G}$. Hence $u \in \operatorname{im} \mathbb{P}_{G} \cap i m Q_{G}=$ $\{0\}$. This proves the theorem.

## 4 Applications

### 4.1 A new sort of elementary functions

Let us start with the introduction of an exponential function. For this section we assume that $x$ is a paravector in $C \ell_{0, n}$, e.g. $x \in P V$ Vec $C \ell_{0, n} \subset$ $C \ell_{0, n}$. Each paravector permits the representation $x=x_{0}+\underline{x}$ with $\underline{x}=$ $\sum_{i=1}^{n} x_{i} e_{i}$. Furthermore, we define a function $\omega: \mathbb{R}^{n} \rightarrow S^{n}$ given by $\omega(x):=$ $\underline{x} /|\underline{x}|$. We will use for our definition a normally convergent series expansion. For an arbitrary $\varepsilon>0$ it is always possible to find a sufficiently large number $N$ such that for any $r, s>N$ holds

$$
\left|\sum_{k=r}^{s} \frac{x^{k}}{k!}\right| \leq \sum_{k=r}^{s} \frac{K^{k}|x|^{k}}{k!}<\varepsilon,
$$

where $K$ is a constant which only depends on $n$ and satisfies the inequality $|x y| \leq K^{2}|x||y|$. Note that $|x|^{2}=x \bar{x}=x_{0}^{2}+x_{1}^{2}+\ldots+x_{n}^{2}$.

Definition 1 Let $x$ be a paravector in $C \ell_{0, n}$. The exponential function $e^{x}$ is defined by the power series

$$
\begin{equation*}
\sum_{k=0}^{\infty} \frac{x^{k}}{k!} \tag{11}
\end{equation*}
$$

Proposition 2 The exponential function permits the representation

$$
e^{x}=e^{x_{0}}(\cos |\underline{x}|+\omega(x) \sin |\underline{x}|)
$$

Proof. Using the CaUchy product of two power series we get

$$
\begin{gathered}
\sum_{k=0}^{\infty} \frac{x^{k}}{k!}=\sum_{k=0}^{\infty} \frac{\left(x_{0}+\underline{x}\right)^{k}}{k!}=\sum_{k=0}^{\infty} \frac{1}{k!} \sum_{\ell=0}^{k}\binom{k}{\ell} x_{0}^{\ell} \underline{x}^{k-\ell} \\
=\sum_{k=0}^{\infty} \sum_{\ell=0}^{k} \frac{x_{0}^{\ell} \underline{x}^{k-\ell}}{\ell!(k-\ell)!}=\sum_{\ell=0}^{\infty} \frac{x_{0}^{\ell}}{\ell!} \sum_{m=0}^{\infty} \frac{\underline{x}^{m}}{m!}=e^{x_{0}} e^{x} .
\end{gathered}
$$

It remains to consider $e^{\underline{x}}$. We have

$$
\begin{aligned}
& \sum_{k=0}^{\infty} \frac{\underline{x}^{k}}{k!}=\sum_{\ell=0}^{\infty} \frac{\underline{x}^{2 \ell}}{(2 \ell)!}+\sum_{\ell=0}^{\infty} \frac{\underline{x}^{2 \ell+1}}{(2 \ell+1)!} \\
= & \sum_{\ell=0}^{\infty}(-1)^{\ell} \frac{|\underline{x}|^{2 \ell}}{(2 \ell)!}+\omega(x) \sum_{\ell=0}^{\infty}(-1)^{\ell} \frac{|\underline{x}|^{2 \ell+1}}{(2 \ell+1)!}=\cos |\underline{x}|+\omega(x) \sin |\underline{x}| .
\end{aligned}
$$

Corollary 3 Let $x, y$ be paravectors in $C \ell_{0, n}$. If $x y=y x$ we get

$$
\begin{equation*}
e^{x+y}=e^{x} e^{y} \tag{12}
\end{equation*}
$$

Corollary 4 For any paravector $x \in C \ell_{0, n}$ we have
(i) $e^{x} \neq 0$,
(ii) $e^{-x} e^{x}=1$,
(iii) $e^{\omega(x) \pi}=-1$,
(iv) $e^{k \underline{z}}=\left(e^{x}\right)^{k} \quad(k \in \mathbb{N})$ (Morvre's formula), (v) $\left|e^{x}\right|=e^{x_{0}}$.

Using an idea in [2] we obtain:
Corollary 5 Let $x$ be a paravector in $C \ell_{0, n}$. Then the exponential function may be described by the limit

$$
e^{x}=\lim _{m \rightarrow \infty}\left(1+\frac{x}{m}\right)^{m}
$$

Proof. Consider the difference

$$
e^{x}-\left(1+\frac{x}{m}\right)^{m}=\sum_{k=0}^{\infty}\left(\frac{1}{k!}-\frac{\binom{k}{m}}{m^{k}}\right) x^{k}
$$

Since $\left(1+\frac{x}{m}\right)^{m}$ is a polynomial this series is normally convergent. The same is valid for the series $e^{x}$. The coefficients in front of $x^{k}$ are positive. Now we obtain

$$
\left|e^{x}-\left(1+\frac{x}{m}\right)^{m}\right| \leq \sum_{k=0}^{\infty}\left(\frac{1}{k!}-\frac{\binom{k}{m}}{m^{k}}\right)|K x|^{k}=e^{|K x|}-\left(1+\frac{|K x|}{m}\right)^{m}
$$

which tends to zero for $m$ to infinity.
Definition 6 Let $x$ be a paravector in $C \ell_{0, n}$. Then hyperbolic and trigonometric functions are defined by the following formulae:

$$
\sinh x:=\frac{e^{x}-e^{-x}}{2}, \quad \cosh x:=\frac{e^{x}+e^{-x}}{2} .
$$

For $|\underline{x}| \neq 0$ we define:

$$
\sin x:=\frac{e^{x \omega(x)}-e^{-x \omega(x)}}{2} \omega(x), \quad \cos x:=\frac{e^{x \omega(x)}+e^{-x \omega(x)}}{2} .
$$

Corollary 7 It holds

$$
\sin x=-\sin (-x) \quad \text { and } \quad \cos x=\cos (-x)
$$

Proposition 8 The hyperbolic functions $\sinh x$ and $\cosh x$ and trigonometric functions $\sin x$ and $\cos x$ permit the following representations:

$$
\begin{aligned}
& \cosh x=\omega(x) \sinh x_{0} \sin |\underline{x}|+\cosh x_{0} \cos |\underline{x}|, \\
& \sinh x=\omega(x) \cosh x_{0} \sin |\underline{x}|+\sinh x_{0} \cos |\underline{x}| .
\end{aligned}
$$

and

$$
\begin{aligned}
& \sin x=\omega(x) \sinh |\underline{x}| \cos x_{0}+\cosh |\underline{x}| \sin x_{0}, \\
& \cos x=\omega(x) \sinh |\underline{x}| \sin x_{0}+\cosh |\underline{x}| \cos x_{0} .
\end{aligned}
$$

The right-hand sides of these representations can be used for the definition of $\sin$ and $\cos$ in the case of $|\underline{x}|=0$.

Definition 9 Let $x \in C \ell_{0, n}$ be a paravector, e.g. $x=x_{0}+\underline{x}, x \neq x_{0}<0$. Then a paravector valued logarithm $\log x$ is defined by

$$
\log x:=\ln |x|+\omega(x) \arccos \frac{x_{0}}{|x|}, \quad|\underline{x}| \neq 0 \quad \text { or }|\underline{x}|=0, x_{0}>0 .
$$

Corollary 10 The above defined logarithm $\log x$ has the following properties:
(i) $\log 1=0$ and $\log e_{i}=\frac{\pi}{2}(i=1, \ldots, n)$,
(ii) $1-\frac{1}{|x|}-\arctan \frac{|\underline{x}|}{\left|x_{0}\right|} \leq|\log x| \leq|x|-1+\arctan \frac{|\underline{x}|}{\left|x_{0}\right|}$.

Theorem 11 The function $\log x$ is the inverse to the exponential function $e^{x}$ introduced above, which means

$$
e^{\log x}=x \quad \text { and } \quad \log e^{x}=x
$$

Proof. The proof is to realize by a straightforward calculation.
Corollary 12 Let $x y=y x$ then the well-known logarithm rule

$$
\log (x y)=\log x+\log y
$$

is valid.
Definition 13 Let $\alpha$ be a real number. The general power function $x^{\alpha}$ is defined by

$$
x^{\alpha}:=e^{\alpha \log x} .
$$

One example will confirm this definition.
Example 14 Let $x=\underline{x}$ and $\alpha=\frac{1}{n}$. We have to line out the following calculation:

$$
\begin{aligned}
\underline{x}^{\frac{1}{n}} & =e^{\frac{1}{n}\left(\ln |\underline{x}|+\omega(x) \arccos \frac{x}{x \mid}+\omega(x) 2 k \pi\right)} \\
& =\sqrt[n]{|\underline{x}|}\left[\cos \left(\frac{1}{n}(\arccos 0+2 k \pi)\right)+\omega(x) \sin \left(\frac{1}{n}(\arcsin 0+2 k \pi)\right)\right] \\
& =\sqrt[n]{|\underline{x}|}\left[\cos \left(\frac{\pi}{2 n}+\frac{2 k \pi}{n}\right)+\omega(x) \sin \left(\frac{\pi}{2 n}+\frac{2 k \pi}{n}\right)\right]
\end{aligned}
$$

for $k=0,1, \ldots, n-1$.

### 4.1.1 Radially differentiable functions and FUETER's mapping

The elementary functions introduced above are not Clifford regular. This is caused by the operator $\Gamma$ in the following representation of the CauchyFueter operator

$$
\partial=\partial_{0}+\omega(x) \partial_{|\underline{x}|}+\omega(x) \frac{1}{|\underline{x}|} \Gamma .
$$

Thus it holds as already proved $\Gamma \omega=(n-1) \omega$. This expression is disturbing the structure of our elementary functions. By the way, we note that scalar and vector fields which are only depending on $|\underline{x}|$ are lying in the kernel of $\Gamma$. It seems to be useful to introduce the following operator:

Definition 15 The differential operator

$$
D_{r a}:=\frac{1}{2}\left(\partial_{0}-\omega(x) \partial_{|x|}\right)
$$

is called radial CAUCHY-FUETER operator.

Note the denotation $D_{r a} u:=u^{\prime}$ for a paravector $u \in C \ell_{0, n}$. Let us now investigate in which manner this operator is acting on the elementary functions.

Proposition 16 We list here the following differential actions:
(i) $\partial_{|\underline{x}|} \omega(x)=0, \quad$ (ii) $\partial_{|x|}|x|=\frac{|\underline{x}|}{|x|}, \quad$ (iii) $\partial_{|x|} \mid x=\omega(x)$.

Proof. For $u \in C^{2}\left(\mathbb{R}^{n}, C \ell_{0, n}\right)$ we have for instance

$$
\partial_{|x|} \omega_{i}(x)=\partial_{|\underline{x}|} \partial_{i}|\underline{x}|=\partial_{i} \partial_{|x|}|\underline{x}|=\partial_{i} 1=0,
$$

where $\omega_{i}(x)=x_{i} /|\underline{x}|$. We note only for (iii) that $\underline{x}=(\underline{x} /|\underline{x}|)|\underline{x}|$. Property (ii) is easy to verify.

Proposition 17 Let $u, v$ be paravector valued functions. If $\omega(x) u=u \omega(x)$ then for $D_{\text {ra }}$ the Leibniz rule holds.

Theorem 18 The radial differential operator $D_{\text {ra }}$ generates in the set of elementary functions the following rules:
(i) $\left(e^{x}\right)^{\prime}=e^{x}$,
(ii) $(\sin x)^{\prime}=\cos x, \quad(\cos x)^{\prime}=-\sin x$,
(iii) $(\sinh x)^{\prime}=\cosh x, \quad(\cosh x)^{\prime}=\sinh x$,
(iv) $(\log x)^{\prime}=\frac{1}{x}$,
(v) $\left(x^{\alpha}\right)^{\prime}=\alpha x^{\alpha-1}$ where $\alpha \in \mathbb{R}$.

Proof. All properties follow straightforward.
Definition 19 Let $f: P \operatorname{Vec} C \ell_{0,1} \mapsto P \operatorname{Vec} C \ell_{0,1}$ be a paravector valued function in $C \ell_{0, n}$ of the type $f=f_{0}+\omega(x) f_{1}$ with $f_{0}, f_{1} \in \mathbb{R}$ and $h=$ $h_{0}+\omega(x)|\underline{h}|$. Such a function is called radially differentiable or radially CLIFFORD regular if and only if

$$
\lim _{h \rightarrow 0} \frac{(f(x+h)-f(x)) \bar{h}}{|h|^{2}}
$$

exists. In the case of existence the radial derivative is just $f^{\prime}(x)=D_{r a}$.
Theorem 20 Let $f_{i}=f_{i}\left(x_{0},|\underline{x}|\right),(i=0,1)$ real-valued functions and $f=f_{0}+\omega(x) f_{1}$. Further, there exist continuous partial derivatives $\partial_{0} f_{1}$ and $\partial_{z \mid} f_{i}(i=0,1)$. Then the function $f$ is radially differentiable if and only if the relations

$$
\partial_{0} f_{0}=\partial_{|x|} f_{1}, \quad \partial_{0} f_{1}=-\partial_{|z|} f_{0}
$$

are valid.
Corollary 21 All above defined elementary functions are radially ClifFORD regular.

Now we will demonstrate how to transform radially regular functions to Clifford regular functions. The basic idea goes back to R. Fueter, who formulated a transformation of holomorphic functions to quaternionic regular functions. Later M. Sce [22] and T. QIAN [20] generalized these results. We will illustrate this mapping in $C \ell_{0, n}$. Let $z:=u+i v \in \mathbb{C}$ and $h:=$ $u+\omega v \in C \ell_{0, n}$. Introducing by $\tau$ the mapping:

$$
q=\tau(z)=\Delta^{(n-1) / 2} h .
$$

Proposition 22 [20] Let $h=u\left(x_{0},|\underline{x}|\right)+\omega v\left(x_{0},|\underline{x}|\right)$ be a function radially regular in an open set of $\mathbb{R}^{n+1}$. Then for any number $k \in \mathbb{N}$ we find

$$
\Delta^{k} h=u_{k}\left(x_{0},|\underline{x}|\right)+\omega v_{k}\left(x_{0},|\underline{x}|\right)
$$

where

$$
u_{k}=2 k\left(\partial_{|\underline{x}|}\right) u_{k-1} \frac{1}{|\underline{x}|}, \quad v_{k}=2 k\left(\partial_{|\underline{x \mid}|} \frac{1}{|\underline{x}|}-\frac{v_{k-1}}{|\underline{x}|^{2}}\right)=2 k \partial_{|\underline{\mid x}|} \frac{v_{k-1}}{|\underline{x}|} .
$$

Proof. The proof given in [20] follows by mathematical induction.
Theorem 23 Let $n=2$. A radially quaternionic regular function fulfils the following differential equation

$$
|\underline{x}|^{2} \Delta h-|\underline{x}| \partial \underline{x} \mid h+\omega \operatorname{Vec} h=0
$$

Proof. The action of the operator $\Delta$ on $h$.

$$
\Delta h=\frac{x}{|\underline{x}|} \partial_{|\underline{x}|} u\left(|\underline{x}|, x_{0}\right)+\omega \partial_{|\underline{x}|}\left(\frac{v\left(|\underline{x}|, x_{0}\right)}{|\underline{x}|}\right) .
$$

Further we obtain

$$
\begin{aligned}
& \Delta h=\frac{1}{|\underline{x}|} \partial_{\mid} \underline{x}\left|u+\omega \partial_{\mid} \underline{x}\right|\left(\frac{v}{|\underline{x}|}\right)=\frac{1}{|\underline{x}|}\left[\partial \underline{x} \left\lvert\, u+\omega \partial_{|\underline{x}| v]}-\omega \frac{v}{|\underline{x}|^{2}}\right.\right. \\
& \left.\Delta h=\frac{1}{|\underline{x}|} \partial_{\mid \underline{x}} \right\rvert\, h-\omega \frac{\operatorname{Vec} h}{|\underline{x}|^{2}} .
\end{aligned}
$$

Hence

$$
\left[|\underline{x}|^{2} \Delta-|\underline{x}| \partial|\underline{x}|+\omega \operatorname{Vec}\right] h=0 .
$$

Corollary 24 The Laplacian realizes the mapping

$$
\Delta: \operatorname{ker} \partial_{|z| a} \longmapsto \operatorname{ker} \partial,
$$

where $\partial_{\mathrm{ra}}=\partial_{0}+D_{\mathrm{ra}}$.

Remark 25 For an arbitrary $n$ we have to substitute the operator $\Delta$ by the operator $\Delta^{(n-1) / 2}$. In this way we have for odd $n$ that this operator is just a pointwise differential operator and for even $n$ we have to consider the FOURIER multiplier operator induced by the symbol

$$
(2 \pi i|\xi|)^{n-1}
$$

We will now restrict our considerations to the case of real quaternions.
Theorem 26 FUETER's mapping delivers us the following quaternionic regular elementary functions.

1. $\exp x:=\left(+\frac{\sin |x|}{|x|}+\omega \frac{-|x| \cos |x|+\sin |x|}{|x|^{2}}\right) e^{x_{0}}$
2. $\sin x:=\sin x_{0} \frac{\sin h|x|}{|x|}+\omega\left(\frac{|x| \cos h|x|-\sin h|x|}{|x|^{2}}\right) \cos x_{0}$
3. $\cos x:=\cos x_{0} \frac{\sin h|x|}{|x|}-\omega\left(\frac{-\sin h|x|+|x| \cos h|x|}{|x|^{2}}\right) \sin x_{0}$
4. $\cosh x:=+\cosh x_{0} \frac{\sin |x|}{|x|}+\omega \frac{-\sin |x|+|x| \cos |x|}{|x|^{2}} \sinh x_{0}$.
5. $\sinh x:=+\sinh x_{0} \frac{\sin |x|}{|x|}+\omega \frac{-\sin |x|+|x| \cos |x|}{|x|^{2}} \cosh x_{0}$

Proof. Using the paper ([20]) the proof is an exercise.

Remark 27 Fueter's method also generates a "logarithm" which we will denote by $\log x$. After a straightforward computation we find:

$$
\log x=\frac{1}{\left(r^{2}+x_{0}^{2}\right)}-\omega\left(\frac{1}{r\left(r^{2}+x_{0}^{2}\right)}+\frac{1}{r^{2}} \arccos \frac{x_{0}}{\sqrt{r^{2}+x_{0}^{2}}}\right) .
$$

Unfortunately log is not the inverse of exp.

Now we will consider some elementary properties of the regular exponential function $\exp x$. Setting

$$
\bar{\partial}:=\left(\frac{\partial_{0}+\bar{D}}{2}\right) .
$$

Theorem 28 The exponential function has the property:

$$
\bar{\partial} \exp \lambda x=\lambda \exp \lambda x, \quad \lambda \in \mathbb{C}
$$

Proof. We obtain $D=\frac{1}{|x|} L+\omega \ell_{\omega}$. As for definition

$$
D \exp \lambda x=\left(\frac{1}{|\underline{x}|} L+\omega \ell_{\omega}\right) \exp \lambda x
$$

using $L f(|\underline{x}|)=0$ (for any real function $f!$ ) and $L \omega=2$ then we get.

$$
\begin{aligned}
D \exp \lambda x= & \left(\frac{1}{|x|} L+\omega \ell_{\omega}\right)\left(\frac{\sin \lambda|x|}{\lambda|x|}+\omega\left(\frac{\sin \lambda|x|-\lambda|x| \cos \lambda|x|}{\lambda^{2}|x|^{2}}\right)\right) \\
= & \frac{2}{|x|}\left(\frac{\sin \lambda|x|-\lambda| | x|\cos \lambda| x \mid}{\lambda^{2}|x|^{2}}\right)+\omega\left(\frac{\lambda \cos \lambda|x|(\lambda|x|)-\lambda \sin \lambda|x|}{(\lambda|\underline{x}|)^{2}}\right) \\
& \frac{\left.\left.|\lambda \cos \lambda| x\left|+\lambda^{2}\right| x|\sin \lambda| x|-\lambda \cos \lambda| x\left|\lambda^{2}\right| x\right|^{2}-2 \lambda^{2}|x|(\sin \lambda|x|-\lambda|x| \cos \lambda|x|)\right]}{(\lambda \mid x)^{4}} \\
= & 2\left(\frac{\sin \lambda|x|-\lambda|x| \cos \lambda|x|}{\left.\lambda^{2}|x|\right|^{3}}\right)-\frac{\lambda \cos \lambda|x|}{(\lambda|\underline{x}|)^{2}}-\frac{\lambda^{2}|x| \sin \lambda|x|}{(\lambda|x|)^{2}}+\frac{\lambda \cos \lambda|x|}{(\lambda|x|)^{2}}- \\
& -\frac{2 \lambda^{2}|x|}{(\lambda|x|)^{4}}[\sin \lambda|\underline{x}|-\lambda|\underline{x}| \cos \lambda|\underline{x}|]+\omega\left(\frac{\lambda \cos \lambda|x|(\lambda|x|)-\lambda \sin \lambda|x|}{(\lambda|\underline{x}|)^{2}}\right) \\
= & 2\left[\frac{\sin \lambda|x|}{\left.\lambda^{2}|x|\right|^{3}}-\frac{\sin \lambda|x|}{\lambda^{2}|x|^{3}}\right]+2\left[\frac{\lambda^{3}|x|^{2}}{(\lambda|x|)^{4}} \cos \lambda|\underline{x}|-\frac{\lambda|x|}{\lambda^{2}|x|^{3}} \cos \lambda|\underline{x}|\right]+\omega[\ldots] \\
= & -\lambda\left[\frac{\sin \lambda|x|}{\lambda|x|}\right]+\omega[\ldots]=-\lambda \exp \lambda x
\end{aligned}
$$

Corollary 29 It holds
(i) $|\exp x|^{2}>0$, (ii) $\lim _{x \rightarrow 0}|\exp x|=e^{x_{0}}$.

### 4.2 Selected boundary value problems in fluid mechanics

We assume for the moment that the domain $G$ is bounded by a pieceweise smooth bounded Liapunov surface. In the last years these assumptions could be considerable weakened.

### 4.2.1 Linear equations of Stokes' type

$$
\begin{aligned}
-\Delta \underline{u}+\frac{1}{\eta} \nabla p & =\frac{\rho}{\eta} \underline{f} \text { in } G \\
\operatorname{div} \underline{u} & =f_{0} \text { in } G \\
\underline{u} & =\underline{g} \text { on } \Gamma .
\end{aligned}
$$

Here $\eta$ is the viscosity and $\rho$ the density of the fluid. We have to look for the velocity $u$ and the hydrostatical pressure $p$. Between $f_{0}$ and $\underline{g}$ we have to fulfil the relation:

$$
\int_{G} f_{0} d x=\int_{\Gamma} n \underline{g} d \Gamma .
$$

For $g=0$ then the measure of the compressibility $f_{0}$ satisfies the identity

$$
\int_{G} f_{0} d x=0
$$

For all such real functions $f_{0}$ can be represented the unique solution ( $p$ is unique up to a real constant) as follows:

Theorem 30 [12] Let $f:=f_{0}+\underline{f} \in W_{p}^{k}(G, \mathbb{H})(k \geq 0,1<p<\infty)$. Then we have

$$
\begin{aligned}
& \underline{u}=\frac{\rho}{\eta} T_{G} \operatorname{Vec} T_{G} \underline{f}-\frac{\rho}{\eta} T_{G} \operatorname{Vec} F_{\Gamma}\left(\operatorname{tr}_{\Gamma} T_{G} \operatorname{Vec} F_{\Gamma}\right)^{-1} \operatorname{tr}_{\Gamma} T_{G} \operatorname{Vec} T_{G} \underline{f}-T_{G} f_{0} \\
& p=\rho \operatorname{Sc} T_{G} \underline{f}-\rho \operatorname{Sc} F_{\Gamma}\left(\operatorname{tr}_{\Gamma} T_{G} \operatorname{Vec} F_{\Gamma}\right)^{-1} \operatorname{tr}_{\Gamma} T_{G} \operatorname{Vec} T_{G} \underline{f}+\eta f_{0}
\end{aligned}
$$

In that way we strongly separated velocity and pressure.

### 4.2.2 Nonlinear equations of Stokes' type

Now we assume that the compressibility depends on the velocity and the nonlinear outer forces. The equations describing this state are the following:

$$
\begin{align*}
-\Delta \underline{u}+\frac{1}{\eta} \nabla_{p} & =\Lambda f(u) \text { in } G  \tag{13}\\
\operatorname{div}\left(\eta^{-1} \underline{u}\right) & =0 \text { in } G \\
\underline{u} & =0 \text { on } \Gamma
\end{align*}
$$

The parameter of viscosity $\eta(\eta>0)$ depends on the position. The main result is given by:

## Theorem 31 [15]

1. Let $f \in L_{2}(G, \mathbb{H}), p \in W_{2}^{1}(G), \eta \in C^{\infty}(G)$. Then every solution of the system (13) permits the integral representation

$$
\begin{align*}
& u=\Lambda R B f-R B D p  \tag{14}\\
& 0=S c \Lambda Q T_{G} B f-S c \mathcal{Q} T_{G} B D p
\end{align*}
$$

Here $B$ is the multiplication operator by $\eta^{-1}$ and $R:=T_{G} Q T_{G}$.
2. If the operator function $f(u)$ satisfies the estimates
(i) $\|f(u)-f(v)\|_{2} \leq L\|u-v\|_{2,1}$, for $\|u\|_{2,1},\|v\|_{2,1} \leq 1$,
(ii) $\|B\|_{2} \leq K$ for positive constants $K, L$
(iii) $\Lambda<\left\{\|T\|_{\text {imen } L_{2}, L_{2}}\|T\|_{L_{2}, L_{2}} K L\right\}^{-1}$
the iteration procedure

$$
\begin{aligned}
& u_{n}=\Lambda R B f\left(u_{n-1}\right)-R B D p_{n} \\
& \Lambda S c D B R f\left(u_{n-1}\right)=\operatorname{ScDBRDp_{n}} \\
& \left\|u_{0}\right\|_{2,1} \leq 1 \quad\left(u_{0} \in \stackrel{\circ}{W}_{2}^{1}(G, \mathbb{H})\right)
\end{aligned}
$$

is converging to a unique solution $\{u, p\} \in \dot{W}_{2}^{1}(G, \mathbb{H}) \cap \operatorname{ker}(\operatorname{div} B) \times$ $L_{2}(G)$ of (14), where $p$ is unique up to a real constant.

### 4.2.3 Problems of Navier-Stokes type

In the stationary case Navier-Stokes equations are described in the following way:

$$
\begin{align*}
-\Delta \underline{u}+\frac{\rho}{\eta}(\underline{u} \cdot \nabla) \underline{u}+\frac{1}{\eta} \nabla p & =\frac{\rho}{\eta} f \text { in } G  \tag{15}\\
\operatorname{div} \underline{u} & =0 \text { in } G \\
\underline{u} & =0 \text { on } \Gamma .
\end{align*}
$$

We will abbreviate with $M(u):=M^{*}(u)-f$, where $M^{*}(u)=(\underline{u} \cdot \operatorname{grad}) \underline{u}$. The main result is now the following:

Theorem 32 [9][11]

1. Let $f \in L_{2}(G), p \in W_{2}^{1}(G)$. Every solution of (15) permits the operator integral representation

$$
\begin{align*}
& u=-\frac{\rho}{\eta} R M(u)-\frac{1}{\eta} T_{G} \mathcal{Q} p  \tag{16}\\
& \frac{\rho}{\eta} S c \mathcal{Q} T_{G} M(u)-\frac{1}{\eta} S c \mathcal{Q} p=0 .
\end{align*}
$$

2. The system (16) has a unique solution $\{u, p\} \in \dot{W}_{2}^{1}(G, \mathbb{H}) \cap \operatorname{ker}(\operatorname{div} B) \times$ $L_{2}(G)$, where $p$ is unique up to a real constant, if
(i) $\|f\|_{p} \leq\left(18 K^{2} C_{1}\right)^{-1}$

$$
\text { with } \quad K:=\frac{\rho}{\eta}\left\|T_{G}\right\|_{\left[L_{2} \text { nim } \mathcal{Q}, W_{2}^{1}\right]}\|T\|_{\left\{L_{p}, L_{2}\right]}
$$

(ii) $u_{0} \in \stackrel{\circ}{W}_{2}^{1}(G, H) \cap \operatorname{ker}(\operatorname{div} B)$

$$
\text { with }\left\|u_{0}\right\|_{2,1} \leq \min \left(V, \frac{1}{4 K C_{1}}+W\right)
$$

holds. Here means $V:=\left(2 K C_{1}\right)^{-1}, W:=\left[\left(4 K C_{1}\right)^{-2}-\frac{p\|f\|_{p}}{\eta C_{1}}\right]^{\frac{1}{2}}$ and $C_{1}=9^{\frac{1}{p}} C$, where $C$ is the embedding constant from $W_{2}^{1}$ in $L_{2}$. The iteration procedure (starting with $u_{0}$ )

$$
\begin{aligned}
& u_{n}=\frac{\rho}{\eta} R M\left(u_{n-1}\right)-\frac{1}{\eta} R D p_{n} \\
& \frac{\rho}{\eta} S c Q T_{G} M\left(u_{n-1}\right)=-\frac{1}{\eta} S c Q p_{n} \\
& \left(u_{0} \in \stackrel{\circ}{W}_{2}^{1}(G, \mathbb{H}) \cap \text { kerdiv }\right)
\end{aligned}
$$

converges in $W_{2}^{1}(G, \mathbb{H}) \times L_{2}(G)$.

### 4.2.4 Navier-Stokes equations with heat conduction

We will now consider the flow of a viscous fluid under the influence of temperature. The corresponding equations read as follows:

$$
-\Delta \underline{u}+\frac{\rho}{\eta}(u \cdot \nabla) \underline{u}+\frac{1}{\eta} \nabla_{p}+\frac{\gamma}{\eta} \underline{\underline{\gamma}} w=f \quad \text { in } \quad G
$$

$$
\begin{array}{rlrl}
-\nabla w+\frac{m}{\kappa}(u \cdot \nabla) w & =\frac{1}{\kappa} h & & \text { in } \\
\operatorname{div} \underline{u} & =0 & & \text { in } \\
\underline{u} & G \\
\underline{u} & =0 & & \text { on } \\
w & =0 & & \text { on }
\end{array}
$$

We denote by $\rho$ the density of the fluid, by $\eta$ the viscosity, by $\gamma$ the Grashof number, by $\kappa$ the temperatur conductivity and by $m$ the Prandl number. As usual $\underline{u}$ stands for the velocity, $w$ for the temperature and $p$ for the hydrostatic pressure. We will here only formulate the main result. It can be shown that the solutions of this system fulfil the following system of operator integral equations, where boundary conditions will be fulfilled automatically. Here we have this system:

$$
\begin{align*}
& u=-R\left[M(u)-\frac{\gamma}{\eta} e_{3} w\right]-\frac{1}{\eta} T_{G} \mathcal{Q} p  \tag{17}\\
& 0=S c D R\left[M(u)-\frac{\gamma}{\eta} e_{3} w\right]-\frac{1}{\eta} \mathcal{Q}_{p} \\
& w=-\frac{m}{\kappa} R S c(u D) w+R g
\end{align*}
$$

where $M(u):=\frac{\rho}{\eta}(\underline{u} \cdot \operatorname{grad}) u+f(u)-F$.
Theorem 33 [13]

1. We consider the following iteration procedure:

$$
\begin{align*}
& u_{n}=-R\left[M\left(u_{n-1}\right)-\frac{\gamma}{\eta} e_{3} w_{n-1}\right]-\frac{1}{\eta} T_{G} Q p_{n}  \tag{18}\\
& 0=S c D R\left[M\left(u_{n-1}\right)-\frac{\gamma}{\eta} e_{3} w_{n-1}\right]-\frac{1}{\eta} Q p_{n} \\
& w_{n}=-\frac{m}{\kappa} R S c\left(u_{n} D\right) w_{n}+R g
\end{align*}
$$

The computation of $w_{n}$ will be done by the inner iteration:

$$
w_{n}^{j}=\frac{m}{\kappa} R S c\left(u_{n} D\right) w_{n}^{j-1}+R g .
$$

2. Let $u_{n} \in \stackrel{\circ}{W}_{2}^{1}$. Further, let $m \neq 4 \kappa$ and $u_{n} \|<\kappa / m K C$. The sequence $\left\{w_{n}^{(\mathcal{)}}\right\}_{j \in N}$ converges in $W_{2}^{1}(G)$.
3. Let $F \in L_{2}(G, \mathbb{H}), g \in L_{2}(G), f: W_{2}^{1}(G, \mathbb{H}) \rightarrow L_{2}(G, \mathbb{H})$ with $\| f(u)-$ $f(v)\left\|_{2} \leq L\right\| u-v \|_{2,1}$ and $f(0)=0$. Under the additional smallness conditions
(i) $\frac{\rho}{\eta}\|F\|_{2}+\frac{\gamma}{\eta} K|d|^{-1}\|g\|_{2}<\frac{1}{16 K^{2} C} \quad\left(d:=\left(4-\frac{m}{\kappa}\right) \kappa\right)$
(ii) $\|g\|_{2}<\left(1-\frac{1}{\sqrt{2}}\right) \eta d^{2}\left(\frac{1}{32 K^{3} C m}\right)$
(iii) $m<4 \kappa$
the sequence $\left\{u_{n}, w_{n}, p_{n}\right\}_{(n \in \mathrm{~N}\}}$ converges in $W_{2}^{1} \times W_{2}^{1} \times L_{2}$ to the unique solution $(u, w, p) \in \stackrel{\circ}{W}_{2}^{1}(G, \mathbb{H}) \times \stackrel{\circ}{W}_{2}^{1}(G, \mathbb{H}) \times L_{2}(G)$ of the originally boundary value problem, where $p$ is unique up to a real constant.

Remark 34 We note that conditions (i) and (ii) can always be realized for fluids with big enough viscosity number.

### 4.3 Eigenvalues and Teodorescu transform

### 4.3.1 On the first eigenvalue of Dirichlet's problem

We are going to establish connections between suitable norms of the Teodorescu transform and the first eigenvalue of Dirichlet's problem. Assume that $G$ is a bounded domain with a boundary $\Gamma$ which satisfies weak smoothness conditions for instance the cone property. Then the Dirac operator

$$
D: \stackrel{\circ}{W_{2}^{1}}(G) \rightarrow i m \mathcal{Q}
$$

is invertible and

$$
D^{-1}:=T_{G}: i m \mathcal{Q} \rightarrow \stackrel{\circ}{W_{2}^{1}}(G)
$$

We obtain in this way the following estimates:

$$
\begin{array}{r}
\left\|T_{G} v\right\|_{2} \leq c\|v\|_{2} \quad(v \in \mathcal{Q}) \\
\|D u\|_{2} \geq c^{-1}\|u\|_{2} \quad\left(u \in \stackrel{\circ}{W}_{2}^{1}(G)\right)
\end{array}
$$

For the smallest eigenvalue $\lambda_{1}$ of Dirichlet's problem we get

Hence

$$
\|D u\|_{2}^{2} \geq \lambda_{1}\|u\|_{2}^{2}
$$

With $u:=T_{G} v$ and $v \in \operatorname{imQ}$ it follows

$$
\left\|T_{G} v\right\|_{2}^{2} \leq \frac{1}{\lambda_{1}}\|v\|_{2}^{2}
$$

and $\left\|T_{G}\right\|_{\left[m \mathcal{Q}, W_{2}^{1}\right]} \leq \lambda_{1}^{-1 / 2}$. Assume now that $u_{1}$ is the corresponding eigenfunction to the eigenvalue $\lambda_{1}$. We obtain

$$
\left\|T_{G} w\right\|_{2}^{2}=\left\|T_{G} D u_{1}\right\|_{2}^{2}=\left\|u_{1}\right\|_{2}^{2}=\left\|D u_{1}\right\|_{2}^{2}=\|w\|_{2}^{2} \frac{1}{\lambda_{1}}
$$

with $w=D u_{1}$ and $\left(u_{1} \in \stackrel{\circ}{W_{2}^{1}}(G)\right)$. As a consequence, we have

$$
\lambda_{1}=\left\|T_{G}\right\|_{\left[i m Q, L_{2} \mid\right.}^{-2} .
$$

In a similar way we find

$$
\left\|T_{G}\right\|_{\left[i m \mathcal{Q}, \dot{W}_{2}^{1}(G) \mid\right.}^{2}=\frac{\lambda_{1}+1}{\lambda_{1}} .
$$

Because of $\left\|T_{G}\right\|_{\left[L_{2}, L_{2}\right]} \geq\left\|T_{G}\right\|_{\left[\text {imQ }, L_{2}\right]}$ we get, applying Schmidt's inequality,

$$
\left\|T_{G}\right\|_{\left|L_{2}, L_{2}\right|} \leq \sqrt[n]{\frac{|G|}{\left|B_{1}\right|}}
$$

Here $B_{1}$ denotes the unit ball in $\mathbb{R}^{n}$ and $\left|B_{1}\right|,|G|$ denotes the volume of the unit ball and the domain $G$, respectively. Thus we have obtained under these very weak conditions the domain $G$ the following lower estimate for $\lambda_{1}$ :

$$
\lambda_{1} \geq \sqrt[n]{\left(\frac{\left|B_{1}\right|}{|G|}\right)^{2}}
$$

Remark 35 Under additional geometrical and smoothness conditions there exist better lower estimates. Lower bounds for eigenvalues were first established in 1996 by E. Trefftz/F.A. Willers, stimulated by problems in engineering. Essential contributions were given by J. Hersch (1960), W.A. Kondratiev (1967), J. Cheeger (1970), L. Payne/I. Stakgold (1973), W.K. Hayman (1978), M. Taylor (1979), R. Ossermann (1979), C. Bandle (1980), C.B. Croke (1981), M.H. Protter (1981), R. Klötzler (1983), J.V. Jegorov/W.A. Kondratiev (1984) and so on.

### 4.3.2 First eigenvalue of Neumann's problem

By modification of the Teodorescu transform we will get a lower estimation of the first eigenvalue of Neumann's problem

$$
\begin{aligned}
-\Delta u=\lambda u & \text { in } \quad G \\
\partial_{n} u=0 & \text { on } \quad \Gamma .
\end{aligned}
$$

For this purpose, we introduce the factor space

$$
\widetilde{W}_{2}^{1}(G):=\left\{u \in W_{2}^{1}(G): \int_{G} u(x) d G=0\right\} / \operatorname{ker} \cdot D
$$

and the modified Teodorescu transform

$$
\widetilde{T} u:=T u-\frac{1}{|G|} \int_{G}(T u)(x) d G \quad\left(u \in L_{2}(G)\right)
$$

Note that

$$
D: \widetilde{W}_{2}^{1}(G) \xrightarrow{o n} L_{2}(G) .
$$

The following properties can be proved:
(i) $\operatorname{im} D\left(\widetilde{W}_{2}^{1}(G)\right)=L_{2}(G)$
(ii) ker $D \cap \widetilde{W}_{2}^{1}(G)=\{0\}$
(iii) $\operatorname{ker} \widetilde{T} \cap L_{2}(G)=\{0\}$
(iv) $\operatorname{im} \widetilde{T}\left(L_{2}(G)\right)=\widetilde{W}_{2}^{1}(G)$

Hence, it follows immediately that the inverse of the operator $D$ as a map between this pair of Banach spaces is given by

$$
D^{-1}:=\widetilde{T}
$$

Using the same technique as before, we prove by consideration of the operator $\widetilde{T}$ (cf. [13]) the estimate

$$
\lambda_{1}^{N} \geq \frac{1}{4} \sqrt[n]{\left(\frac{\left|B_{1}\right|}{|G|}\right)^{2}}
$$

where $\lambda_{1}^{N}$ denotes the first eigenvalue of Neumann's problem.

### 4.4 An eigenvalue problem of Stokes type

Let $f: W_{2}^{1}(G) \rightarrow V$ ec $L_{2}(G)$ be a non-linear operator which fulfils a Lipschitz condition

$$
\|f(u)-f(v)\|_{2} \leq L\|u-v\|, \quad(u, v) \in\left\{u \in W_{2}^{1}(G):\|u\| \leq 1\right\} .
$$

We are going to consider the following non-linear Stokes' eigenvalue problem :

$$
\begin{array}{r}
-\Delta \underline{u}+\eta^{-1} \operatorname{grad} p=\Lambda f(\underline{u}) \\
\text { in } \quad G \\
\operatorname{div} \underline{u}=0 \\
\text { in } \quad G \\
\underline{u}=0
\end{array} \quad \text { on } \Gamma .
$$

The real number $\Lambda$ is called eigenvalue parameter. $\eta=\eta(x)$ describes the viscosity function. After transformation in a corresponding quaternionic language we get the equivalent operator-integral formulation

$$
\begin{align*}
u+\eta^{-1} T_{G} Q p & =\Lambda T_{G} \mathcal{Q} T_{G} f(u)  \tag{19}\\
\eta^{-1} S c Q p & =\Lambda S c Q T_{G} f(u) \tag{20}
\end{align*}
$$

where $u=u_{0}+\underline{u}$ and $f(\underline{u})=f(u)$.

Definition 36 If for problem formulated above the parameter $\Lambda$ there exist non-zero functions $u$ and $p$ which solve the problem formulated above then $\Lambda$ is called a point of the spectrum $\sigma$ of our problem. If $u\|u\|_{2,1} \leq r$ then we denote this part of the spectrum by $\sigma_{r}$.

Theorem 37 Let $\Lambda \in \sigma_{1}, f(0)=0$. Abbreviate $\|T\|_{1}:=\|T\|_{\left[i m \mathcal{Q}, W_{2}^{1}\right]}$ and $\|T\|_{2}:=\|T\|_{\left|\mathrm{m}, L_{2}\right|}$. Then

$$
|\Lambda| \geq\left(L\|T\|_{1}\|T\|_{2}\right)^{-1} .
$$

Proof. The differentiation of the representation formula (19) yields

$$
D u+\frac{1}{\eta} \mathcal{Q} p=\Lambda \mathcal{Q} T f(u)
$$

As in the case of the linear Stokes system we have

$$
\begin{equation*}
\|D u\|_{2}^{2}+\frac{1}{\eta}\|\mathcal{Q} p\|_{2}^{2}=|\Lambda|^{2}\|\mathcal{Q} T f(u)\|_{2}^{2} \tag{21}
\end{equation*}
$$

It can be shown (cf. [12]) that

$$
\sqrt{\frac{\lambda_{1}}{1+\lambda_{1}}}\|u\|_{2,1} \leq|\Lambda|\|T\|_{2}\|f(u)\|_{2} \leq|\Lambda|\|T\|_{2} L
$$

Here, we used that $\|f(u)\|_{2} \leq L$ for $\|u\|_{1,2} \leq 1$ as a consequence of the local Lipschitz condition. Hence,

$$
\|u\|_{2,1} \leq|\lambda|\|T\|_{1}\|T\|_{2} L .
$$

Now the following iteration procedure can be formulated:

$$
\begin{aligned}
& u_{0} \in \stackrel{\circ}{W}_{2}^{1}(G) \cap \text { ker div, } \quad\left\|u_{0}\right\| \leq 1 \\
& u_{n}=\lambda T Q T f\left(u_{n-1}\right)-\frac{1}{\eta} T \mathcal{Q} p_{n} \quad(n=1,2, \ldots)
\end{aligned}
$$

At each step a linear STOKES problem has to be solved. Making some operator estimations, we obtain:

$$
\left\|u_{n}-u_{n-1}\right\|_{2,1} \leq|\Lambda|\|T\|_{1}\|T\|_{2} L\left\|u_{n-1}-u_{n-2}\right\|_{2,1} .
$$

Assuming $|\Lambda|\|T\|_{1}\|T\|_{2} L \leq 1$ we conclude the contractivity of the nonlinear mapping. With the help of (21) it is ensured that $\left\|u_{n}\right\|_{1.2} \leq 1(\forall u \in$ N ). Now, we can apply BANACH's fixed-point theorem and obtain that $\left\{u_{n}, p_{n}\right\}$ converges to the unique solution. Therefore, $\Lambda$ such that

$$
|\Lambda| \leq \frac{1}{L\|T\|_{1} \mid T \|_{2}}
$$

can not belong to $\sigma_{1}$.
What happens when we omit the restrictive condition $f(0)=0$ ?
Remark 38 If $f(0) \neq 0$ then, it can not be ensured that $u_{n} \rightarrow 0$, although the proof of convergence goes through. If we assume that $f: \stackrel{\circ}{W}_{2}^{1} \rightarrow L_{p}(G)$ $\left(\frac{6}{5}<p<\frac{3}{2}\right)$ and

$$
\begin{aligned}
& \|f(u)-f(v)\|_{p} \leq L\|u-v\|_{2,1} \\
& \forall u, v \in W_{2}^{1}(G):\|u\|_{2,1} \leq 1,\|v\|_{2,1} \leq 1
\end{aligned}
$$

We can prove that

$$
\inf _{\Lambda \in \sigma_{1}}|\Lambda| \geq\left[L\|T\|_{1}\|T\|_{\left[L_{p}, L_{2}\right]}\right]^{-1} .
$$

This is a consequence of

$$
\sqrt{\frac{\lambda_{1}}{1+\lambda_{1}}}\|u\|_{2,1} \leq|\lambda|\|T\|_{\left[L_{p}, L_{2}\right]} L
$$

which can be concluded from (21). The assumption $p>\frac{6}{5}$ guarantees that $T: L_{p} \rightarrow L_{2}$ is bounded.
one has the following theorem:
Theorem 39 Let $\Lambda \in \sigma_{1}, f(0)=0$. Then under the assumption

$$
\|f(u)-f(v)\|_{2} \leq L\|u-v\|_{2}, \quad \forall u, v:\|u\|_{2} \leq 1,\|v\|_{2} \leq 1 .
$$

we have that

$$
\inf _{\Lambda \in \sigma_{1}}|\Lambda| \geq \lambda_{1} L^{-1}
$$

Proof. We already know that $\|D u\|_{2} \geq \sqrt{\lambda_{1}}\|u\|_{2} \quad\left(\forall u \in{\stackrel{\circ}{W_{2}}}^{1}\right)$. Therefore, from (21) we get

$$
\lambda_{1}\|u\|_{2}^{2} \leq|\lambda|^{2}\|T\|_{2}^{2}\|f(u)\|_{2}^{2} \leq|\Lambda|^{2}\|T\|_{2}^{2} L^{2}\|u\|_{2}^{2}
$$

Now $\|T\|_{2} \leq 1 / \sqrt{\lambda_{1}}$ makes the proof complete.

Remark 40 The LIPSCHITZ condition in $L_{2}$ is more useful than the condition between $L_{2}$ and $W_{2}^{1}$. But we have to pay for this convenience. Namely, we have another eigenvalue in our final result. Nevertheless, all estimates are constructive, because $\lambda_{1}$ as well as the $T$-norms can be explicitly estimated by properties of $G$.
4.4.1 Lower bound of the spectrum of a non-linear problem in elasticity

Let us consider the following eigenvalue problem (cf. [?][13]):

$$
\begin{aligned}
-\Delta \underline{u}-\frac{m}{m-2} \operatorname{grad} \operatorname{div} \underline{u} & =\Lambda f(\underline{u}) \quad \text { in } G \\
\underline{u} & =0 \text { on } \Gamma
\end{aligned}
$$

The parameter $m$ denotes Poisson's contraction number and $\underline{u}$ is the vector of displacements. As in the previous section we complete the vector function $\underline{u}$ to the quaternionic function $u=u_{0}+\underline{u}$. We obtain the quaternionic formulation

$$
\begin{aligned}
D M D u & =\Lambda f(u) \\
u & =0,
\end{aligned}
$$

where

$$
M u:=\frac{m-2}{2(m-1)} u_{0}+\underline{u} .
$$

An iteration procedure is now defined by

$$
u_{n}=T_{G} \mathcal{Q}_{M} M^{-1} \Lambda T_{G} f\left(u_{n-1}\right.
$$

It is easy to show that $t r_{\Gamma} u_{n}=0!$ ). In [12] it is verified that the sequence $\left(u_{n}\right)$ converges to zero if

$$
\inf _{\Lambda \in \sigma_{1}}|\Lambda| \geq \frac{m-2}{2 L(m-1)\|T\|_{1}\|T\|_{2}}
$$

In that way we get lower bound of $\sigma_{1}$.
Remark 41 Similarly, one get an estimation of the first eigenvalue of the oscillation problem in linear elasticity:

$$
\begin{aligned}
-\Delta \underline{u}-\frac{m}{m-2} \operatorname{grad} \operatorname{div} \underline{u}-\theta^{2} \rho & =\Lambda f(\underline{u}) \quad \text { in } G \\
\underline{u} & =0 \text { on } \Gamma
\end{aligned}
$$

Here $\theta$ denotes the oscillation frequency and $\rho$ the density of the material. The estimation reads :

$$
\Lambda_{1} \geq \frac{m-2}{2(m-1)} \frac{\lambda_{1}}{\sqrt{\lambda_{1}+1}}+\theta^{2} \rho
$$

where $\lambda_{1}$ denotes the first eigenvalue of Dirichlet's problem.

### 4.4.2 Comparison of eigenvalues of different eigenvalue problems

In this section we want to deduce relations between the first eigenvalue of Diri-
chlet' s problem, Stokes' problem and Lamé' s problem. The results were obtained by K. Gürlebeck. Confere our book [12]. Lamè's system

$$
\begin{aligned}
-\Delta \underline{u}-\frac{m}{m-2} \operatorname{grad} \operatorname{div} \underline{u} & =\Lambda \underline{u} \text { in } G, \\
\underline{u} & =0 \text { on } \Gamma,
\end{aligned}
$$

where $\operatorname{Sc} u=0$ for $\Lambda \in \mathbb{R}$. This system has only the trivial solution if

$$
|\Lambda|\left\|T_{G} \mathcal{Q}_{M} M^{-1} T_{G}\right\|_{\left|L_{2}, L_{2}\right|}<1
$$

Hence

$$
\Lambda_{1}(m, G) \geq\left\|T_{G}\right\|_{\left|L_{2}, L_{2}\right|}^{-2} \frac{m-2}{2(m-1)}
$$

Now we can estimate, on the one hand,

On the other hand
$\Lambda_{1}(m, G)=\inf \left\{\frac{\left(\|D u\|_{2}^{2}+\frac{m}{m-2}\|S c D u\|_{2}^{2}\right)}{\|u\|_{2}^{2}}, \quad u \in \stackrel{W}{W}_{2}^{1}\right.$, Sc $\left.u=0, u \neq 0\right\}$.
Therefore, we have monotonicity relatively to $m$ i.e.

$$
\Lambda_{1}(m, G) \leq \Lambda_{1}\left(m^{\prime}, G\right) \text { for } m<m^{\prime} .
$$

Then there exists the limit

$$
\lim _{m \rightarrow 2} \Lambda_{1}=\Lambda . . \leq \Lambda
$$

We are able to formulate the following result:
Theorem 42 (Gürlebeck[10]) Let $\Lambda$. be the first eigenvalue of Stokes equation under Dirichlet's condition and $m>2$. Then we have the inclusion

$$
\left(\frac{\left|B_{1}\right|}{|G|}\right)^{2 / 3} \pi^{2} \leq \lambda_{1}(G)<\Lambda_{1}(m, G) \leq \Lambda \cdot(G)
$$

and $\Lambda_{1}(m, G) \rightarrow \Lambda,(G)$. In other words: The first eigenvalue $\lambda_{1}$ of Dirichlet's problem is smaller than the first eigenvalue $\Lambda_{1}$ of Lamé's problem and this again is smaller then first eigenvalue $\Lambda$. of Stokes' problem.

Remark 43 B. Kawohl obtained in 1987 (cf. [16] the estimation

$$
\left(1+\frac{m}{n(m-2)}\right) \lambda_{1}(G) \geq \Lambda_{1}(m, G) \geq \lambda_{1}(G)
$$

Remark 44 If one has a solution $u$ of Stokes' problem

$$
\begin{aligned}
-\Delta u+\eta^{-1} \operatorname{grad} p & =f \text { in } G, \\
\operatorname{div} u & =0 \text { in } G, \\
u & =0 \text { on } \Gamma,
\end{aligned}
$$

then it is possible to get an upper estimate for $\lambda_{1}$. It holds

$$
0<\lambda_{1}(G) \leq \frac{K^{2}}{1-K^{2}}
$$

where $K:=\left(\sqrt{2}\left\|T_{G} f\right\|_{2}-\eta^{-1}\left\|\mathcal{Q}_{p}\right\|_{2}\right)\|u\|_{2,1}^{-1}$ has to be small enough.

## References

[1] Altman, S.L. (1986) Rotations, quaternions and double groups, Clarendon Press, Oxford.
[2] Arnold W.I. (1973) Ordinary differential equations. MIT Press, Cambridge, Mass.
[3] Bernstein S. (1991) Operator calculus for elliptic boundary value problems in unbounded domains. Zeitschrift $f$. Analysis $u$. Anwend. 10, 4: 447-460.
[4] Bernstein S. (1993) Elliptic boundary value problems in unbounded domains. In: F. Brackx, et al (eds), Clifford algebras and their applications in mathematical physics, Kluwer, Dordrecht : 45-53.
[5] Brackx F. Delanghe R. and Sommen F. (1982) Clifford analysis. Pitman Research Notes in Math., Boston, London, Melbourne.
[6] Cnops J. (1989) Orthogonal functions associated with the Dirac operator. Thesis, State University of Ghent, Academic year 1988-89, 129 pages.
[7] Delanghe R. Sommen F. and Souček V. (1992) Clifford algebra and spinor-valued functions Kluwer, Dordrecht.
[8] Gilbert J.E. and Murray M.A.M. (1990) Clifford algebras and Dirac operators in harmonic analysis. Cambr. Studies in advanced mathematics, Cambridge 26.
[9] Gürlebeck K. (1988) Grundlagen einer diskreten räumlich verallgemeinerten Funktionentheorie und ihrer Anwendungen. Habilitationsschrift, TU Karl-Marx-Stadt.
[10] Gürlebeck K. (1991) Lower and upper bounds for the first eigenvalue of the Lamè system In : Kühnau R. and W. Tutschke (eds.), Boundary value and initial value problems in complex analysis and its applications to differential equations 1, Pitman Research Notes in Mathematics Series 256, 184-192.
[11] Gürlebeck K. and W. Sprössig, (1990) Quaternionic Analysis and Boundary Value Problems, Birkhäuser Verlag, Basel.
[12] Gürlebeck K. and W. Sprössig (1997) Quaternionic Calculus for Physicists and Engineers, Mathematical Methods in Practice: Volume 1, Chichester: John Wiley\& Sons.
[13] Gürlebeck K. and Sprössig W. (1987) A unified approach to estimation of lower bounds for the first eigenvalue of several elliptic boundary value problems. Math. Nachr. 131: 183-199.
[14] Gürlebeck K., Kähler U., Ryan J. Sprössig W. (1997) Clifford analysis over unbounded domains, Advances in Applied Mathematics 19, 216-239.
[15] Gürlebeck K., Sprössig W. and Wimmer U. (1993) Hypercomplex function theory for consideration of non-linear Stokes problems with variable viscosity. Complex Variables, Theory and Appl. 22: 195-202.
[16] Kawohl B. (1987) Estimates for the first eigenvalue of a special elliptic system, Preprint der Universität Heidelberg.
[17] Kravchenko V. and M.V. Shapiro (1996)Integral representations for spatial models of mathematical physics, Pitman Research Notes in Mathematics Series, Addison Wesley Longman, Harlow.
[18] Misner, Ch.W.,Thorne, K.S. and J. A. Wheeler Gravitation, W.H. Freemann and Company, San Francisco.
[19] Porteous I. R. (1995) Clifford Algebras and the Classical Groups, Cambridge Studies in Advanced Mathematics: 50 .
[20] Qian, T.(1997) Generalization of Fueter's result to $\mathbb{R}^{n+1}$. Rend. Mat. Acc. Lincei 8 (1997)9, 111 - 117.
[21] Riesz M. (1958) Clifford numbers and spinors. Lecture series, Institut of Physical Science and Technology, Nr. 38, Maryland.
[22] Sce, M.(1957)Osservazioni sulle serie di potenze nei moduli quadtratici. Atti Acc. Lincei Rend. fis. 23, 220-225.
[23] Sprössig W. and Gürlebeck K. (1996) On the treatment of fluid problems by methods of Clifford analysis. In: Deville M., Gavrilakis S. and Ryhming I.L.(eds), Computation three-dimensional complex flows Notes on Numerical Fluid Mechanics, Vieweg, Braunschweig, Wiesbaden 53: 304-310.

