# Some Rigidity Results Related to Monge-Ampère Functions 

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#### Abstract

The space of Monge-Ampère functions, introduced by J. H. G. Fu, is a space of rather rough functions in which the map $u \mapsto \operatorname{Det} D^{2} u$ is well defined and weakly continuous with respect to a natural notion of weak convergence. We prove a rigidity theorem for Lagrangian integral currents that allows us to extend the original definition of Monge-Ampère functions. We also prove that if a MongeAmpère function $u$ on a bounded set $\Omega \subset \mathbb{R}^{2}$ satisfies the equation Det $D^{2} u=0$ in a particular weak sense, then the graph of $u$ is a developable surface, and moreover $u$ enjoys somewhat better regularity properties than an arbitrary Monge-Ampère function of 2 variables.


## 1 Introduction

The space of Monge-Ampère functions, introduced by J. H. G. Fu [7, 8], is the largest known space of functions $u: \Omega \subset \mathbb{R}^{n} \rightarrow \mathbb{R}$ for which all minors of the Hessian $D^{2} u$, including, in particular, the determinant $\operatorname{det} D^{2} u$, are well defined as signed Radon measures and weakly continuous in a certain natural sense. This makes it an interesting function space from the point of view of analysis and nonlinear potential theory, and also possibly useful for some problems in the calculus of variations and non-smooth geometry.

Technical restrictions in Fu's work forced him to work with Monge-Ampère functions that are locally Lipschitz. The first goal of this paper is to show that basic properties of Monge-Ampère functions, in particular, an underlying theorem about rigidity of Lagrangian intergal currents, which guarantees that the measures associated with det $D^{2} u$ and with other minors of the Hessian are in some sense canonical, remain valid without this local Lipschitz condition. This is carried out in Theorem4.1 and its corollaries, and it allows us to expand the space of Monge-Ampère functions to what we believe is its natural generality and to strengthen Fu's weak continuity results. In [13], examples were constructed showing that in $n$-dimensions there exists a Monge-Ampère function $u$ that is not $C_{\text {loc }}^{0, \gamma}$ for any $\gamma>\frac{2}{n+1}$, so the local Lipschitz assumption that we remove is genuinely restrictive.

The other main result of this paper, Theorem 6.1 establishes a rigidity property of a function $u: \Omega \subset \mathbb{R}^{2} \rightarrow \mathbb{R}$ solving the equation $\operatorname{det} D^{2} u=0$, where $\operatorname{det} D^{2} u$ is now understood in the sense of Monge-Ampère functions. If $u: \Omega \rightarrow \mathbb{R}$ is a sufficiently smooth function and $\operatorname{det} D^{2} u(x)=0$ in $\Omega$, then it is a classical fact that the graph of $u$ is a developable surface, in the sense that for every $x \in \Omega$, either $u$

[^0]is affine in a neighborhood of $x$ or $x$ belongs to a line segment that intersects $\partial \Omega$ at both ends and along which $D u$ is constant. This was proved by Hartman and Nirenberg [11] for $u \in C^{2}$ and by Kirchheim [15] when $u \in W^{2, \infty}$. Pakzad [16] showed that Kirchheim's proof can be extended to $u \in W^{2,2}$ via a lemma that shows that if $u \in W^{2,2}(\Omega)$ solves the equation $\operatorname{det} D^{2} u=0$, then $u$ is $C^{1}$. Pogorelov [17] established a similar developability property for functions that are merely $C^{1}$, without any kind of condition about det $D^{2} u$, but assuming that the image of the gradient map has Lebesgue measure 0 . Here we prove an analogous developability result when $u$ is merely a Monge-Ampère function. In this case $u$ need not be $C^{1}$, in fact the gradient $D u$ is in general merely a function of bounded variation, so that the statement " $D u$ is constant along a line segment" here means that either every point of the segment is a Lebesgue point, or every point is a jump point of $D u$, with the same jump at all points on the segment. The theorem thus implies a modest but optimal regularity property: every point of $\Omega$ is either a Lebesgue point of $D u$ or belongs to the jump set of $D u$.

A Monge-Ampère function $u: \mathbb{R}^{n} \rightarrow \mathbb{R}$ is defined in terms of an $n$-dimensional integral current in $\mathbb{R}^{n} \times \mathbb{R}^{n}$, denoted [ $d u$ ], that can be thought of as a generalized graph of the gradient $D u$. This current is required to be Lagrangian with respect to the canonical symplectic form, see (2.6); roughly speaking, this means that it is weakly curl-free. It should be noted that in the language of Cartesian currents, (see Giaquinta, Modica, Souček [10]) a Monge-Ampère function $u$ is precisely a function whose gradient supports a Lagrangian Cartesian current. Thus, many of our results can be stated as theorems about Cartesian currents. For example, Corollary 4.2 implies that a Lagrangian Cartesian current is uniquely determined by its support function.

The measures associated with minors of the Hessian are defined using this current [du] (see (4.2)); this is motivated by the fact, recalled in (2.4), that if $u$ is smooth, one can recover all minors of the Hessian by integrating suitable $n$-forms over the graph of the gradient.

### 1.1 Some Related Work

Fu established a rigidity result for Legendrian currents, as a corollary of his result about Lagrangian currents [7]. This Legendrian version of the theorem has subsequently been used in a number of applications, including works that develop a theory of curvature measures for a number of classes of rather irregular subsets of Euclidean space, including subanalytic sets [9], Lipschitz manifolds [18], and o-minimal sets [3]. These works rely on the notion of a normal cycle, which is a Legendrian current in $\mathbb{R}^{n} \times S^{n-1}$ that bears roughly the same relation to the graph of the Gauss map as the current $[d u]$ associated with a Monge-Ampère function has to the graph of the gradient. The uniqueness theorem for Legendrian cycles has also been used for further developments of general theory related to Legendrian cycles (see for example [2]) and in problems, arising in computational geometry, relating to estimating curvatures in polygonal approximations of smooth surfaces (see [4]). By contrast, the Lagrangian version of the theorem, and corresponding results about Monge-Ampère functions and weak continuity of minors of the Hessian, have to date received less attention. In
fact, the authors of several recent papers $[6,12,14]$ that discuss the weak continuity of the map $u \mapsto \operatorname{det} D^{2} u$ seem to be unaware of Fu's earlier work: this is certainly the case for the work of N. Jung and the author [14]. These papers interpret det $D^{2} u$ in the sense of distributions, which has the advantage of making it possible to extend the theory to certain functions for which $\operatorname{det} D^{2} u$ is not a measure. In situations where it is natural to require that $\operatorname{det} D^{2} u$ be a measure, however, the geometric measure theory framework of Monge-Ampère functions (equivalently, Lagrangian Cartesian currents) yields sharper results.

We finally mention the note [13], which presents some examples and attempts to give an elementary treatment of some aspects of Monge-Ampère functions.

### 1.2 Organization of This Paper

In Section 2 we recall some background and fix some notation. The definition of Monge-Ampère functions is given in Section 2.5 Section 3 contains some general geometric measure theory results that are used throughout the rest of the paper.

Section 4 contains the proof of our version of Fu's rigidity theorem for Lagrangian currents. This theorem says, heuristically, that a Lagrangian integral current in $\mathbb{R}^{n} \times \mathbb{R}^{n}$ with no boundary and with finite mass in sets of the form $K \times \mathbb{R}^{n}, K$ compact, is uniquely determined by its "most horizontal part." We also deduce some corollaries, including results about the weak continuity of the map $u \mapsto \operatorname{det} D^{2} u$, as well as corresponding results for other minors of the Hessian. In particular, the definition of det $D^{2} u$ for a Monge-Ampère function $u$ is given in (4.2); see also Remark 3, where the relation between our notion of $\operatorname{det} D^{2} u$ and the distributional determinant of the Hessian is discussed.

Section 5 establishes a result that completely characterizes certain 1-dimensional slices of the current $[d u]$ associated with a Monge-Amère function $u$, see Proposition 5.3. This is essentially equivalent to a description of that part of the currrent [ $d u$ ] corresponding to the second derivatives of $u$, i.e., the $1 \times 1$ minors of $D^{2} u$ (see Proposition 5.1). These rather technical results are used in Section6 in the proof of our second main result, Theorem6.1 which shows that if $u: \Omega \subset \mathbb{R}^{2} \rightarrow \mathbb{R}$ is a weak solution (in the sense of Monge-Ampère functions) of the equation $\operatorname{det} D^{2} u=0$, then the graph of $u$ is a developable surface in the sense described above.

## 2 Notation and Background

### 2.1 General Notation

If $\mu$ is a Radon measure, we write $\mu \geq 0$ to mean that $\mu(A) \geq 0$ for all measurable $A$, and we write $\mu_{1} \leq \mu_{2}$ when $\mu_{2}-\mu_{1} \geq 0$.

We write $I(k, n):=\left\{\alpha=\left(\alpha_{1}, \ldots, \alpha_{k}\right) \in \mathbb{Z}^{k}: 1 \leq \alpha_{1}<\cdots<\alpha_{k} \leq n\right\}$. For $\alpha \in I(k, n)$ we write $\bar{\alpha}$ to denote the element of $I(n-k, n)$ with the property that $\left(\alpha_{1}, \ldots, \alpha_{k}, \bar{\alpha}_{1}, \ldots, \bar{\alpha}_{n-k}\right)$ is a permutation of $(1, \ldots, n)$. We write $\operatorname{sgn}(\alpha, \bar{\alpha})$ to denote the sign of the permutation. We write 0 to denote the unique element of $I(0, n)$, and $\overline{0}:=(1, \ldots, n)$. If $\alpha \in I(k, n)$ is a multiindex, then $|\alpha|=k$.

If $A=\left(a_{i j}\right)_{i, j=1}^{n}$ is an $n \times n$ matrix and $\alpha, \beta \in I(k, n)$, then $M^{\alpha \beta}(A)=\operatorname{det}\left[A^{\alpha \beta}\right]$,
where $A^{\alpha \beta}$ is the $k \times k$ matrix whose $i, j$ entry is $a_{\alpha_{i} \beta_{j}}$. We use the convention $M^{00}(A)=1$.

We write $B_{r}(a)$ to denote the open ball $\{x:|x-a|<r\}$; the ambient space is normally clear from the context.

We often implicitly sum over repeated indices. However, when we sum over multiindices, we normally indicate the sum explicitly.

### 2.2 Geometric Measure Theory Notation

We assume some familiarity with basic definitions and facts of geometric measure theory, such as currents, rectifiability, the coarea formula, and properties of functions of bounded variation. Here we recall some notions that will be used often, and we point out some ways in which our notation differs from that found in standard references such as Federer [5].

If $v: \Omega \rightarrow \mathbb{R}^{\ell}$ is a BV function, then we write $\mathcal{J}_{v}$ to denote the jump set of $v$, and for $x \in \mathcal{J}_{v}$, we write $v^{+}(x), v^{-}(x)$ to denote the approximate limits of $v$ on the two sides of $\mathcal{J}_{v}$ ( see [1, Proposition 3.69].)

We say " $j$-rectifiable", or if no confusion can result, simply "rectifiable" to mean what Federer (see $[5,3.2 .14])$ calls "countably $\left(\mathcal{H}^{j}, j\right)$ rectifiable."

We follow convention and write $\|T\|$ to denote the total variation measure associated with a current $T$ of locally finite mass.

If $\Gamma$ is an $\ell$-rectifiable subset of some Euclidean space $\mathbb{R}^{M}$, and $m: \Gamma \rightarrow(0, \infty)$ and $\tau: \Gamma \rightarrow \wedge_{\ell} \mathbb{R}^{M}$ are $\mathcal{H}^{\ell}$ measurable, locally integrable functions such that $\tau(x)$ is a simple $\ell$-vector associated with the approximate tangent space $T_{x} \Gamma$ at $\mathcal{H}^{\ell}$ a.e. $x \in \Gamma$, then we write $\mathrm{T}(\Gamma, m, \tau)$ to denote the current defined by

$$
\begin{equation*}
\underline{\mathrm{T}}(\Gamma, m, \tau)(\phi)=\int_{\Gamma}\langle\phi(x), \tau(x)\rangle m(x) d \mathcal{H}^{\ell}(x) \tag{2.1}
\end{equation*}
$$

A current of the form (2.1) is said to be rectifiable. If $|\tau|=1$ and $m(x) \in \mathbb{Z}$ for $\mathcal{H}^{\ell}$ a.e. $x \in \Gamma$, then we say that $\underline{T}(\Gamma, m, \tau)$ is integer multiplicity rectifiable, abbreviated as i.m. rectifiable.

If $W$ is a $j$-rectifiable subset of some $\mathbb{R}^{\ell}$, and $Z$ is a $k$-rectifiable subset of $\mathbb{R}^{m}$, and $F: \mathbb{R}^{\ell} \rightarrow \mathbb{R}^{m}$ is a Lipschitz map such that $F(x) \in Z$ for $\mathcal{H}^{j}$ almost every $x \in W$, then we will sometimes write $J_{W \rightarrow Z} F$ to denote the Jacobian as appearing in the coarea formula; see [5, 3.2.22]. This might be written by Federer as " $J_{k} f$, where $f$ is the restriction of $F$ to $W$." We will normally omit the subscripts when no confusion can result.

### 2.3 Notation Related to Product Space Structure

The setting for most of our results is a product space, which we will often write as $\Omega_{h} \times \Omega_{v}$, with $\Omega_{h}$ an open subset of $\mathbb{R}^{n}$ and $\Omega_{v}$ an open subset of $\mathbb{R}^{m}$. We will refer to $\Omega_{h}$ and $\Omega_{v}$ as horizontal and vertical, respectively. Except in Section 3, we require that $m=n$. In cases when $\Omega_{v}$ is a Euclidean space, we will often drop the subscripts and simply write $\Omega \subset \mathbb{R}^{n}$ for the horizontal space, and $\Omega \times \mathbb{R}^{m}$ for the product space.

We write $p_{h}: \Omega_{h} \times \Omega_{v} \rightarrow \Omega_{h}$ and $p_{v}: \Omega_{h} \times \Omega_{v} \rightarrow \Omega_{v}$ to denote the natural projections.

We always write $\left(x_{1}, \ldots, x_{n}\right)$ and $\left(\xi_{1} \ldots, \xi_{m}\right)$ for coordinates on the horizontal space $\Omega_{h}$ and the vertical space $\Omega_{v}$, respectively. Thus $d x_{1}, \ldots, d x_{n}$ will denote horizonal covectors, and $d \xi_{1}, \ldots, d \xi_{m}$ vertical covectors. We also write $\left\{e_{1}, \ldots, e_{n}\right\}$ to denote the standard basis for the tangent space to $\Omega_{h}$, and $\left\{\varepsilon_{1}, \ldots, \varepsilon_{m}\right\}$ to denote the basis for the tangent space to $\Omega_{v}$. The bases for the spaces of vectors and covectors are assumed to be dual in the sense that $\left\langle d x_{\alpha} \wedge d \xi_{\beta}, e_{\gamma} \wedge \varepsilon_{\delta}\right\rangle=1$ if $\alpha=\gamma$ and $\delta=\beta$ and 0 if not. Here, for example, $d x_{\alpha}:=d x_{\alpha_{1}} \wedge \cdots \wedge d x_{\alpha_{j}}$ if $\alpha=\left(\alpha_{1}, \ldots, \alpha_{j}\right) \in I(j, n)$.

We write $\mathcal{D}^{\ell}\left(\Omega_{h} \times \Omega_{v}\right)$ to denote the space of all $C^{\infty} \ell$-forms with compact support in $\Omega_{h} \times \Omega_{v}$. For a covector $\phi=\sum_{|\alpha|+|\beta|=j+k} \phi^{\alpha \beta} d x_{\alpha} \wedge d \xi_{\beta}$, we will write

$$
\begin{equation*}
P^{j, k} \phi=\sum_{\substack{|\alpha|=j \\|\beta|=k}} \phi^{\alpha \beta} d x_{\alpha} \wedge d \xi_{\beta} . \tag{2.2}
\end{equation*}
$$

For a differential form $\phi \in \mathcal{D}^{j+k}\left(\Omega_{h} \times \Omega_{v}\right)$, we define $\left(P^{j, k} \phi\right)(x, \xi)=P^{j, k} \phi(x, \xi)$. For a current $T \in \mathcal{D}_{j+k}\left(\Omega_{h} \times \Omega_{v}\right)$, we define $P_{j, k} T(\phi)=T\left(P^{j, k} \phi\right)$. Note that $P_{j, k} \underline{\mathrm{~T}}(\Gamma, m, \tau)=\underline{\mathrm{T}}\left(\Gamma, m, P_{j, k} \tau\right)$, where $P_{j, k} \tau$ is defined as in (2.2). Given a vector $w \in T_{(x, \xi)}\left(\Omega_{h} \times \Omega_{v}\right)$, we will often use the notation

$$
\begin{equation*}
w_{h}=\sum_{i=1}^{n}\left(w \cdot e_{i}\right) e_{i}=P_{1,0} w, \quad w_{v}=\sum_{i=1}^{m}\left(w \cdot \varepsilon_{i}\right) \varepsilon_{i}=P_{0,1} w . \tag{2.3}
\end{equation*}
$$

For $v \in C^{1}\left(\Omega ; \mathbb{R}^{m}\right)$, following Giaquinta, Modica, and Souček [10] we write $G_{v}$ to denote the current associated with integration over the graph of $v$.

$$
\begin{equation*}
G_{v}(\phi)=\int_{\Omega} \sum_{|\alpha|+|\beta|=n} \phi^{\alpha \beta}(x, v(x)) \operatorname{sgn}(\alpha, \bar{\alpha}) M^{\bar{\alpha} \beta}(D v(x)) d x \tag{2.4}
\end{equation*}
$$

Thus $P_{n-k, k} G_{v}$ encodes the $k \times k$ minors of $D v$. If $v \in W^{1, p}$ for $p \geq \min \{m, n\}$, then the above expression still makes sense and $\partial G_{v}=0$ in $\Omega \times \mathbb{R}^{m}$ (see [10]).

We remark that if $T$ is an $\ell$-current of locally finite mass in $\Omega \times \mathbb{R}^{m}$ such that $\|T\|\left(K \times \mathbb{R}^{m}\right)<\infty$ for every compact $K \subset \Omega$, then $T(\phi)$ is well defined whenever $\phi$ is a smooth $\ell$-form with support in $K \times \mathbb{R}^{m}$, the point being that compact support is not required in the vertical directions. Indeed, let $\chi_{R}$ be a family of functions such that $\chi_{R}(x, \xi)=1$ if $|\xi| \leq R, \chi_{R}(x, \xi)=0$ if $|\xi| \geq 2 R$, and $\left|\nabla_{\xi} \chi_{R}\right| \leq C / R$ and $\nabla_{x} \chi_{R} \equiv 0$. Then it is easy to check that $\lim _{R \rightarrow \infty} T\left(\chi_{R} \phi\right)$ exists and is independent of the specific choice of $\left\{\chi_{R}\right\}$; this is how we define $T(\phi)$.

For such $T$ it follows that $p_{h \#} T$ is well defined, where $p_{h \#} T(\phi)=T\left(p_{h}^{\#} \phi\right)$ for $\phi \in \mathcal{D}^{\ell}(\Omega)$. One can also check that the standard identity $\partial p_{h \#} T=p_{h \#} \partial T$ still holds, as long as $(\|T\|+\|\partial T\|)\left(K \times \mathbb{R}^{m}\right)<\infty$ for $K$ compact. Indeed, if we let $\chi_{R}$ be as above, then

$$
\begin{aligned}
\partial p_{h \#} T(\phi) & =p_{h \#} T(d \phi)=\lim _{R \rightarrow \infty} T\left(\chi_{R} p_{h}^{\#} d \phi\right)=\lim _{R \rightarrow \infty} T\left(\chi_{R} d p_{h}^{\#} \phi\right) \\
& =\lim _{R \rightarrow \infty}\left[T\left(d\left(\chi_{R} p_{h}^{\#} \phi\right)\right)-T\left(\left(d \chi_{R}\right) p_{h}^{\#} \phi\right)\right]=p_{h \#} \partial T .
\end{aligned}
$$

### 2.4 Lagrangian Currents

We write $\omega$ to designate the standard symplectic form $\omega=\sum_{i=1}^{n} d x_{i} \wedge d \xi_{i}$ on $\mathbb{R}^{n} \times \mathbb{R}^{n}$ or any open subset thereof, so that if $v, w$ are vectors, then

$$
\omega(v \wedge w)=\sum_{i=1}^{n}\left[\left(v \cdot e_{i}\right)\left(w \cdot \varepsilon_{i}\right)-\left(v \cdot \varepsilon_{i}\right)\left(w \cdot e_{i}\right)\right]=v \cdot \mathrm{~J} w,
$$

where $J \varepsilon_{i}=e_{i}, J e_{i}=-\varepsilon_{i}$. For $n \geq 2$, an $n$-plane $P$ in $\mathbb{R}^{n} \times \mathbb{R}^{n}$ is said to be Lagrangian if $\left\langle\omega, \tau \wedge \tau^{\prime}\right\rangle=0$ for any two vectors $\tau, \tau^{\prime}$ tangent to $P$. (When $n=1$ we consider every 1-plane in $\mathbb{R} \times \mathbb{R}$ to be Lagrangian.) A rectifiable $n$-current in $\mathbb{R}^{n} \times \mathbb{R}^{n}$ is Lagrangian if $\mathcal{H}^{n}$ a.e. approximate tangent plane is Lagrangian.

Note that the current $G_{v}$ associated as in (2.4) with a smooth map $v: \Omega_{h} \rightarrow \mathbb{R}^{n}$ is Lagrangian if and only if $v_{x_{j}}^{i}=v_{x_{i}}^{j}$ for all $i, j$. This is not hard to check. In particular, if $u: \Omega_{h} \rightarrow \mathbb{R}$ is a smooth function, then $G_{D u}$ is always Lagrangian.

An alternate definition is sometimes given, whereby if $U$ is an open subset of $\mathbb{R}^{n} \times \mathbb{R}^{n}, n \geq 2$, then a current $T \in \mathcal{D}_{n}(U)$ is said to be Lagrangian if $T(\omega \wedge \eta)=0$ for every $\eta \in \mathcal{D}^{n-2}(U)$. It is clear that this condition is preserved under weak convergence.

The alternate definition makes sense for currents that are not necessarily rectifiable, and it agrees with the one we have given for rectifiable currents. We sketch the well-known argument: recall that

$$
\left\langle\omega \wedge \eta, \tau_{1} \wedge \cdots \wedge \tau_{n}\right\rangle=\sum_{\alpha \in I(2, n)} \operatorname{sgn}(\alpha, \bar{\alpha})\left\langle\omega, \tau_{\alpha_{1}} \wedge \tau_{\alpha_{2}}\right\rangle\left\langle\eta, \tau_{\bar{\alpha}_{1}} \wedge \cdots \wedge \tau_{\bar{\alpha}_{n-2}}\right\rangle=0
$$

for every $n-2$-covector $\eta$. From this it is not hard to see that an $n$-plane is Lagrangian if and only if $\langle\omega \wedge \eta, \tau\rangle=0$ for every $n-2$-covector $\eta$ and every orienting $n$-vector $\tau$. Then the equivalence of the two definitions (whenever both make sense) can be verified by rather standard measure theoretic arguments.

### 2.5 Definition of Monge-Ampère Functions

If $\Omega$ is an open subset of $\mathbb{R}^{n}$, then $u \in W_{\text {loc }}^{1,1}(\Omega)$ is said to be a Monge-Ampère function if there exists an $n$-dimensional i.m. rectifiable current $[d u]$ in $\Omega \times \mathbb{R}^{n}$ such that

$$
\begin{gather*}
\partial[d u]=0  \tag{2.5}\\
{[d u] \text { is Lagrangian },}  \tag{2.6}\\
\|[d u]\|\left(K \times \mathbb{R}^{n}\right)<\infty \text { whenever } K \subset \Omega \text { is compact }  \tag{2.7}\\
{[d u]\left(\phi d x_{1} \wedge \cdots \wedge d x_{n}\right)=\int \phi(x, D u(x)) d x} \tag{2.8}
\end{gather*}
$$

for every $\phi \in C_{c}^{\infty}\left(\Omega \times \mathbb{R}^{n}\right)$. If $u$ is, for example, $C^{2}$, then the current $G_{D u}$ associated with integration over the graph of $D u$ satisfies these conditions. For example, in this
case (2.8) is an immediate consequence of (2.4). Property (2.5) and rectifiability are well known (see for example [7,10].

The earlier work of $\mathrm{Fu}[7,8]$ gave a different definition of Monge-Ampère functions in which (2.7) was replaced by the stronger condition that [du] be "locally vertically bounded", which can only hold if $u$ is locally Lipschitz. The terminology here also differs slightly from that used in [13], where we reserved the term Monge-Ampère for functions $u$ such that $\mathbf{M}([d u])<\infty$; functions satisfying (2.5)-(2.8) were called locally Monge-Ampère functions.

If $u$ is a Monge-Ampère function, then $D u$ has locally bounded variation (see [7] for the proof). Examples of Monge-Ampère functions include convex functions, or more generally functions of the form $\min \left\{u_{1}, \ldots, u_{k}\right\}$ where $u_{1}, \ldots, u_{k}$ are semiconvex (see again [7]). From these examples it follows that Monge-Ampère functions need not belong to $W^{2, p}$ for any $p \geq 1$.

## 3 Decomposition of a Stratum of $T$

In this section we assume $\Omega_{h}$ is an open subset of $\mathbb{R}^{n}, \Omega_{v}$ is an open subset of $\mathbb{R}^{m}$, and $j, k$ are nonnegative integers with $j \leq n$ and $k \leq m$. We first state a lemma that assembles some results from [5] and fixes some notation that we will use in this section.

Lemma 3.1 Assume that $T=\underline{T}(\Gamma, m, \tau)$ is a rectifiable $j+k$-current in $\Omega_{h} \times \Omega_{v} \subset$ $\mathbb{R}^{n} \times \mathbb{R}^{m}$ with $\mathbf{M}(T)<\infty$, and define

$$
\begin{equation*}
\Gamma_{h}:=\left\{x \in \Omega_{h}: \mathcal{H}^{k}\left(p_{h}^{-1}(x) \cap \Gamma\right)>0\right\} . \tag{3.1}
\end{equation*}
$$

Then $\Gamma_{h}$ is $j$-rectifiable, and the following hold for $\mathcal{H}^{j}$ a.e. $x \in \Gamma_{h}$. First, we can write $\tau(x, \xi)$ in the form

$$
\begin{equation*}
\tau(x, \xi)=\left(\tau_{v}^{1}+\tau_{h}^{1}\right) \wedge \cdots \wedge\left(\tau_{v}^{j}+\tau_{h}^{j}\right) \wedge \tau_{v}^{j+1} \wedge \cdots \wedge \tau_{v}^{j+k} \tag{3.2}
\end{equation*}
$$

for $\mathcal{H}^{k}$ a.e. $\xi \in p_{h}^{-1}(x) \cap \Gamma$, using the notation (2.3), where $\left\{\tau_{h}^{i}\right\}_{h=1}^{j}$ are orthogonal and $\left\{\tau_{v}^{i}\right\}_{i=j+1}^{n}$ are orthonormal, and $\tau^{i} \perp \tau_{v}^{i^{\prime}}$ for all $i \leq j$ and $i^{\prime}>j$. Next, if we write

$$
\begin{equation*}
\Gamma_{v}(x):=\left\{(x, \xi) \in p_{h}^{-1}(x) \cap \Gamma: J_{\Gamma \rightarrow \Gamma_{h}} p_{h}(x, \xi)>0\right\} \tag{3.3}
\end{equation*}
$$

then $\Gamma_{v}(x)$ is $k$-rectifiable, and at $\mathcal{H}^{k}$ a.e. point in $\Gamma_{v}(x)$,

$$
\begin{equation*}
\operatorname{span}\left\{\tau_{h}^{1}, \ldots, \tau_{h}^{j}\right\}=T_{x} \Gamma_{h}, \quad \operatorname{span}\left\{\tau_{v}^{j+1}, \ldots, \tau_{v}^{j+k}\right\}=T_{(x, \xi)} \Gamma_{v}(x) \tag{3.4}
\end{equation*}
$$

Finally,

$$
\begin{equation*}
J_{\Gamma \rightarrow \Gamma_{h}} p_{h}(x, \xi)=\left|\tau_{h}^{1} \wedge \cdots \wedge \tau_{h}^{j}\right| \tag{3.5}
\end{equation*}
$$

The proof, consisting mostly of a string of references to Federer [5], is given at the end of this section.

The main result of this section is the following.

Proposition 3.2 Assume that $T=\underline{T}(\Gamma, m, \tau)$ is an i.m. rectifiable $j+k$-current in $\Omega_{h} \times \Omega_{v}$, and that $\mathbf{M}(T)+\mathbf{M}(\partial T)<\infty$. Further assume that $P_{j+1, k-1} T=0$.

Define $\Gamma_{h}$ and $\Gamma_{v}(x)$ as in Lemma 3.1 and let $\tau_{h}: \Gamma_{h} \rightarrow \wedge_{j} \mathbb{R}^{n}$ be a measurable map such that $\tau_{h}(x)$ is for $\mathcal{H}^{j}$ a.e. $x$ a unit simple $j$-vector associated with $T_{x} \Gamma_{h}$. Then for $\mathcal{H}^{j}$ a.e. $x \in \Gamma_{h}$, there exists an i.m. rectifiable $k$-current $V_{x}$ in $\Omega_{h} \times \Omega_{v}$ carried by $\Gamma_{v}(x)$, such that if $\alpha \in I(j, n), \beta \in I(k, m)$, and $\phi \in C_{c}^{\infty}\left(\Omega_{h} \times \Omega_{v}\right)$, then the map $x \mapsto V_{x}\left(\phi d \xi_{\beta}\right)$ is $\mathcal{H}^{j}\left\llcorner\Gamma_{h}\right.$ integrable, and

$$
\begin{equation*}
T\left(\phi d x_{\alpha} \wedge d \xi_{\beta}\right)=\int_{\Gamma_{h}} V_{x}\left(\phi d \xi_{\beta}\right)\left\langle d x_{\alpha}, \tau_{h}(x)\right\rangle d \mathcal{H}^{j}(x) \tag{3.6}
\end{equation*}
$$

If in addition $\partial T=0$ in $\Omega_{h} \times \Omega_{v}$, then $\partial V_{x}=0$ in $\Omega_{h} \times \Omega_{v}$ for $\mathcal{H}^{j}$ a.e. $x \in \Gamma_{h}$. Finally, if $A$ is any Borel subset of $\Omega_{h} \times \Omega_{v}$, then

$$
\begin{equation*}
\int_{\Gamma_{h}}\left\|V_{x}\right\|(A) \mathcal{H}^{j}(d x) \leq\|T\|(A) \tag{3.7}
\end{equation*}
$$

If $j=n$, then the condition $P_{j+1, k-1} T=0$ is automatically satisfied, and for $\mathcal{L}^{n}$ a.e. $x \in \Omega_{h}$, the current $V_{x}$ is just a slice $\left\langle T, p_{h}, x\right\rangle$ of $T$ by a level set of the projection $p_{h}: \Omega_{h} \times \Omega_{v} \rightarrow \Omega_{h}$.

Remark 1 One can use the rectifiable slices theorem of B. White [20] to show that $P_{j, k} T$ can be identified with a rectifiable flat $j$-chain in $\Omega_{h}$ with coefficients in the (normed abelian) group of flat $k$-chains in $\Omega_{v}$. In fact this argument can be used to prove Proposition 3.2. although that proof would be more difficult than the one we give here. Note that on a purely formal level, the expression on the right-hand side of (3.6) looks like a rectifiable $j$-current in $\Omega_{h}$, carried by the set $\Gamma_{h}$, oriented by $\tau_{h}(\xi)$, and with "multiplicity" at $x \in \Gamma_{h}$ given by $V_{x}$.

For the proof of the proposition we will need the following.
Lemma 3.3 Assume that $T=\underline{T}(\Gamma, m, \tau)$ is an i.m. rectifiable $j+k$-current in $\Omega_{h} \times \Omega_{v}$, and that $\mathbf{N}(T):=\mathbf{M}(T)+\mathbf{M}(\partial T)<\infty$. Further assume that $P_{j+\ell, k-\ell} T=0$ for all $\ell \geq 1$. Then $\|T\|=\|T\|\left\llcorner\left(\Gamma_{h} \times \Omega_{v}\right)\right.$, for $\Gamma_{h}$ as defined in (3.1) above.

The conclusion of the lemma can fail if we do not assume $\mathbf{M}(\partial T)<\infty$ or if for example $P_{j+\ell, k-\ell} T \neq 0$ for some $\ell \geq 1$.

Proof First suppose that $j=0$, in which case the hypotheses imply that $T=P_{0, k} T$. We may suppose that $k \geq 1$, since the conclusion is clear if $j=k=0$. It suffices to show that there exists a finite or countable set $\left\{x^{i}\right\} \subset \Omega_{h}$ such that

$$
\begin{equation*}
T=T\left\llcorner\left(\cup\left\{x^{i}\right\} \times \Omega_{v}\right)\right. \tag{3.8}
\end{equation*}
$$

since then the definition (3.1) implies that $\Gamma_{h} \subset\left\{x_{i}\right\}$. As remarked by Fu [7], (3.8) follows from Lemma 3.3 in Solomon [19]. We recall the argument for the reader's convenience, and because we will need it later: first, using [5, 4.2.25] and the assumption that $T$ is integer multiplicity, we can write $T$ as a countable sum $T=\sum T_{i}$ of
indecomposable components, with $\mathbf{N}(T)=\sum \mathbf{N}\left(T_{i}\right)$. Solomon's lemma asserts that if $f$ is any Lipschitz real-valued function such that $\langle T, f, r\rangle=0$ for a.e. $r \in \mathbb{R}$, then each $T_{i}$ is supported in a level set of $f$. Let $f_{q}(x, \xi)=e_{q} \cdot x=x_{q}$ for $q \in\{1, \ldots, n\}$. Then for every $\phi \in \mathcal{D}^{k-1}\left(\Omega_{h} \times \Omega_{v}\right)$

$$
\int\left\langle T, f_{q}, r\right\rangle(\phi) d r=T\left(d x_{q} \wedge \phi\right)=0
$$

since $P_{0, k} T=T$. It follows that $\left\langle T, f_{q}, r\right\rangle=0$ for a.e. $r$, and hence that for each $T_{i}$, there exists some $r_{1}^{i}, \ldots, r_{n}^{i}$ such that $T_{i}$ is supported in $\bigcap_{q=1}^{n} f_{q}^{-1}\left(r_{q}^{i}\right)$. In other words, if we write $x^{i}=\left(r_{1}^{i}, \ldots, r_{n}^{i}\right)$, then $T_{i}=T_{i}\left\llcorner\left(\left\{x^{i}\right\} \times \Omega_{v}\right)\right.$, establishing the lemma when $j=0$.

Next we prove the lemma for arbitrary positive $j \leq n$. For $\alpha \in I(j, n)$, let $p_{\alpha}(x, \xi)=\left(x_{\alpha_{1}}, \ldots, x_{\alpha_{j}}\right)$. For every $\alpha$ and a.e. $y \in \mathbb{R}^{j}$, the slice $\left\langle T, p_{\alpha}, y\right\rangle$ is a $k$-dimensional rectifiable current with finite mass and finite boundary mass, and it is easy to check that $P_{\ell, k-\ell}\left\langle T, p_{\alpha}, y\right\rangle=0$ for all $\ell \geq 1$. Thus the $j=0$ case implies that $\left\langle T, p_{\alpha}, y\right\rangle$ has the form (3.8). In particular the definition (3.1) of $\Gamma_{h}$ then implies that $\left\langle T, p_{\alpha}, y\right\rangle$ is supported in $\Gamma_{h} \times \Omega_{v}$. Then for $B=\left(\Omega_{h} \backslash \Gamma_{h}\right) \times \Omega_{v}$ we have

$$
\mathbf{M}\left(T\llcorner B) \leq \sum_{\alpha \in I(j, n)} \int_{\mathbb{R}^{j}} \mathbf{M}\left(\left\langle T, p_{\alpha}, y\right\rangle\llcorner B) d \mathcal{L}^{j}(y)=0\right.\right.
$$

This completes the proof of the lemma.
Proof of Proposition 3.2 In this proof we write $J p_{h}$ instead of $J_{\Gamma \rightarrow \Gamma_{h}} p_{h}$. We use notation from Lemma 3.1throughout.

It is convenient initially to assume that $T$ satisfies the hypotheses of Lemma 3.3, which are stronger than those of the theorem in that we require $P_{j+\ell, k-\ell} T=0$ for all $\ell \geq 1$, rather than only for $\ell=1$. This assumption will be relaxed in Step 3.
Step 1. Let us write $\Gamma^{*}:=\left\{(x, \xi) \in \Gamma: J p_{h}(x, \xi)>0\right\}$, so that $\Gamma_{v}(x)=$ $\Gamma^{*} \cap p_{h}^{-1}(x)$. Lemma3.1 implies that $\mathcal{H}^{j+k}$ a.e. in $\Gamma^{*}$,

$$
\frac{\tau_{h}^{1} \wedge \cdots \wedge \tau_{h}^{j}}{\left|\tau_{h}^{1} \wedge \cdots \wedge \tau_{h}^{j}\right|}=\frac{\tau_{h}^{1} \wedge \cdots \wedge \tau_{h}^{j}}{J p_{h}}
$$

is a unit simple vector orienting $T_{x} \Gamma_{h}$, so there exists some function $\sigma(x, \xi): \Gamma^{*} \rightarrow$ $\pm 1$ such that

$$
\frac{\tau_{h}^{1} \wedge \cdots \wedge \tau_{h}^{j}(x, \xi)}{J p_{h}(x, \xi)}=\sigma(x, \xi) \tau_{h}(x)
$$

a.e. in $\Gamma^{*}$. Next, let us write $\tau_{v}(x, \xi):=\sigma(x, \xi) \tau_{v}^{j+1} \wedge \cdots \wedge \tau_{v}^{n}$. Then if $|\alpha|=j,|\beta|=k$,

$$
\begin{align*}
\left\langle d x_{\alpha} \wedge d \xi_{\beta}, \tau\right\rangle & =\left\langle d x_{\alpha} \wedge d \xi_{\beta}, P_{j, k} \tau\right\rangle  \tag{3.9}\\
& =\left\langle d x_{\alpha} \wedge d \xi_{\beta}, \tau_{h}^{1} \wedge \cdots \wedge \tau_{h}^{j} \wedge \tau_{v}^{j+1} \wedge \cdots \wedge \tau_{v}^{n}\right\rangle \\
& =J p_{h}(x, \xi)\left\langle d x_{\alpha}, \tau_{h}\right\rangle\left\langle d \xi_{\beta}, \tau_{v}\right\rangle
\end{align*}
$$

Lemma 3.3 implies that $\mathcal{H}^{j+k}$ almost all of $\Gamma$ is contained in $\Gamma_{h} \times \Omega_{v}$, so that $p_{h}$ maps almost all of $\Gamma$ into $\Gamma_{h}$. Using the coarea formula [5,3.2.22] and (3.9) we can write, still for $|\alpha|=j,|\beta|=k$,

$$
\begin{align*}
T(\phi & \left.d x_{\alpha} \wedge d \xi_{\beta}\right)  \tag{3.10}\\
& =\int_{\Gamma} \phi(x, \xi)\left\langle d x_{\alpha} \wedge d \xi_{\beta}, \tau(x, \xi)\right\rangle m(x, \xi) d \mathcal{H}^{j+k}(x, \xi) \\
& =\int_{\Gamma_{h}} \int_{\Gamma^{*} \cap p_{h}^{-1}(x)} \phi \frac{\left\langle d x_{\alpha} \wedge d \xi_{\beta}, \tau(x, \xi)\right\rangle}{J p_{h}(x, \xi)} m d \mathcal{H}^{k}(\xi) d \mathcal{H}^{j}(x) \\
& =\int_{\Gamma_{h}}\left(\int_{\Gamma_{v}(x)} \phi(x, \xi)\left\langle d \xi_{\beta}, \tau_{v}\right\rangle m(x, \xi) d \mathcal{H}^{k}(\xi)\right)\left\langle d x_{\alpha}, \tau_{h}\right\rangle d \mathcal{H}^{j}(x)
\end{align*}
$$

Recall from Lemma3.1 that for $\mathcal{H}^{j}$ a.e. $x \in \Gamma_{h}, \Gamma_{v}(x)$ is $\mathcal{H}^{k}$-rectifiable and $\tau_{v}(x, \xi)$ is a unit simple tangent vector orienting $T_{(x, \xi)} \Gamma_{v}(x)$ for $\mathcal{H}^{k}$ a.e. $(x, \xi) \in \Gamma_{v}(x)$. So the current $V_{x}$ defined by

$$
\begin{equation*}
V_{x}\left(\psi d \xi_{\beta}\right):=\int_{\Gamma_{v}(x)} \psi(x, \xi)\left\langle d \xi_{\beta}, \tau_{v}\right\rangle m(x, \xi) d \mathcal{H}^{k}(\xi) \tag{3.11}
\end{equation*}
$$

is rectifiable for $\mathcal{H}^{j}$ a.e. $x$. Combining (3.10) and (3.11), we have proved (3.6). Note that $x \mapsto V_{x}\left(\phi d \xi_{\beta}\right)$ is $\mathcal{H}^{j}\left\llcorner\Gamma_{h}\right.$-integrable; this too is a consequence of the coarea formula [5, 3.2.22].

Step 2.
Note that $T$ is i.m. rectifiable if and only if $m(x, \xi)$ is an integer $\mathcal{F}^{j+k}$ almost everywhere, and similarly $V_{x}$ is i.m. rectifiable for $\mathcal{H}^{j}$ a.e. $x$ if and only if for $\mathcal{H}^{j}$ a.e. $x$, $m(x, \xi)$ is an integer for $\mathcal{H}^{k}$ a.e. $\xi$. By the coarea formula again, the former condition implies the latter.
Step 3. We now prove that the same conclusions remain valid if we merely assume that $P_{j+1, k-1} T=0$. If this holds, define $T^{\prime}=\sum_{\ell \leq 0} P_{j+\ell, k-\ell} T$. We claim $T^{\prime}$ verifies all the hypotheses of Steps 1 and 2 above. It is clear that $P_{j+\ell, k-\ell} T^{\prime}=0$ for every $\ell \geq 1$ and that $\mathbf{M}\left(T^{\prime}\right) \leq \mathbf{M}(T)<\infty$. To see that $\mathbf{M}\left(\partial T^{\prime}\right)<\infty$, one checks from the definitions (and using the hypothesis $P_{j+1, k-1} T=0$ for the third equality below) that if we write $P^{\prime} \phi=\sum_{\ell \leq 0} P^{j+\ell, k-\ell-1} \phi$, then

$$
\partial T^{\prime}(\phi)=T^{\prime}(d \phi)=T^{\prime}\left(d P^{\prime} \phi\right)=T\left(d P^{\prime} \phi\right)=\partial T\left(P^{\prime} \phi\right)
$$

Since $\left|P^{\prime} \phi(x)\right| \leq|\phi(x)|$ at every $x$, this implies that $\mathbf{M}\left(\partial T^{\prime}\right) \leq \mathbf{M}(\partial T)$. In addition, we prove in Lemma 3.4 below that $T^{\prime}$ is i.m. rectifiable, so we have checked all the relevant hypotheses.

Let us write $\Gamma^{\prime}$ to denote the set that carries $T^{\prime}$. Clearly $\Gamma^{\prime} \subset \Gamma$ up to a set of $\mathcal{H}^{j+k}$ measure zero. Applying Steps 1 and 2 to $T^{\prime}$, we find that for $\mathcal{H}^{j}$ a.e. $x \in \Gamma_{h}^{\prime}:=$ $\left\{x \in \Omega_{h}: \mathcal{H}^{k}\left(p_{h}^{-1}(x) \cap \Gamma^{\prime}\right)>0\right\}$ there exists a current $V_{x}^{\prime}$ such that (3.6) holds with $T, \Gamma_{h}, V_{x}$ replaced by $T^{\prime}, \Gamma_{h}^{\prime}, V_{x}^{\prime}$.

Since $\Gamma^{\prime} \subset \Gamma$, it is clear that $\Gamma_{h}^{\prime} \subset \Gamma_{h}$. It is clear from the definition of $T^{\prime}$ that $P_{j, k} T=P_{j, k} T^{\prime}$, so if we define

$$
V_{x}= \begin{cases}V_{x}^{\prime} & \text { if } x \in \Gamma_{h}^{\prime} \\ 0 & \text { if not }\end{cases}
$$

then (3.6) is satisfied by $T$.
Step 4. We now prove that

$$
\begin{equation*}
\text { if } \partial T=0 \text { in } \Omega_{h} \times \Omega_{v} \text {, then } \partial V_{x}=0 \text { in } \Omega_{h} \times \Omega_{v} \text { for } \mathcal{H}^{j} \text { a.e. } x \text {. } \tag{3.12}
\end{equation*}
$$

To see this, fix $\psi \in \mathcal{D}^{j+k-1}\left(\Omega_{h} \times \Omega_{v}\right)$ of the form $\psi(x, \xi)=f(x) g(x, \xi) d x_{\alpha} \wedge d \xi_{\beta}$ where $|\alpha|=j$ and $|\beta|=k-1$. The fact that $P_{j+1, k-1} T=0$ and (3.6) imply that

$$
\begin{aligned}
0=\partial T(\psi)=T(d \psi) & =T\left(\sum_{\ell} f g_{\xi_{\ell}} d \xi_{\ell} \wedge d x_{\alpha} \wedge d \xi_{\beta}\right) \\
& =(-1)^{j} \int_{\Gamma_{h}} V_{x}\left(\sum_{\ell} f g_{\xi_{\ell}} d \xi_{\ell} \wedge d \xi_{\beta}\right)\left\langle d x_{\alpha}, \tau_{h}(x)\right\rangle d \mathcal{H}^{j}(x)
\end{aligned}
$$

Similarly, since $V_{x}$ is supported in $\{x\} \times \Omega_{h}$ with $P_{1, k-1} V_{x}=0$,

$$
V_{x}\left(\sum_{\ell} f g_{\xi_{\ell}} d \xi_{\ell} \wedge d \xi_{\beta}\right)=f(x) \partial V_{x}\left(g d \xi_{\beta}\right)
$$

It follows that

$$
\int_{\Gamma_{h}} \partial V_{x}\left(g d \xi_{\beta}\right)\left\langle d x_{\alpha}, \tau_{h}(x)\right\rangle f(x) d \mathcal{H}^{j}(x)=0
$$

for all $f, \alpha$ as above, and hence that $\partial V_{x}\left(g d \xi^{\beta}\right)=0$ for $\mathcal{H}^{j}$ a.e. $x \in \Gamma_{h}$. Since $g, \beta$ were arbitrary, linearity and the fact that $V_{x}=P_{0, k} V_{x}$ imply that for every $\psi \in$ $\mathcal{D}^{k-1}\left(\Omega_{h} \times \Omega_{v}\right)$,

$$
\begin{equation*}
\partial V_{x}(\psi)=0 \text { for } \mathcal{H}^{j} \text { a.e. } x \in \Gamma_{h} \tag{3.13}
\end{equation*}
$$

Now let $\left\{\psi^{q}\right\}_{q=1}^{\infty}$ be a countable dense subset of $\mathcal{D}^{k-1}\left(\Omega_{h} \times \Omega_{v}\right)$. In view of (3.13), there is a subset of $\Gamma_{h}$ of full $\mathcal{H}^{j}$ measure, in which $\partial V_{x}\left(\psi^{q}\right)=0$ for every $q$. By density, $\partial V_{x}=0$ at every $x$ in this set, proving (3.12).
Step 5. Finally, to verify (3.7), note from (3.5) that $J p_{h} \leq 1$ a.e. in $\Gamma$, so that the coarea formula implies that

$$
\begin{aligned}
\int_{\Gamma_{h}}\left\|V_{x}\right\|(A) \mathcal{H}^{j}(d x) & =\int_{\Gamma_{h}} \int_{\Gamma_{v}(x) \cap A} m(x, \xi) \mathcal{H}^{k}(d \xi) \mathcal{H}^{j}(d x) \\
& =\int_{\Gamma \cap A} m(x, \xi) J p_{h}(x, \xi) \mathcal{H}^{j+k}(d x d \xi) \\
& \leq \int_{\Gamma \cap A} m(x, \xi) \mathcal{H}^{j+k}(d x d \xi)=\|T\|(A) .
\end{aligned}
$$

This completes the proof of the proposition.

We conclude this section with the proof of Lemma 3.1
Proof of Lemma 3.1 The rectifiability of $\Gamma_{h}$ is a special case of [5, 3.2.31]. It then follows from [5, 3.2.22] that for $\mathcal{H}^{j}$ a.e. $x \in \Gamma_{h}$, at $\mathcal{H}^{k}$ a.e. $\xi \in p_{h}^{-1}(x) \cap$ a.e. $\Gamma$, the approximate tangent space $T_{(x, \xi)} \Gamma$ is mapped by $p_{h}$ into a set of dimension at most $j$. The expression (3.2) for $\tau$ follows at such points by elementary linear algebra: given a basis $\tilde{\tau}^{1}, \ldots, \tilde{\tau}^{j+k}$ for $T_{(x, \xi)} \Gamma$, we can construct a new basis $\left\{\tau^{i}\right\}$ with the property that the horizontal parts $\left\{\tau_{h}^{i}\right\}$ are orthogonal. Since the horizontal parts span a space of dimension at most $j$, they can now be nonzero for at most $j$ of the resulting vectors. The remaining vectors, which are necessarily purely vertical, can be taken to be orthogonal by a similar argument, and can further be taken to be orthogonal to $\operatorname{span}\left(\tau_{v}^{1}, \ldots, \tau_{v}^{j}\right)$. We obtain (3.2) after relabelling and normalizing suitably.

Next, (3.4) follows by writing conclusions from [5, 3.2.22 (1); 4.3.8 (3)] in terms of the basis appearing on the right-hand side of (3.2).

Lastly, to compute $J p_{h}$, let $\left\{\tilde{\tau}^{i}\right\}_{i=1}^{j+k}$ be an orthonormal basis for $T_{(x, \xi)} \Gamma$ such that $\tilde{\tau}^{1} \wedge \cdots \wedge \tilde{\tau}^{j}=\left(\tau_{h}^{1}+\tau_{v}^{1}\right) \wedge \cdots \wedge\left(\tau_{h}^{j}+\tau_{v}^{j}\right)$, and with $\tilde{\tau}^{i}=\tau^{i}$ for $i>j$. Then $d p_{h}\left(\tilde{\tau}^{i}\right)=\tilde{\tau}_{h}^{i}=0$ for $i>j$, so the Cauchy-Binet formula yields

$$
\begin{aligned}
J p_{h}(x, \xi)^{2} & =\sum_{\alpha \in I(j, j+k)}\left|d p_{h}\left(\tilde{\tau}^{\alpha_{1}}\right) \wedge \cdots \wedge d p_{h}\left(\tilde{\tau}^{\alpha_{j}}\right)\right|^{2} \\
& =\left|d p_{h}\left(\tilde{\tau}^{1}\right) \wedge \cdots \wedge d p_{h}\left(\tilde{\tau}^{j}\right)\right|^{2}=\left|d p_{h}\left(\tau^{1}\right) \wedge \cdots \wedge d p_{h}\left(\tau^{j}\right)\right|^{2} \\
& =\left|\tau_{h}^{1} \wedge \cdots \wedge \tau_{h}^{j}\right|^{2}
\end{aligned}
$$

which is (3.5).
Finally, we have the following.
Lemma 3.4 Assume that $T=\underline{T}(\Gamma, m, \tau)$ is an i.m. rectifiable $j+k$-current in $\Omega_{h} \times \Omega_{v}$, and that $\mathbf{M}(T)+\mathbf{M}(\partial T)<\infty$. Further assume that $P_{j+1, k-1} T=0$. Let $T^{\prime}=\sum_{\ell \leq 0} P_{j+\ell, k-\ell} T$. Then $T^{\prime}$ is i.m. rectifiable.

Proof Step 1. We first claim that if $\eta=\eta^{1} \wedge \cdots \wedge \eta^{j+k}$ is any nonzero simple vector such that $P_{j+1, k-1} \eta=0$, then either $\sum_{\ell \geq 2} P_{j+\ell, k-\ell} \eta=0$ or $\sum_{\ell \leq 0} P_{j+\ell, k-\ell} \eta=0$.

To prove this, assume toward a contradiction that both $\sum_{\ell \geq 2}^{\leq} P_{j+\ell, k-\ell} \eta \neq 0$ and $\sum_{\ell \leq 0} P_{j+\ell, k-\ell} \eta \neq 0$. We write $\eta^{i}=\eta_{h}^{i}+\eta_{v}^{i}$, where $\eta_{h}^{i}$ is the horizontal part of $\eta^{i}$. By a Gram-Schmidt argument we can assume that $\eta_{h}^{i} \cdot \eta_{h}^{j}=0$ if $i \neq j$. For $\alpha \in I(q, j+k)$ we write $\eta_{h}^{\alpha}=\eta_{h}^{\alpha_{1}} \wedge \cdots \wedge \eta_{h}^{\alpha_{q}}$ and similarly $\eta_{v}^{\bar{\alpha}}$. In this notation, $\eta=\sum_{q=0}^{j+k} \sum_{|\alpha|=q} \sigma(\alpha, \bar{\alpha}) \eta_{h}^{\alpha} \wedge \eta_{v}^{\bar{\alpha}}$. The orthogonality of $\left\{\eta_{h}^{i}\right\}$ implies that all nonzero terms in the above sum are linearly independent. In particular, since $P_{j+1, k-1} \eta=0$, it follows that

$$
\begin{equation*}
\eta_{h}^{\alpha} \wedge \eta_{v}^{\bar{\alpha}}=0 \quad \text { for all } \alpha \in I(j+1, j+k) \tag{3.14}
\end{equation*}
$$

And because $\sum_{\ell \leq 0} P_{j+\ell, k-\ell} \eta \neq 0$, we may assume (after relabeling if necessary) that $\eta_{h}^{1} \wedge \cdots \wedge \eta_{h}^{j-\ell} \wedge \eta_{v}^{j-\ell+1} \wedge \cdots \wedge \eta_{v}^{j+k} \neq 0$ for some $\ell \geq 0$. The fact that
$\sum_{\ell \geq 2} P_{j+\ell, k-\ell} \eta \neq 0$ similarly implies that span $\left\{\eta_{h}^{i}\right\}$ has dimension at least $j+2$, so after further relabelling, we may assume that $\left\{\eta_{h}^{1}, \ldots, \eta_{h}^{j+2}\right\}$ are nonzero and hence (by orthogonality) linearly independent. It follows that $\eta_{h}^{1} \wedge \cdots \wedge \eta_{h}^{j+1} \wedge \eta_{v}^{j+2} \wedge \cdots \wedge \eta_{v}^{j+k} \neq$ 0 , contradicting (3.14).
Step 2. The hypothesis that $P_{j+1, k-1} T=0$ implies that $P_{j+1, k-1} \tau=0 \mathcal{H}^{j+k}$ a.e. in $\Gamma$. Thus Step 1 implies that $\sum_{\ell \leq 0} P_{j+\ell, k-\ell} \tau$ equals either $\tau$ or 0 , a.e. in $\Gamma$. From this it is clear that $T^{\prime}=T\left\llcorner\Gamma^{\prime}=\underline{T}\left(\Gamma^{\prime}, m, \tau\right)\right.$, where $\Gamma^{\prime} \subset \Gamma$ is the set of points at which $\sum_{\ell \leq 0} P_{j+\ell, k-\ell} \tau=\tau$.

## 4 Fu's Theorem Revisited

In this section we prove the following.
Theorem 4.1 Suppose that $\Omega \subset \mathbb{R}^{n}$ is an open set and that $T$ is an i.m. rectifiable Lagrangian current in $\Omega \times \mathbb{R}^{n}$ such that $\partial T=0$ in $\Omega \times \mathbb{R}^{n}$, and moreover that $\|T\|\left(K \times \mathbb{R}^{n}\right)<\infty$ for every compact $K \subset \Omega$.

If $P_{n-k, k} T=0$, then $P_{n-\ell, \ell} T=0$ for all $\ell>k$. In particular, if $P_{n, 0} T=0$, then $T=0$.

Fu [7] proved the $k=0$ case (which is the main case) of the same result with the stronger hypothesis that $T$ is locally vertically bounded, which means that whenever $K \subset \Omega$ is compact, $T\left\llcorner\left(K \times \mathbb{R}^{n}\right)\right.$ has compact support in $\mathbb{R}^{n} \times \mathbb{R}^{n}$.

The theorem implies that an i.m. rectifiable Lagrangian current in $\Omega \times \mathbb{R}^{n}$ is determined by its "most horizontal" part. In particular, we have the following.
Corollary 4.2 Let $T_{1}, T_{2}$ be Lagrangian rectifiable currents in $\Omega \times \mathbb{R}^{n}$ with no boundary in $\Omega \times \mathbb{R}^{n}$, and such that $\left\|T_{i}\right\|\left(K \times \mathbb{R}^{n}\right)<\infty$ for every compact $K \subset \Omega$, for $i=1,2$.

If $P_{n, 0} T_{1}=P_{n, 0} T_{2}$, then $T_{1}=T_{2}$. In particular, if $u$ is a Monge-Ampère function, then there is a unique current $[d u]$ satisfying (2.5)-(2.8)
Proof Apply Theorem 4.1, with $k=0$, to $T=T_{1}-T_{2}$. Uniqueness of the current [ $d u$ ] for a Monge-Ampère function follows immediately.

We have already noted in Section 2.5 that if $u$ is $C^{2}$, then

$$
\begin{equation*}
[d u]\left(\phi d x_{\alpha} \wedge d \xi_{\beta}\right)=\sigma(\alpha, \bar{\alpha}) \int_{\Omega} \phi(x, D u(x)) M^{\bar{\alpha} \beta}\left(D^{2} u(x)\right) d x \tag{4.1}
\end{equation*}
$$

is an i.m. rectifiable current satisfying (2.5)-(2.8), hence the unique such current, in view of the above corollary. An approximation argument then shows that (4.1) continues to hold for $u \in W_{\text {loc }}^{2, n}(\Omega)$. Motivated by (4.1), given a Monge-Ampère function $u: \Omega \rightarrow \mathbb{R}$ we define signed measures $\mu^{\bar{\alpha} \beta}\left(D^{2} u\right)$ in $\Omega \times \mathbb{R}^{n}$ by

$$
\begin{equation*}
\int_{\Omega \times \mathbb{R}^{n}} \phi(x, \xi) \mu^{\bar{\alpha} \beta}\left(D^{2} u\right)(d x, d \xi)=\sigma(\alpha, \bar{\alpha})[d u]\left(\phi d x_{\alpha} \wedge d \xi_{\beta}\right) \tag{4.2}
\end{equation*}
$$

In view of (4.1), these measures correspond to the minors of $D^{2} u$. In particular $\int \phi d \mu^{\bar{\alpha} \beta}\left(D^{2} u\right)=\int \phi(x, D u) M^{\bar{\alpha} \beta}\left(D^{2} u\right) d x$ whenever $u \in W_{\text {loc }}^{2, n}$. These measures possess good weak continuity properties.

Corollary 4.3 If $u_{k}$ is a sequence of Monge-Ampère functions on a domain $\Omega \subset \mathbb{R}^{n}$ and if $\left\|\left[d u_{k}\right]\right\|\left(K \times \mathbb{R}^{n}\right) \leq C_{K}$ for $K$ compact, and $u_{k} \rightarrow u$ in $L_{\text {loc }}^{1}$, then $u$ is MongeAmpère with $\|[d u]\|(O) \leq \lim \sup _{k}\left\|\left[d u_{k}\right]\right\|(O)$ for every open $O \subset \Omega \times \mathbb{R}^{n}$. Moreover, for every $\alpha, \beta$ such that $|\alpha|+|\beta|=n, \mu^{\bar{\alpha} \beta}\left(D^{2} u_{k}\right) \rightharpoonup \mu^{\bar{\alpha} \beta}\left(D^{2} u\right)$ weakly as measures.

An analogous result is established in [7] with the additional hypothesis that

$$
\left\|D u_{k}\right\|_{L^{\infty}(K)} \leq C_{K}
$$

for $K$ compact.
Proof It follows from the compactness theorem for integral currents (see [10, §2.2.4, Theorem 2] for a version adapted to the present setting) that there exists an i.m. rectifiable current $T$ such that $\left[d u_{k}\right] \rightharpoonup T$ weakly in $\mathcal{D}_{n}\left(\Omega \times \mathbb{R}^{n}\right)$. The fact that [ $d u_{k}$ ] satisfies (2.5)-(2.8) for every $k$, together with the assumed local uniform mass bounds on $\left[d u_{k}\right]$ and standard facts about weak lower semicontinuity, imply that $T$ satisfies (2.5)-(2.8), with $\|T\|(O) \leq \lim \sup _{k}\left\|\left[d u_{k}\right]\right\|(O)$ for every open $O \subset \Omega \times \mathbb{R}^{n}$. Hence $u$ is Monge-Ampère and $T=[d u]$.

Remark 2 The corollary implies that if $\left\{u_{k}\right\}$ is a sequence of $C^{\infty}$ functions on a domain $\Omega$ such that $\left\|M^{\bar{\alpha} \beta}\left(D^{2} u_{k}\right)\right\|_{L^{1}(K)} \leq C_{K}$ for every $\alpha, \beta$ and every compact $K \subset \Omega$, then the $L^{1}$ limit of any convergent subsequence is Monge-Ampère. It is not known whether every Monge-Ampère function arises in this way.

Remark 3 When $\alpha=\beta=(1, \ldots, n)$ we will write $\operatorname{det} D^{2} u$ instead of $\mu^{\alpha \beta}\left(D^{2} u\right)$. Let us temporarily write $\operatorname{det}^{\prime} D^{2} u$ to denote the distributional determinant, when it exists, (see, for example, $[6,12]$ ). Note that as we have defined it, $\operatorname{det} D^{2} u$ is a measure in the product space $\Omega \times \mathbb{R}^{n}$, whereas $\operatorname{det}^{\prime} D^{2} u$ is a distribution on $\Omega$.

If $u$ is Monge-Ampère and $\operatorname{det}^{\prime} D^{2} u$ is well defined, it is natural to ask whether $\operatorname{det}^{\prime} D^{2} u$ is a measure and

$$
\int_{\Omega} \phi(x) \operatorname{det}^{\prime} D^{2} u(d x)=\int_{\Omega \times \mathbb{R}^{n}} \phi(x) \operatorname{det} D^{2} u(d x, d \xi)
$$

This holds if $u$ is smooth, and by approximation if $u$ is a limit of smooth functions in the sense of Corollary 4.3 and in addition $u$ belongs to what is called in [6] an admissible domain for $\operatorname{det}^{\prime} D^{2} u$.

Remark 4 As noted in the introduction, Fu [7] deduced from his version of Theorem 4.1 a uniqueness theorem for Legendrian cycles. Going through the same argument but taking our stronger version of Theorem4.1 as a starting point, we end up with a uniqueness theorem for Legendrian cycles with finite mass, whereas Fu's version of the theorem instead applies to compactly supported Legendrian cycles. The finite mass assumption at first sight appears a bit weaker, but in fact (together with other hypotheses) implies compact support, so here we do not gain any new generality.

Proof of Theorem4.1 Step 1. It clearly suffices to show that if

$$
\begin{equation*}
P_{n-k+1, k-1} T=0 \quad \text { for some } k, 1 \leq k \leq n \tag{4.3}
\end{equation*}
$$

and if $\Omega^{\prime} \subset \Omega$ is an open subset with compact closure, then $P_{n-k, k} T=0$ in $\Omega^{\prime} \times \mathbb{R}^{n}$. Replacing $\Omega$ by $\Omega^{\prime}$ for convenience, we assume that $T$ is a rectifiable current with finite mass and no boundary in $\Omega \times \mathbb{R}^{n}$, and that (4.3) holds.

We apply Lemma3.1 with $j=n-k$, and we use the notation from that lemma. In view of (4.3), Proposition 3.2 asserts that there exists an $n-k$-dimensional rectifiable set $\Gamma_{h} \subset \Omega$ with unit tangent $n-k$-vectorfield $\tau_{h}: \Gamma_{h} \rightarrow \bigwedge_{n-k} \mathbb{R}^{n}$, and for $\mathcal{H}^{n-k}$ a.e. $x \in \Gamma_{h}$, a vertical i.m. rectifiable $k$-current $V_{x}$, such that

$$
\begin{equation*}
T\left(\phi d x_{\alpha} \wedge d \xi_{\beta}\right)=\int_{\Gamma_{h}} V_{x}\left(\phi d \xi_{\beta}\right)\left\langle d x_{\alpha}, \tau_{h}(x)\right\rangle d \mathcal{H}^{j}(x) \tag{4.4}
\end{equation*}
$$

Appealing again to Proposition 3.2 and the definition of Lagrangian currents (from Section 2.4), we find that the following hold at $\mathcal{H}^{n-k}$ a.e. $x \in \Gamma_{h}$ :

- $T_{x} \Gamma_{h}$ exists;
- $V_{x}$ is i.m. rectifiable and is carried by $\Gamma_{v}(x)$ (defined in (3.3)), with $\partial V_{x}=0$;
- at $\mathcal{H}^{k}$ a.e. $(x, \xi) \in \Gamma_{v}(x)$, (3.2) and (3.4) hold and $T_{(x, \xi)} \Gamma$ is Lagrangian.

Step 2. For $x$ satisfying the above conditions, we claim that either $\mathbf{M}\left(V_{x}\right)=0$ or $\mathbf{M}\left(V_{x}\right)=+\infty$. We may assume by choosing coordinates suitably that $T_{x} \Gamma_{h}=$ $\operatorname{span}\left\{e_{1}, \ldots, e_{n-k}\right\}$. The Lagrangian condition implies that at $\mathcal{H}^{k}$ a.e. $(x, \xi) \in \Gamma_{v}(x)$, if $i \leq n-k$ and $i^{\prime}>n-k$, then

$$
0=\omega\left(\tau^{i} \wedge \tau^{i^{\prime}}\right)=\tau^{i} \cdot J \tau^{i^{\prime}}=-J \tau_{h}^{i} \cdot \tau_{v}^{i^{\prime}}
$$

(since $\tau^{i^{\prime}}=\tau_{v}^{i^{\prime}}$ ). Then this says that $\tau_{v}^{i^{\prime}} \cdot \varepsilon_{\ell}=0$ for all $\ell \leq n-k$, so that $T_{(x, \xi)} \Gamma_{v}(x)=$ $\operatorname{span}\left\{\tau_{v}^{i^{\prime}}: i^{\prime}>n-k\right\}=\operatorname{span}\left\{\varepsilon_{n-k+1}, \ldots, \varepsilon_{n}\right\}$ at $\mathcal{H}^{k}$ a.e. $(x, \xi) \in \Gamma_{v}(x)$. In particular, the approximate tangent space to $\Gamma_{v}(x)$ is a.e. constant.

We now demonstrate that this implies that $V_{x}$ is a union of $k$-planes with integer multiplicities. Indeed, it follows from the above that $V_{x}\left(\phi d \xi^{\beta}\right)=0$ for $\beta \in I(k, n)$, unless $\beta=(n-k+1, \ldots, n)$. For any $q \leq n-k$, let $p_{q}(x, \xi)=\xi_{q}$; it follows that $\left\langle V_{x}, p_{q}, s\right\rangle=0$ for $\mathcal{L}^{1}$ a.e. $s \in \mathbb{R}$, and hence (as in the proof of Lemma3.3) via [19, Lemma 3.3] that the indecomposable components of $V_{x}$ are contained in level sets of $p_{q}$. This holds for every $q \leq n-k$, for each indecomposable component $V_{x, i}$ of $V_{x}$, so there exists $r_{1}^{i}, \ldots, r_{n-k}^{i}$ such that $V_{x, i}$ is supported in the $k$-plane $P_{x, i}:=$ $\left\{(x, \xi): \xi_{q}=r_{q}^{i}, q=1, \ldots, n-k\right\}$. Then recalling that $\partial V_{x}=0$, we infer from the constancy theorem that each $V_{x, i}$ corresponds to integration over the $k$-plane $P_{x, i}$, with an integer multiplicity and suitable orientation.

It follows that either $\mathbf{M}\left(V_{x}\right)=0$ or $\mathbf{M}\left(V_{x}\right)=+\infty$ as claimed.
Step 3. However, it is clear from (3.7) that for any compact $K \subset \Omega$,

$$
\int_{\Gamma_{h}} \mathbf{M}\left(V_{x}\right) d \mathcal{H}^{j}(x) \leq\|T\|\left(\mathbb{R}^{n}\right)<\infty
$$

so that $\mathbf{M}\left(V_{x}\right)<\infty$ for $\mathcal{H}^{j}$ a.e. $x$. Consequently $\mathbf{M}\left(V_{x}\right)=0$, and hence $V_{x}=0$, for $\mathcal{H}^{j}$ a.e. $x$. It then follows immediately from (4.4) that $T\left(\phi d x_{\alpha} \wedge d \xi_{\beta}\right)=0$, whenever $|\alpha|=n-k,|\beta|=k$. In other words, $P_{n-k, k} T=0$.

## 5 Description of $P_{n-1,1}[d u]$

The main result of this section has two essentially equivalent forms, the first of which is the following.

Proposition 5.1 Suppose that $u$ is a Monge-Ampère function on an open set $\Omega \subset \mathbb{R}^{n}$. Then

$$
\begin{align*}
{[d u]\left(\phi d x_{i} \wedge d \xi_{j}\right) } & =\int_{\Omega \backslash \mathcal{I}_{D u}} \phi(x, D u(x)) \sigma(\bar{i}, i) u_{x_{i} x_{j}}(d x)  \tag{5.1}\\
& +\int_{\mathcal{J}_{D u}}\left(\int_{0}^{1} \phi\left(x, \theta D u^{+}+(1-\theta) D u^{-}\right) d \theta\right) \sigma(\bar{i}, i) u_{x_{i} x_{j}}(d x)
\end{align*}
$$

for every smooth $\phi$ with compact support in $\Omega \times \mathbb{R}^{n}$.
Recall that if $u$ is Monge-Ampère, then $D^{2} u$ is a measure. The Cantor part of $D^{2} u$ is contained in the first term on the right-hand side. The difficulty in proving an analogous result for $P_{n-k, k}[d u], k \geq 2$, lies partly in dealing with the Cantor part of $k \times k$ minors of $D^{2} u$.

Remark 5 Recall from (4.2) that we write $\mu^{i j}$ to denote the measure in $\Omega \times \mathbb{R}^{n}$ defined by $\int \phi \mu^{i j}(d x, d \xi)=\sigma(\bar{i}, i)[d u]\left(\phi d x_{\bar{i}} \wedge d \xi_{j}\right)$. The main result of [14] implies that among all matrix-valued measures $\left(\nu^{i j}\right)$ satisfying

$$
\int_{\Omega} \phi_{x_{i}}(x, D u(x)) d x+\int_{\Omega \times \mathbb{R}^{n}} \phi_{\xi_{j}}(x, \xi) \nu^{i j}(d x, d \xi)=0 \quad \forall \phi \in C_{c}^{\infty}\left(V \Omega \times \mathbb{R}^{n}\right),
$$

there is a unique measure with minimal mass. It turns out that the minimizing measure is exactly $\left(\mu^{i j}\right)$; this follows from combining (5.1) with results from [14]. Note that $\left(\mu^{i j}\right)$ satisfies the above identity as a consequence of the fact that $\partial[d u]=0$.

Proposition 5.1 is very closely related to Proposition 5.3 below, which gives a description of certain 1-dimensional slices of $[d u]$. In order to state the latter result, it is convenient to use the following.

Lemma 5.2 For an interval $(a, b) \subset \mathbb{R}$, suppose that $v:(a, b) \rightarrow \mathbb{R}^{k}$ is a function of bounded variation, with total variation L. Define a 1 -current $I_{v}^{*}$ in $\mathbb{R}^{k}$ by

$$
\begin{align*}
I_{v}^{*}\left(\phi^{j} d y^{j}\right)= & \int_{(a, b) \backslash \mathcal{J}_{v}} \phi^{j}(v(s)) v_{j}^{\prime}(d s)  \tag{5.2}\\
& +\int_{\mathcal{J}_{v}}\left(\int_{0}^{1} \phi^{j}\left(\theta v^{+}(s)+(1-\theta) v^{-}(s)\right) d \theta\right) v_{j}^{\prime}(d s)
\end{align*}
$$

Then there exists a Lipschitz curve $\gamma:(0, L) \rightarrow \mathbb{R}^{k}$ such that

$$
\begin{equation*}
I_{v}^{*}\left(\phi^{j} d y^{j}\right)=\int_{0}^{L} \phi^{j}(\gamma(s)) \gamma_{j}^{\prime}(s) d s \tag{5.3}
\end{equation*}
$$

In particular, $I_{v}^{*}$ is an integral 1-current, and

$$
\partial I_{v}^{*}(\psi)=\lim _{s \nearrow b} \psi(v(s))-\lim _{t \searrow a} \psi(v(t))=\psi(\gamma(L))-\psi(\gamma(0)) .
$$

Any integral 1-current can be represented as a sum of terms having the same form as the right-hand side of (5.3); here only one such term is needed.

Note that $I_{v}^{*}$ is the current corresponding to the image of $v$, with jumps "filled in" in the simplest possible way.
Proof If $v$ is $C^{1}$, then $I_{v}^{*}$ is just the current associated with integration over the image of $v$, and the conclusions are clear. If not, let $v_{\ell}:(a, b) \rightarrow \mathbb{R}^{k}$ be $C^{1}$ functions such that $v_{\ell} \rightarrow v$ strictly in BV , which means that $v_{\ell} \rightarrow v$ in $L^{1}$ and

$$
L_{\ell}:=\int_{(a, b)}\left|v_{\ell}^{\prime}\right| d x \rightarrow L:=\left\|v^{\prime}\right\|((a, b))
$$

It is well known that such sequences exist. It then follows from [14, Theorem 1.1] that $I_{v_{\ell}}^{*} \rightarrow I_{v}^{*}$ weakly. On the other hand, let $\gamma_{\ell}$ be an arclength reparametrization of $v_{\ell}$, so that $\gamma_{\ell}(s)=v_{\ell}(\sigma(s))$, where $\sigma:(a, b) \rightarrow\left(0, L_{\ell}\right)$ is nondecreasing and is chosen so that $\left|\gamma_{\ell}^{\prime}\right|=1$ a.e. Let $\gamma$ be a limit of a uniformly convergent subsequence. Then (5.3) and the assertions about $\partial I_{v}^{*}$ follow by passing to limits from the corresponding statements for $v_{\ell}, \gamma_{\ell}$.

Using notation from the lemma, we can state the following.
Proposition 5.3 Let $U \subset \mathbb{R}^{n}$ be an open subset, and let $f: U \rightarrow \mathbb{R}^{n}$ be a $C^{1,1}$ diffeomorphism onto its image, which we call $V$, with inverse $g: V \rightarrow U$. Also, let $q\left(y_{1}, \ldots, y_{n}\right):=\left(y_{1}, \ldots, y_{n-1}\right)=y^{\prime}$.

If $u$ is a Monge-Ampère function on $U$, then

$$
\begin{equation*}
\left\langle[d u], q \circ f \circ p_{h}, y^{\prime}\right\rangle=I_{w\left(\cdot ; y^{\prime}\right)}^{*} \quad \text { for } \mathcal{L}^{n-1} \text { a.e. } y^{\prime} \in \mathbb{R}^{n-1} \tag{5.4}
\end{equation*}
$$

in the notation of Lemma 5.2 for $w\left(y_{n} ; y^{\prime}\right):=\left(g\left(y^{\prime}, y_{n}\right), D u\left(g\left(y^{\prime}, y_{n}\right)\right)\right)$.
For $\mathcal{L}^{n-1}$ a.e. $y^{\prime} \in \mathbb{R}^{n-1}$, the level set $(q \circ f)^{-1}\left(y^{\prime}\right)$ is a Lipschitz curve (or union of Lipschitz curves) in $U$ and $\left\langle[d u], q \circ f \circ p_{h}, y^{\prime}\right\rangle$ is the slice of [du] sitting above this curve. Note also that $y_{n} \mapsto g\left(y^{\prime}, y_{n}\right)$ is a parametrization of $(q \circ f)^{-1}\left(y^{\prime}\right)$, so that the current $I_{w\left(\cdot ; y^{\prime}\right)}^{*}$ on the left-hand side of (5.4) corresponds to the graph of Du above $(q \circ f)^{-1}\left(y^{\prime}\right)$, with jumps filled in in the natural way.

We start with the proof of Proposition 5.1, which mostly amounts to establishing Proposition 5.3 in the special case $f(x)=x$.

Proof of Proposition 5.1 We suppose that $(\Gamma, m, \tau)$ are an $n$-rectifiable set, integervalued multiplicity function, and orienting unit tangent $n$-vectorfield such that $[d u]=\underline{T}(\Gamma, m, \tau)$. By using a partition of unity, we see that it suffices to consider test functions $\phi$ supported in $B \times \mathbb{R}^{n}$, where $B \subset \Omega$ is a ball. Thus in fact we can assume that $\Omega$ is an open ball and that $\mathbf{M}([d u])<\infty$.

It suffices to prove the proposition for $i=n$. Note that $\sigma(\bar{n}, n)=1$, so the signs will vanish from our calculations.

Given $x=\left(x_{1}, \ldots, x_{n}\right) \in \mathbb{R}^{n}$, we will write $x=\left(x^{\prime}, x_{n}\right)$ with $x^{\prime}=\left(x_{1}, \ldots, x_{n-1}\right)$. We define $q: \Omega \rightarrow \mathbb{R}^{n-1}$ by $q(x)=x^{\prime}$, and $Q(x, \xi)=q(x)$. Throughout the proof we will write $J Q$ to denote the Jacobian $J_{\Gamma \rightarrow \mathbb{R}^{n-1}} Q$. Because we have assumed that $\Omega$ is a ball, $q^{-1}\left(x^{\prime}\right)$ is connected for every $x^{\prime}$.

Let $\left\langle[d u], Q, x^{\prime}\right\rangle$ denote as usual the slice of $[d u]$ by the level set $Q^{-1}\left(x^{\prime}\right)$. Recall that

$$
\begin{equation*}
[d u]\left(\phi d x_{\bar{n}} \wedge d \xi_{j}\right)=\int_{\mathbb{R}^{n-1}}\left\langle[d u], Q, x^{\prime}\right\rangle\left(\phi d \xi_{j}\right) \mathcal{L}^{n-1}\left(d x^{\prime}\right) \tag{5.5}
\end{equation*}
$$

So it suffices to describe $\mathcal{L}^{n-1}$ a.e. slice $\left\langle[d u], Q, x^{\prime}\right\rangle$. We will do this as follows.
First, we record a number of properties that $\left\langle[d u], Q, x^{\prime}\right\rangle$ inherits, for $\mathcal{L}^{n-1}$ a.e. $x^{\prime}$, from the defining attributes (2.5)-(2.8) of $[d u]$. Most important are tangent properties, which follow from the crucial Lagrangian assumption.

Next, we write down a family $\left\{R_{x^{\prime}}\right\}$ of integral 1-currents with the property that the right-hand side of (5.1) is exactly $\int_{\mathbb{R}^{n-1}} R_{x^{\prime}}\left(\phi d \xi_{j}\right) \mathcal{L}^{n-1}\left(d x^{\prime}\right)$. Thus to prove the theorem we must show that $R_{x^{\prime}}=\left\langle[d u], Q, x^{\prime}\right\rangle$ almost everywhere. Toward this end, we deduce a number of properties of a.e. $R_{x^{\prime}}$, like those already found for the slices of $[d u]$.

Finally, we define $S_{x^{\prime}}=\left\langle[d u], Q, x^{\prime}\right\rangle-R_{x^{\prime}}$, and we argue that $S_{x^{\prime}}=0$ for a.e. $x^{\prime}$. This is similar in spirit to the proof of the Uniqueness Theorem 4.1.
Step 1. Properties of a.e. slice. For $\mathcal{L}^{n-1}$ a.e. $x^{\prime} \in \mathbb{R}^{n-1}$, since $\partial[d u]=0$ and $\mathbf{M}([d u])<\infty$, it follows from $[5,4.3 .2]$ that $\partial\left\langle[d u], Q, x^{\prime}\right\rangle=0$ and $\left\langle[d u], Q, x^{\prime}\right\rangle$ is i.m. rectifiable, with finite mass. We claim that in addition, for $\mathcal{L}^{n-1}$ a.e. $x^{\prime} \in \mathbb{R}^{n-1}$,

$$
\begin{equation*}
\left\langle[d u], Q, x^{\prime}\right\rangle\left(\phi d x_{i}\right)=\delta_{i n} \int_{q^{-1}\left(x^{\prime}\right)} \phi(x, D u) \mathcal{H}^{1}(d x) \tag{5.6}
\end{equation*}
$$

for all $\phi \in C_{c}^{\infty}(\Omega)$. To see this, we use (2.8) to compute

$$
\begin{aligned}
\int_{\mathbb{R}^{n-1}} & \left\langle[d u], Q, x^{\prime}\right\rangle\left(\phi d x_{i}\right) \psi\left(x^{\prime}\right) \mathcal{L}^{n-1}\left(d x^{\prime}\right)=[d u]\left(Q^{\#}\left(\psi(\cdot) d x^{\prime}\right) \wedge \phi d x_{i}\right) \\
& =\int_{\mathbb{R}^{n}} \phi(x, D u) \psi \circ q(x) d x_{\bar{n}} \wedge d x_{i} \\
& =\delta_{i n} \int_{\mathbb{R}^{n}} \phi(x, D u) \psi\left(x^{\prime}\right) \mathcal{L}^{n}(d x) \\
& =\delta_{i n} \int_{\mathbb{R}^{n-1}}\left(\int_{q^{-1}\left(x^{\prime}\right)} \phi(x, D u) \mathcal{H}^{1}(d x)\right) \psi\left(x^{\prime}\right) \mathcal{L}^{n-1}\left(d x^{\prime}\right)
\end{aligned}
$$

Since this holds for all $\psi \in C_{c}^{\infty}\left(\mathbb{R}^{n-1}\right)$ and $\phi$ as above, it implies (5.6).
Another fact from [5] is that for $\mathcal{L}^{n-1}$ a.e. $x^{\prime},\left\langle[d u], Q, x^{\prime}\right\rangle$ has the explicit representation

$$
\begin{equation*}
\left\langle[d u], Q, x^{\prime}\right\rangle=\int_{\Gamma\left(x^{\prime}\right)}\langle\phi, \zeta\rangle m d \mathcal{H}^{1} \tag{5.7}
\end{equation*}
$$

where

$$
\begin{equation*}
\Gamma\left(x^{\prime}\right):=Q^{-1}\left(x^{\prime}\right) \cap\{(x, \xi) \in \Gamma: J Q>0\} \tag{5.8}
\end{equation*}
$$

and for $(x, \xi) \in \Gamma\left(x^{\prime}\right)$,

$$
\begin{equation*}
\zeta(x, \xi) \text { is a unit vector in } T_{(x, \xi)} \Gamma \text { such that } d Q(x, \xi)(\zeta)=0 \tag{5.9}
\end{equation*}
$$

Recall that $\operatorname{JQ}(x, \xi)>0$ exactly when $d Q(x, \xi): T_{(x, \xi)} \Gamma \rightarrow \mathbb{R}^{n-1}$ is of full rank. Thus (5.8) implies that $\zeta$ is uniquely specified, up to a sign, by (5.9). Properties (5.7)-(5.9) follow directly from [5, 4.3.8], where the sign of $\zeta$ is also specified; we will not need to keep track of this sign in our later arguments.

Following notation in Lemma3.1 (with $k=1, j=n-1$ ), we will write

$$
\begin{aligned}
\Gamma_{h} & :=\left\{x \in \Omega: \mathcal{H}^{1}\left(p_{h}^{-1}(x) \cap \Gamma\right)>0\right\} \\
\Gamma_{h}\left(x^{\prime}\right) & :=\left\{x \in \Omega: \mathcal{H}^{1}\left(p_{h}^{-1}(x) \cap \Gamma\left(x^{\prime}\right)\right)>0\right\}
\end{aligned}
$$

Clearly $\Gamma_{h}\left(x^{\prime}\right) \subset q^{-1}\left(x^{\prime}\right) \cap \Gamma_{h}$, and the fact that $\Gamma_{h}$ is $n-1$-rectifiable (see Lemma3.1) implies that for $\mathcal{L}^{n-1}$ a.e. $x^{\prime}, \Gamma_{h}\left(x^{\prime}\right)$ is at most countable and $T_{x} \Gamma_{h}$ exists at every $x \in \Gamma_{h}\left(x^{\prime}\right)$. We finally claim that for $\mathcal{L}^{n-1}$ a.e. $x^{\prime}$, at every $x \in \Gamma_{h}\left(x^{\prime}\right)$,

$$
\begin{equation*}
\zeta(x, \xi) \in \mathbb{J}\left(T_{x} \Gamma_{h}^{\perp}\right) \quad \text { for } \mathcal{H}^{1} \text { a.e. } \xi \in \Gamma\left(x^{\prime}\right) \cap p_{h}^{-1}(x) \tag{5.10}
\end{equation*}
$$

where $T_{x} \Gamma_{h}{ }^{\perp}$ denotes the orthogonal complement in the horizontal directions only, so that $T_{x} \Omega=T_{x} \Gamma_{h} \oplus T_{x} \Gamma_{h}{ }^{\perp}$.

Observe that $\zeta(x, \cdot)$ is independent of $\xi$, up to a sign, for $x$ satisfying (5.10).
To prove (5.10), observe that, for $\mathcal{L}^{n-1}$ a.e. $x^{\prime}$, the Lagrangian condition (2.6) and Lemma 3.1 imply that at every $x \in \Gamma_{h}\left(x^{\prime}\right)$,

$$
\begin{equation*}
\text { for } \mathcal{H}^{1} \text { a.e. } \xi \in p_{h}^{-1}(x) \cap \Gamma \text {, (3.2) holds and } T_{(x, \xi)} \Gamma \text { is Lagrangian, } \tag{5.11}
\end{equation*}
$$

and

$$
\begin{equation*}
\text { (3.4) holds (with } j=n-1, k=1 \text { ) a.e. on the set } \Gamma_{v}(x) \text {, see (3.3). } \tag{5.12}
\end{equation*}
$$

So it suffices to prove that (5.10) holds at any $x$ where (5.11), (5.12) are verified.
Fix such an $x$, and fix also $\xi \in p_{h}^{-1}(x) \cap \Gamma_{h}\left(x^{\prime}\right)$ at which (3.2) holds and where the tangent $n$-plane is Lagrangian. Since $Q=q \circ p_{h}$, the vector $\tau^{n}=\tau_{v}^{n} \in T_{(x, \xi)} \Gamma$ appearing on the right-hand side of (3.2) satisfies $d Q(x, \xi) \tau_{v}^{n}=0$. Because this
equation defines $\zeta(x, \xi)$ up to a sign, we find that $\zeta(x, \xi)= \pm \tau_{v}^{n}(x, \xi)$. It follows also that $\tau_{h}^{1}, \ldots, \tau_{h}^{n-1}$ appearing on the right-hand side of (3.2) are nonzero (since otherwise $d Q(x, \xi)$ would have a nullspace of dimension at least two) and hence that $(x, \xi) \in \Gamma_{v}(x)$ as defined in (3.3). This in turn implies that (3.4) holds, and then (as in the proof of Theorem 4.1) the Lagrangian condition implies that (5.10) is satisfied.
Step 2. Definition and properties of $R_{x^{\prime}}$. In this step we appeal to results about one-dimensional sections of BV functions as found for example in [5, $\S 4.5$ ] or [1, Chapter 3].

We henceforth identify $D u$ with its precise representative; see [1, Corollary 3.80] which in particular implies that $D u(x)$ equals its Lebesgue value whenever $x$ is a Lebesgue point of $D u$. For $x^{\prime} \in \mathbb{R}^{n-1}$, we define

$$
\Omega_{x^{\prime}}=\left\{x_{n} \in \mathbb{R}:\left(x^{\prime}, x_{n}\right) \in \Omega\right\}
$$

We will write $D u\left(x_{n} ; x^{\prime}\right)=D u\left(x^{\prime}, x_{n}\right)$, so that we view $D u\left(\cdot ; x^{\prime}\right): \Omega_{x^{\prime}} \rightarrow \mathbb{R}^{n}$ as functions of a single variable, parametrized by $x^{\prime} \in \mathbb{R}^{n-1}$. Then for $\mathcal{L}^{n-1}$ a.e. $x^{\prime}$, $D u\left(\cdot ; x^{\prime}\right)$ is a BV function on $\Omega_{x^{\prime}}$. We will write $\partial_{x_{n}} D u\left(x^{\prime}\right)$ to indicate the associated vector-valued derivative measure on $\Omega_{x^{\prime}}$, and we write $\partial_{x_{n}} u_{x_{j}}\left(x^{\prime}\right), j=1, \ldots, n$ for the components of $\partial_{x_{n}} D u\left(x^{\prime}\right)$. We define a BV function $v\left(\cdot ; x^{\prime}\right): \Omega_{x^{\prime}} \rightarrow \mathbb{R}^{n} \times \mathbb{R}^{n}$, given by

$$
x_{n} \mapsto\left(x^{\prime}, x_{n}, D u\left(x_{n} ; x^{\prime}\right)\right)=v\left(x_{n} ; x^{\prime}\right)
$$

and a 1-current $R_{x^{\prime}}=I_{v\left(\cdot ; x^{\prime}\right)}^{*}$, using notation from Lemma.5.2. We will write $\mathcal{J}_{x^{\prime}}$ to denote the jump set $D u\left(\cdot, x^{\prime}\right)$, which clearly coincindes with the jump set of $v\left(\cdot ; x^{\prime}\right)$. From the definition in Lemma 5.2 we check that

$$
\begin{aligned}
R_{x^{\prime}}\left(\phi d x_{i}\right)= & \delta_{i n} \int_{\Omega_{x^{\prime}}} \phi\left(v\left(x_{n} ; x^{\prime}\right)\right) \mathcal{L}^{1}\left(d x_{n}\right) \\
R_{x^{\prime}}\left(\phi d \xi_{j}\right)= & \int_{\Omega_{x^{\prime}} \backslash \mathcal{J}_{x^{\prime}}} \phi\left(v\left(x_{n} ; x^{\prime}\right)\right) \partial_{x_{n}} u_{x_{j}}\left(x^{\prime}\right)\left(d x_{n}\right) \\
& +\int_{\mathcal{J}_{x^{\prime}}}\left(\int_{0}^{1} \phi\left(\theta v^{+}+(1-\theta) v^{-}\right) d \theta\right) \partial_{x_{n}} u_{x_{j}}\left(x^{\prime}\right)\left(d x_{n}\right)
\end{aligned}
$$

where in the last integral, we write $v^{ \pm}:=\lim _{s \rightarrow x_{n}^{ \pm}} v\left(s ; x^{\prime}\right)$ for $x_{n} \in \mathcal{J}_{x^{\prime}}$.
It follows from Lemma 5.2 that $R_{x^{\prime}}$ is an integral current with $\partial R_{x^{\prime}}=0$ in $\Omega \times \mathbb{R}^{n}$ and $\mathbf{M}\left(R_{x^{\prime}}\right)<\infty$ for $\mathcal{L}^{n-1}$ a.e. $x^{\prime}$. We also claim that

$$
\begin{equation*}
\text { the right-hand side of (5.1) }=\int_{\mathbb{R}^{n-1}} R_{x^{\prime}}\left(\phi d \xi_{j}\right) \mathcal{L}^{n-1}\left(d x^{\prime}\right) \tag{5.13}
\end{equation*}
$$

To verify this, note that $(x, D u(x))=v\left(x_{n} ; x^{\prime}\right)$, so that upon comparing (5.1) and the formulas given above for $R_{x^{\prime}}$, we see that (5.13) follows from the facts that when $f, g: \Omega \rightarrow \mathbb{R}$ are bounded and $D^{2} u$-measurable,

$$
\int_{\Omega \backslash \mathcal{I}_{D u}} f(x) u_{x_{j} x_{n}}(d x)=\int_{\mathbb{R}^{n-1}} \int_{\Omega_{x^{\prime}} \backslash \partial_{x^{\prime}}} f\left(x^{\prime}, x_{n}\right) \partial_{x_{n}} u_{x_{j}}\left(x^{\prime}\right)\left(d x_{n}\right) \mathcal{L}^{n-1}\left(d x^{\prime}\right)
$$

and

$$
\int_{\mathcal{D}_{D u}} g(x) u_{x_{j} x_{n}}(d x)=\int_{\mathbb{R}^{n-1}} \int_{\mathcal{J}_{x^{\prime}}} g\left(x^{\prime}, x_{n}\right) \partial_{x_{n}} u_{x_{j}}\left(x^{\prime}\right)\left(d x_{n}\right) \mathcal{L}^{n-1}\left(d x^{\prime}\right)
$$

Proofs of these identities can be found, for example, in [1, Theorems 3.107, 3.108].
We note for future reference that for $\mathcal{L}^{n-1}$ a.e. $x^{\prime}$, by its definition $R_{x^{\prime}}$ is supported in $Q^{-1}\left(x^{\prime}\right)$, and

## (5.14)

$$
R_{x^{\prime}}\left\llcorner p_{h}^{-1}\left(x^{\prime}, x_{n}\right) \neq 0 \Leftrightarrow x_{n} \in \mathcal{J}_{x^{\prime}} \Leftrightarrow R_{x^{\prime}}\left\llcorner p_{h}^{-1}\left(x^{\prime}, x_{n}\right)=\left[v\left(x_{n}^{-} ; x^{\prime}\right), v\left(x_{n}^{+} ; x^{\prime}\right)\right]\right.\right.
$$

where $\left[v\left(x_{n}^{-} ; x^{\prime}\right), v\left(x_{n}^{+} ; x^{\prime}\right)\right]$ denotes the oriented line segment joining $v\left(x_{n}^{-} ; x^{\prime}\right)$ to $v\left(x_{n}^{+} ; x^{\prime}\right)$. Moreover, [1, Theorem 3.108] together with classical results about the rectifiability of the jump set imply that for $\mathcal{L}^{n-1}$ a.e. $x^{\prime}$,

$$
x_{n} \in \mathcal{J}_{x^{\prime}} \Rightarrow\left(x^{\prime}, x_{n}\right) \in \mathcal{J}_{D u} \text { and } T_{\left(x^{\prime}, x_{n}\right)} \mathcal{J}_{D u} \text { exists, }
$$

and moreover, the fact that $D u$ is a gradient implies that the jump in $D u$ across $\mathcal{J}_{D u}$ is normal to $T_{\left(x^{\prime}, x_{n}\right)} \mathcal{J}_{D u}$, if we identify the vertical directions (in which the jump occurs) and the horizontal directions (in which $T_{\left(x^{\prime}, x_{n}\right)} \mathcal{J}_{D u}$ lives). If we do not identify vertical and horizontal, this says that

$$
\begin{equation*}
v\left(x_{n}^{+}, x^{\prime}\right)-v\left(x_{n}^{-}, x^{\prime}\right) \in \mathbb{J}\left[\left(T_{\left(x^{\prime}, x_{n}\right)} \mathcal{\partial}_{D u}\right)^{\perp}\right] \tag{5.15}
\end{equation*}
$$

where $\left(T_{\left(x^{\prime}, x_{n}\right)} \mathcal{J}_{D u}\right)^{\perp}$ denotes the orthogonal complement in the horizontal directions only, as in (5.10).
Step 3. Conclusion of proof. Now we define $S_{x^{\prime}}=\left\langle[d u], Q, x^{\prime}\right\rangle-R_{x^{\prime}}$. In view of (5.5) and (5.13), it suffices to show that $S_{x^{\prime}}=0$ for $\mathcal{L}^{n-1}$ a.e. $x^{\prime}$. Note that this is also the conclusion of Proposition [5.3] in the case $f(x)=x$.

The various facts about $\left\langle[d u], Q, x^{\prime}\right\rangle$ and $R_{x^{\prime}}$ assembled above imply that for a.e. $x^{\prime}, S_{x^{\prime}}$ is an i.m. rectifiable 1-current with finite mass, and such that

$$
\partial S_{x^{\prime}}=0 \text { in } \Omega \times \mathbb{R}^{n}, \quad P_{1,0} S_{x^{\prime}}=0
$$

If $S_{x^{\prime}}$ has these properties, then it follows from Lemma 3.3, or from the $j=0$ case of Proposition 3.2, that there exist points $x^{i} \in q^{-1}\left(x^{\prime}\right)$ and i.m. rectifiable 1-currents $V_{i}$ (depending on $x^{\prime}$ in a way not captured by our notation) such that $V_{i}$ is supported in $\left\{x^{i}\right\} \times \mathbb{R}^{n}$ and

$$
\begin{equation*}
S_{x^{\prime}}=\sum V_{i}, \quad \partial V_{i}=0, \mathbf{M}\left(V_{i}\right)<\infty \tag{5.16}
\end{equation*}
$$

We fix one of these points $x^{i}$. Note that

$$
\begin{equation*}
V_{i}=S_{x^{\prime}}\left\llcorner p_{h}^{-1}\left(x^{i}\right)=\left[\left\langle[d u], Q, x^{\prime}\right\rangle\left\llcorner p_{h}^{-1}\left(x^{i}\right)\right]-\left[R_{x^{\prime}}\left\llcorner p_{h}^{-1}\left(x^{i}\right)\right]\right.\right.\right. \tag{5.17}
\end{equation*}
$$

It suffices to prove that
the unit tangent to $V_{i}$ is $\mathcal{H}^{1}$ a.e. constant, up to a sign
(i.e., independent of $\xi$ ), up to a sign, since then we can deduce from Solomon's Separation Lemma [19] and the Constancy Theorem that $V_{i}$ is either zero or has infinite mass, exactly as in Step 2 of the proof of Theorem4.1 Since $\mathbf{M}\left(V_{i}\right)<\infty$, we conclude that $V_{i}=0$, and since this holds for all $i$, it will follow that $S_{x^{\prime}}=0$ as needed.

We define

$$
\Sigma_{1}:=\left\{x=\left(x^{\prime}, x_{n}\right) \in \Omega: x_{n} \in \mathcal{J}_{x^{\prime}}\right\}, \quad \Sigma_{2}:=\left\{x=\left(x^{\prime}, x_{n}\right) \in \Omega: x \in \Gamma_{h}\left(x^{\prime}\right)\right\} .
$$

From what we have said above, $\left\langle[d u], Q, x^{\prime}\right\rangle\left\llcorner p_{h}^{-1}\left(x^{i}\right) \neq 0\right.$ if and only if $x^{i} \in \Sigma_{2}$, and $R_{x^{\prime}}\left\llcorner p_{h}^{-1}\left(x^{i}\right)=0\right.$ if and only if $x^{i} \in \Sigma_{1}$. In particular, $V_{i}=0$ unless $x^{i} \in$ $\Sigma_{1} \cup \Sigma_{2}$.

The unit tangent vector $\zeta$ to $\left\langle[d u], Q, x^{\prime}\right\rangle\left\llcorner p_{h}^{-1}\left(x^{i}\right)\right.$ for $x^{i} \in \Sigma_{2}$ is characterized in (5.10). And according to (5.14), for $x^{i} \in \Sigma_{1}$ the unit tangent vector to $R_{x^{\prime}}\left\llcorner p_{h}^{-1}\left(x^{i}\right)\right.$ is

$$
\frac{v\left(x_{n}^{+}, x^{\prime}\right)-v\left(x_{n}^{-}, x^{\prime}\right)}{\left|v\left(x_{n}^{+}, x^{\prime}\right)-v\left(x_{n}^{-}, x^{\prime}\right)\right|}=: \zeta_{R}\left(x^{i}\right)
$$

and is characterized in (5.15). Both $\zeta$ and $\zeta_{R}$ are independent of $\left(x^{i}, \xi\right) \in p_{h}^{-1}\left(x^{i}\right)$ up to a sign, so that (5.18) is immediate unless both terms on the right-hand side of (5.17) are nonzero. When this holds, we must show that $\zeta\left(x^{i}, \xi\right)= \pm \zeta_{R}\left(x^{i}\right)$ for a.e. $\xi$, and in view of (5.10), (5.15), it suffices to show that the approximate tangent spaces of $\Gamma_{h}$ and of $\mathcal{J}_{D u}$ coincide at $x^{i}$.

To do this, we claim that

$$
\mathcal{H}^{n-1}\left(\Sigma_{1} \backslash \Sigma_{2}\right)=0
$$

This follows from what we have already said, because if $x=\left(x^{\prime}, x_{n}\right) \in \Sigma_{1} \backslash \Sigma_{2}$, then (unless $x^{\prime}$ belongs to a set of $\mathcal{L}^{n-1}$ measure zero at which (5.16) fails to hold) (5.16) and (5.17) imply that $S_{x^{\prime}}=R_{x^{\prime}}\left\llcorner p_{h}^{-1}(x)\right.$ and hence that $\partial R_{x^{\prime}}\left\llcorner p_{h}^{-1}(x)=0\right.$, and as remarked earlier, it follows that $\mathbf{M}\left(R_{x^{\prime}}\left\llcorner p_{h}^{-1}(x)\right)=+\infty\right.$. This is impossible away from a set of $\mathcal{H}^{n-1}$ measure zero.

Now we appeal to the fact that if $\mathcal{H}^{k}(A \backslash B)=0$ and $A, B$ are $\mathcal{H}^{k}$ rectifiable, then $T_{x} A=T_{x} B$ at $\mathcal{H}^{k}$ a.e. point of $A$. Clearly $\mathcal{H}^{n-1}\left(\Sigma_{1} \backslash \mathcal{J}_{D u}\right)=0$, and also $0=\mathcal{H}^{n-1}\left(\Sigma_{1} \backslash \Sigma_{2}\right)=\mathcal{H}^{n-1}\left(\Sigma_{2} \backslash \Gamma_{h}\right)$. It follows that $T_{x} \Sigma_{1}=T_{x} \mathcal{J}_{D u}=T_{x} \Gamma_{h}$ at $\mathcal{H}^{n-1}$ a.e. $x \in \Sigma_{1}$, which proves that $(5.18)$ holds at every $x^{i}$, for $\mathcal{L}^{n-1}$ a.e. $x^{\prime}$.

The proof of Proposition5.1 has already established Proposition 5.3 when $f(x)=$ $x$. We deduce the general case of Proposition 5.3 from this special case by a change of variables, using a result from [7] that describes the behaviour of Monge-Ampère functions under coordinate transformations.

Proof of Proposition 5.3 We write $F: U \times \mathbb{R}^{n} \rightarrow V \times \mathbb{R}^{n}$ and $G: V \times \mathbb{R}^{n} \rightarrow U \times \mathbb{R}^{n}$ by $F(x, \xi)=\left(f(x), D g(f(x))^{T} \xi\right)$ and $G(y, \eta)=\left(g(y), D f(g(y))^{T} \eta\right)$. Note that $G=F^{-1}$. According to Fu [7, Proposition 2.5], $u \circ g$ is a Monge-Ampère function on $V$, and $[d(u \circ g)]=F_{\#}[d u]$. (This is proved for locally Lipschitz MongeAmpère functions, but the proof remains valid without that restriction.) Thus $[d u]=$ $G_{\#} F_{\#}[d u]=G_{\#}[d(u \circ g)]$, so that for a.e. $y^{\prime} \in \mathbb{R}^{n-1}$,
$\left\langle[d u], q \circ f \circ p_{h}, y^{\prime}\right\rangle=\left\langle G_{\#}[d(u \circ g)], q \circ f \circ p_{h}, y^{\prime}\right\rangle=G_{\#}\left\langle[d(u \circ g)], q \circ f \circ p_{h} \circ G, y^{\prime}\right\rangle$
using basic properties of slices, see Federer [5, 4.3.2 (7)]. From the definitions one checks that $q \circ f \circ p_{h} \circ G=q \circ p_{h}=Q$, in the notation of Proposition5.1, where we have proved that

$$
\left\langle[d(u \circ g)], Q, y^{\prime}\right\rangle=I_{v\left(\cdot ; y^{\prime}\right)}^{*}
$$

for a.e. $y^{\prime}$, with $\left.v\left(y_{n} ; y^{\prime}\right):=\left(y^{\prime}, y_{n}, D(u \circ g)\left(y^{\prime}, y_{n}\right)\right)\right)$.
To prove the proposition, we must therefore verify that $G_{\#} I_{v\left(\cdot ; y^{\prime}\right)}^{*}=I_{w\left(; y^{\prime}\right)}^{*}$, where $w$ is defined following (5.4). If $v:(0, L) \rightarrow V \times \mathbb{R}^{n}$ is smooth, then $I_{v\left(\cdot ; y^{\prime}\right)}^{*}=v_{\#}[[0, L]]$, where $[[0, L]]$ denotes the 1 -current corresponding to integration over $[0, L]$, and so

$$
G_{\#} I_{v\left(\cdot ; y^{\prime}\right)}^{*}=G_{\#} v_{\#}[[0, L]]=(G \circ v)_{\#}[[0, L]] .
$$

It is easy to check that $G \circ v\left(y_{n} ; y^{\prime}\right)=w\left(y_{n} ; y^{\prime}\right)$, so the corollary follows in this case. This remains valid if $v$ is continuous with bounded variation. If $v$ is merely a function of bounded variation, let us write $\mathcal{J}_{v}$ for the jump set of $v$. We split $I_{v}^{*}$ into a continuous part and a jump part:

$$
I_{v}^{*}=v_{\#}\left[[0, L] \backslash \mathcal{J}_{v}\right]+\left(I_{v}^{*}-v_{\#}\left[[0, L] \backslash \mathcal{J}_{v}\right]\right)
$$

These are the first and second terms, respectively, on the right-hand side of (5.2). It follows from what we have said that $G_{\#} v_{\#}\left[[0, L] \backslash \mathcal{J}_{v}\right]=w_{\#}\left[[0, L] \backslash \mathcal{J}_{v}\right]$. The other term $\left(I_{v}^{*}-v_{\#}\left[[0, L] \backslash \partial_{v}\right]\right)$ consists of a sum of straight line segments, each having the form $\left[\left(y, D(u \circ g)^{-}(y)\right),\left(y, D(u \circ g)^{+}(y)\right)\right]$ for some $y \in V$, and from the explicit form of $G$ one can verify that

$$
\begin{aligned}
& G_{\#}\left[\left(y, D(u \circ g)^{-}(y)\right),\left(y, D(u \circ g)^{+}(y)\right)\right] \\
&=\left[\left(g(y), D u^{-}(g(y))\right),\left(g(y), D u^{+}(g(y))\right)\right]
\end{aligned}
$$

Combining these, we obtain the desired result.

## 6 Weak Solutions of a Degenerate Monge-Ampère Equation

In this section we give the proof of the second main result of our paper.
Theorem 6.1 Suppose that $\Omega$ is a bounded open subset of $\mathbb{R}^{2}$. Assume that $u$ is a Monge-Ampère function on $\Omega$ that satisfies

$$
\begin{equation*}
[d u]\left(\phi d \xi_{1} \wedge d \xi_{2}\right)=0 \quad \text { for all } \phi \in C_{c}^{\infty}\left(\Omega \times \mathbb{R}^{2}\right) \tag{6.1}
\end{equation*}
$$

Then for every $x \in \Omega$, at least one of the following must hold:
(i) $u$ is affine in an open neighborhood of $x$.
(ii) There exists a line segment $\ell_{x}$, passing through $x$ and meeting $\partial \Omega$ at both endpoints, along which $D u$ is constant in the sense that
(a) every point along $\ell_{x}$ is a Lebesgue point of $D u$, with the same Lebesgue value; or
(b) every point on $\ell_{x}$ belongs to the jump set $\mathcal{J}_{D u}$ of $D u$, with same approximate limits on both sides of $\ell_{x}$.
In particular, every point in $\Omega$ is either a Lebesgue point of $D u$ or belongs to the jump set of $D u$.

As mentioned in the introduction, this extends earlier work of Hartman and Nirenberg [11], Kirchheim [15], and Pakzad [16].

The theorem shows that a function $u$ satisfying the hypotheses has the regularity of a BV function of a single variable. This is optimal: if $f: \mathbb{R} \rightarrow \mathbb{R}$ is an arbitrary function of bounded variation and $u(x, y)=f(x)$, then $u$ satisfies all the hypotheses of the theorem.

Our assumption (6.1) should be compared with the weaker condition:

$$
\begin{equation*}
[d u]\left(\phi d \xi_{1} \wedge d \xi_{2}\right)=0 \quad \text { for all } \phi(x, \xi)=\tilde{\phi}(x), \tilde{\phi} \in C_{c}^{\infty}(\Omega) \tag{6.2}
\end{equation*}
$$

This weaker condition requires only that the marginal on $\Omega$ of the measure in the product space vanishes. The conclusions of the theorem need not hold under assumption (6.2). To see this, let $B$ denote the unit ball in $\mathbb{R}^{2}$, and suppose that $u: B \rightarrow \mathbb{R}$ is the restriction to $B$ of a function that is homogeneous of degree 1 and smooth away from the origin. Then one can check that $u$ is Monge-Ampére, and that if $\phi(x, \xi)=\tilde{\phi}(x)$, then $[d u]\left(\phi d \xi_{1} \wedge d \xi_{2}\right)$ makes sense, and

$$
\begin{equation*}
[d u]\left(\phi d \xi_{1} \wedge d \xi_{2}\right)=A \tilde{\phi}(0), \quad A=\frac{1}{2} \int_{0}^{2 \pi} \gamma(\theta) \wedge \gamma^{\prime}(\theta) d \theta \tag{6.3}
\end{equation*}
$$

where $\gamma(\theta)=D u(\cos \theta, \sin \theta)$. This is proved, for example, in [13, Lemma 4.1, Remark 4]. From this one can check that, given any $u$ as above, one can find $c \in R$ such that $A=0$ for $u_{c}(x):=u(x)+c|x|$. Such a function $u_{c}$ satisfies (6.2) but not in general the conclusions of Theorem6.1

For functions $u$ as described above, the distributional determinant $\operatorname{det}^{\prime} D^{2} u$ exists and is given by $\operatorname{det}^{\prime} D^{2} u=A \delta_{0}$; see Remark 3 Thus the conclusions of the theorem do not hold if we assume that $u$ is a Monge-Ampère function such that $\operatorname{det}^{\prime} D^{2} u$ is well defined and vanishes.

Throughout the proof we will write $[d u]=\underline{\mathrm{T}}(\Gamma, m, \tau)$. The starting point of the proof is the following.

Lemma 6.2 Suppose that $\Omega$ is a bounded, open subset of $\mathbb{R}^{2}$, and assume that $u: \Omega \rightarrow \mathbb{R}$ is a Monge-Ampère function that satisfies (6.1). Then there exists a 1-rectifiable set $\Gamma_{v} \subset \mathbb{R}^{2}$ and an $\mathcal{H}^{1}$-measurable mapping $\tau_{v}: \Gamma_{v} \rightarrow \wedge_{1} \mathbb{R}^{2}$ such that $\left|\tau_{v}\right|=1$ a.e.,

$$
\text { at } \mathcal{H}^{1} \text { a.e. } \xi \in \Gamma_{v}, T_{\xi} \Gamma_{v} \text { exists and equals } \operatorname{span}\left\{\tau_{v}(\xi)\right\} ;
$$

and for $\mathcal{H}^{1}$ a.e. $\xi \in \Gamma_{v}$, there exists a horizontal i.m. rectifiable 1-current $H_{\xi}$ in $\Omega \times \mathbb{R}^{2}$ supported in $\Omega \times\{\xi\}$ and satisfying

$$
\begin{gather*}
\partial H_{\xi}=0 \quad \text { in } \Omega \times \mathbb{R}^{2}  \tag{6.4}\\
{[d u]\left(\phi d x_{i} \wedge d \xi_{j}\right)=\int_{\Gamma_{v}} H_{\xi}\left(\phi d x_{i}\right)\left\langle d \xi_{j}, \tau_{v}(\xi)\right\rangle \mathcal{H}^{1}(d \xi)} \tag{6.5}
\end{gather*}
$$

for every $\phi \in C_{c}^{\infty}\left(\Omega \times \mathbb{R}^{2}\right)$. (In particular, $\xi \in \Gamma_{v} \mapsto H_{\xi}\left(\phi d x_{i}\right)$ is $\mathcal{H}^{1}$-measurable for every such $\phi$.) Moreover, for every $\xi \in \Gamma_{v}$ at which $H_{\xi}$ is defined, there exists a horizontal unit vector $\tau_{h}(\xi)$, a collection of nonzero integers $\left\{m_{i}(\xi)\right\}$, and a collection of line segments $\left\{\ell_{i}(\xi)\right\}_{i}$, each parallel to $\tau_{h}(\xi)$ and with its endpoints in $\partial \Omega$, such that $p_{v}^{-1}(\xi) \cap \Gamma=\bigcup_{i} \ell_{i}(\xi) \times\{\xi\}$, and such that

$$
\begin{equation*}
H_{\xi}\left(\phi d x_{i}\right)=\sum_{i} \int_{\ell_{i}(\xi) \times\{\xi\}} \phi(x, \xi)\left\langle d x_{i}, \tau_{h}(\xi)\right\rangle m_{i}(\xi) \mathcal{H}^{1}(d x), \quad H_{\xi}\left(\phi d \xi_{j}\right)=0 \tag{6.6}
\end{equation*}
$$

for $\phi \in C_{c}^{\infty}\left(\Omega \times \mathbb{R}^{2}\right)$. And finally, $\tau_{h}(\xi) \cdot J \tau_{v}(\xi)=0$ for $\mathcal{H}^{1}$ a.e. $\xi \in \Gamma_{v}$.
We will eventually show that for a.e. $\xi$, each $\ell_{i}(\xi)$ is contained entirely in either the Lebesgue set of $D u$ or the jump set of $D u$. Moreover, we will prove that in the former case, $D u \equiv \xi$ along $\ell_{i}(\xi)$, and in the latter case, $\xi \in\left[D u^{-}(x), D u^{+}(x)\right]$ for every $x \in \ell_{i}(\xi)$. A main point will be to show that for a.e. $\xi, \xi^{\prime}$ and every $i, i^{\prime}$ either $\ell_{i}(\xi)=\ell_{i^{\prime}}\left(\xi^{\prime}\right)$ or $\ell_{i}(\xi) \cap \ell_{i^{\prime}}\left(\xi^{\prime}\right)=\varnothing$. This will be established by carrying out a blowup argument, and then classifying all homogeneous functions satisfying (6.1); these are the key points in the proof of the theorem.
Proof of Lemma 6.2 The proof follows closely that of Theorem 4.1, with the roles of horizontal and vertical reversed. We define

$$
\Gamma_{v}:=\left\{\xi \in \mathbb{R}^{2}: \mathcal{H}^{1}\left(p_{v}^{-1}(\xi) \cap \Gamma\right)>0\right\}
$$

which in view of Lemma 3.1 is a 1-rectifiable set. Proposition 3.2 with $p_{h}$ replaced by $p_{v}$, implies the existence of a current $H_{\xi}$ for $\mathcal{H}^{1}$ a.e. $\xi \in \Gamma_{v}$, such that (6.4), (6.5) hold. Then as in the proof of Theorem 4.1, the Lagrangian condition (2.6) and facts assembled in Lemma 3.1 again with $p_{h}$ and $p_{v}$ switched, imply that for $\mathcal{H}^{1}$ a.e. $\xi$, there exists a horizontal $\tau_{h}(\xi)$, determined up to a sign by the condition $\tau_{h}(\xi) \cdot J \tau_{\nu}(\xi)=0$, and such that unit tangent vectors to $H_{\xi}$ equal $\pm \tau_{h}(\xi), \mathcal{H}^{1}$ a.e. Then Solomon's Separation Lemma [19] and the constancy theorem imply that $H_{\xi}$ is a sum of indecomposables, each of which is supported in $\Omega \times\{\xi\}$ and consists of an oriented integer multiplicity line segment parallel to $\tau_{h}(\xi)$ and with no boundary in $\Omega$. Each such segment is bounded, since $\Omega$ is bounded, and so the endpoints must lie in $\partial \Omega$. These facts are summarized in (6.6). Since (6.5) is insensitive to the behavior of $H_{\xi}$ on $\mathcal{H}^{1}$ null sets, we can modify $H_{\xi}$ on such a set to arrange that (6.6) holds at every point where $H_{\xi}$ is defined. (This is simply for convenience.)

Fix $a \in \Omega$ and let $R=\operatorname{dist}(a, \partial \Omega)$. Let $\rho(x, \xi)=|x-a|$. For $r \in(0, R)$, we let $x_{r}(s)=a+r\left(\cos \frac{s}{r}, \sin \frac{s}{r}\right), s \in \mathbb{R} /(2 \pi r \mathbb{Z})$ be an arclength parametrization of $\partial B_{r}(a)$. We write $D u(\cdot ; r)$, or simply $D u(r)$, to denote the function $\mathbb{R} /(2 \pi r \mathbb{Z}) \rightarrow \mathbb{R}^{2}$ defined by $s \mapsto D u(s ; r)=D u\left(x_{r}(s)\right)$. Then for $\mathcal{L}^{1}$ a.e. $r \in(0, R), D u(\cdot ; r)$ is a function of bounded variation; see again [1, Chapter 3]. We will write $\partial_{s} D u(r)$ to denote the associated vector-valued derivative measure, with components $\partial_{s} u_{x_{j}}(r), j=1,2$. We also write $\mathcal{J}_{r}$ to denote the jump set of $D u(r)$.

Note that $\rho, R, x_{r}, D u(r)$ all depend on $a$ in a way that is not indicated in our notation.

Lemma 6.3 For $\mathcal{L}^{1}$ a.e. $r \in(0, R)$, if we define $v_{r}: \mathbb{R} /(2 \pi r \mathbb{Z}) \rightarrow \Omega \times \mathbb{R}^{2}$ by $v_{r}(s)=\left(x_{r}(s), D u\left(x_{r}(s)\right)\right)$, then $\langle[d u], \rho, r\rangle=I_{v_{r}}^{*}$, using notation from Lemma 5.2 In particular.

$$
\begin{align*}
& \langle[d u], \rho, r\rangle\left(\phi d \xi_{j}\right)=\int_{\mathbb{R} /(2 \pi r \mathbb{Z}) \backslash \mathcal{J}_{r}} \phi^{j}\left(x_{r}(s), D u\left(x_{r}(s)\right)\right) \partial_{s} u_{x_{j}}(r)(d s)  \tag{6.7}\\
& \quad+\int_{\mathcal{J}_{r}}\left(\int_{0}^{1} \phi^{j}\left(x_{r}(s), \theta D u^{+}\left(x_{r}(s)\right)+(1-\theta) D u^{-}\left(x_{r}(s)\right)\right) d \theta\right) \partial_{s} u_{x_{j}}(r)(d s)
\end{align*}
$$

where as usual $D u^{ \pm}\left(x_{r}(s)\right)=\lim _{\sigma \rightarrow s^{ \pm}} D u\left(x_{r}(\sigma)\right)$. Note from Lemma 5.2 that $I_{v}^{*}$ is the current associated with integration over a single Lipschitz curve, so that its support is just the image of this curve.

The above lemma is just a special case of Proposition5.3, with the the right-hand side written out in detail. (Strictly speaking, we need to cover $\Omega \backslash\{a\}$ by simply connected open sets, on each of which $\left(x_{1}, x_{2}\right)=x_{r}(s) \mapsto(r, s)$ defines a smooth change of coordinates, and apply Proposition 5.3 on each such open set.)

Proof of Theorem6.1 Step 1. We first claim that

$$
\begin{equation*}
\langle[d u], \rho, r\rangle\left(\phi d \xi_{j}\right)=\int_{\Gamma_{v}}\left\langle H_{\xi}, \rho, r\right\rangle(\phi)\left\langle d \xi_{j}, \tau_{v}(\xi)\right\rangle \mathcal{H}^{1}(d \xi) \tag{6.8}
\end{equation*}
$$

for $\mathcal{L}^{1}$ a.e. $r \in(0, R)$ and $\phi \in C_{c}^{\infty}\left(\Omega \times \mathbb{R}^{2}\right)$. This follows by slicing (6.5). To see this, fix $f \in C_{c}^{\infty}(0, R)$; then

$$
\begin{aligned}
\int_{0}^{R}\langle[d u], \rho, r\rangle\left(\phi d \xi_{j}\right) f(r) d r & =[d u]\left(\rho^{\#}(f(\cdot) d r) \wedge \phi d \xi_{j}\right) \\
& =\int_{\Gamma_{v}} H_{\xi}(\phi f \circ \rho d \rho)\left\langle d \xi_{j}, \tau_{v}(\xi)\right\rangle \mathcal{H}^{1}(d \xi) \\
& =\int_{0}^{R} \int_{\Gamma_{v}}\left\langle H_{\xi}, \rho, r\right\rangle(\phi)\left\langle d \xi_{j}, \tau_{v}(\xi)\right\rangle \mathcal{H}^{1}(d \xi) f(r) d r
\end{aligned}
$$

Since this holds for all $f$ as above, we deduce (6.8).
Step 2. We now define

$$
S_{r}(a):=\left\{\xi \in \Gamma_{v}: H_{\xi} \text { is well defined and }\left\langle H_{\xi}, \rho, r\right\rangle \neq 0\right\}
$$

We will normally write $S_{r}$ when there is no possibility of confusion. It follows from the explicit description of $H_{\xi}$ in Lemma6.2 that

$$
\begin{equation*}
S_{r}=\left\{\xi \in \Gamma_{v}: H_{\xi} \text { is well defined and }\left(\bigcup_{i} \ell_{i}(\xi)\right) \cap B_{r}(a) \neq \varnothing\right\} \tag{6.9}
\end{equation*}
$$

In particular, it follows that

$$
\begin{equation*}
S_{r^{\prime}} \subset S_{r} \text { when } r^{\prime}<r \tag{6.10}
\end{equation*}
$$

We will say $r$ is a "good radius" if $D u(r)$ is a function of bounded variation and 6.7), (6.8) hold. For a good radius $r$, we define

$$
D u^{*}\left(\partial B_{r}(a)\right):=\left\{D u\left(x_{r}(s)\right): s \notin \mathcal{J}_{r}\right\} \cup\left(\bigcup_{s \in \mathcal{J}_{r}}\left[D u^{-}\left(x_{r}(s)\right), D u^{+}\left(x_{r}(s)\right)\right]\right)
$$

where $[p, q]$ denotes the line segment $\{\theta p+(1-\theta) q: 0 \leq \theta \leq 1\}$. This is just the image via $D u$ of $\partial B_{r}(a)$, with the jumps filled in as usual. Where no confusion can result, we will often omit the center $a$ and write simply $D u^{*}\left(\partial B_{r}\right)$. It follows from Lemma6.3that $D u^{*}\left(\partial B_{r}\right)=p_{v}(\operatorname{supp}\langle[d u], \rho, r\rangle)$. Thus 6.8) implies that

$$
\begin{equation*}
\mathcal{H}^{1}\left(S_{r} \backslash D u^{*}\left(\partial B_{r}\right)\right)+\mathcal{H}^{1}\left(D u^{*}\left(\partial B_{r}\right) \backslash S_{r}\right)=0 \tag{6.11}
\end{equation*}
$$

Note that $D u^{*}\left(\partial B_{r}(a)\right)$ and $S_{r}(a)$ have complementary good properties. On one hand, $D u^{*}\left(\partial B_{r}(a)\right)$ directly encodes information about the pointwise behavior of the gradient of $u$. Also, it is not hard to see from the definition that $D u^{*}\left(\partial B_{r}(a)\right)$ is closed and connected; in fact it is the image of a single Lipschitz curve. Note in particular that $\operatorname{diam}\left(D u^{*}\left(\partial B_{r}\right)\right) \leq \mathcal{H}^{1}\left(D u^{*}\left(\partial B_{r}\right)\right)$. This need not hold for $S_{r}$ over which we have little or no control on $\mathcal{H}^{1}$ null sets. On the other hand, $S_{r}$ is directly related through (6.9) to the line segments $\ell_{i}(\xi)$ that we seek to understand.

In any case, (6.11) and (6.10) imply that if $r, r^{\prime}$ are good, then

$$
\begin{equation*}
D u^{*}\left(\partial B_{r^{\prime}}\right) \subset D u^{*}\left(\partial B_{r}\right) \text { when } r^{\prime}<r \tag{6.12}
\end{equation*}
$$

Note: From now on, in every assertion we make involving any $D u^{*}\left(\partial B_{r}(a)\right)$, we implicitly assume that $r$ is a good radius for the given center $a$.

Step 3. It follows from (6.10), 6.11) that $r \mapsto \mathcal{H}^{1}\left(D u^{*}\left(\partial B_{r}(a)\right)\right)$ is almost everywhere nondecreasing, for every $a$. We will prove that
(6.13) if $\mathcal{H}^{1}\left(D u^{*}\left(\partial B_{r}(a)\right)=0\right.$ for some $r>0$, then $u$ is affine in $B_{r}(a)$,
(6.14) if $\mathcal{H}^{1}\left(D u^{*}\left(\partial B_{r}(a)\right)\right) \rightarrow 0$ as $r \rightarrow 0$, then $a$ is a Lebesgue point of $D u$,

$$
\begin{equation*}
\text { if } \mathcal{H}^{1}\left(D u^{*}\left(\partial B_{r}(a)\right)\right) \geq \alpha>0 \forall r>0 \text {, then } a \text { is a jump point of } D u \text {. } \tag{6.15}
\end{equation*}
$$

These imply in particular that every point of $\Omega$ is either a Lebesgue point or a jump point of $D u$. Conclusions (6.13) and (6.14) follow easily from what we have already said. Indeed, if $\mathcal{H}^{1}\left(D u^{*}\left(\partial B_{r}\right)\right)=0$ for some $r \in(0, R)$, then since $D u^{*}\left(\partial B_{r}\right)$ is connected, it consists of a single point, say $D u^{*}\left(\partial B_{r}\right)=\left\{\xi_{a}\right\}$. Then (6.12) implies that $D u^{*}\left(\partial B_{r^{\prime}}\right)=\left\{\xi_{a}\right\}$ for $r<r^{\prime}$. Thus, in view of the definition of $D u^{*}\left(\partial B_{r}\right)$, $D u=\xi_{a} \mathcal{L}^{2}$ a.e. in $B_{r}(a)$, and so $u$ is affine in $B_{r}(a)$, with gradient $\xi_{a}$.

Similarly, if $\mathcal{H}^{1}\left(D u^{*}\left(\partial B_{r}\right)\right) \rightarrow 0$ as $r \rightarrow 0$, then (6.12) implies that there exists a point $\xi_{a}$ such that $\operatorname{dist}\left(D u^{*}\left(\partial B_{r}\right), \xi_{a}\right) \rightarrow 0$ as $r \rightarrow 0$, and then the definition of $D u^{*}\left(\partial B_{r}\right)$ implies that $a$ is a Lebesgue point of $D u$, with $D u(a)=\xi_{a}$.

Step 4. Blowup. In the next two steps we use a blowup argument to prove 6.15). Fix $a \in \Omega$ at which

$$
\begin{equation*}
\mathcal{H}^{1}\left(S_{r}\right)=\mathcal{H}^{1}\left(D u^{*}\left(\partial B_{r}\right)(a)\right) \geq \alpha>0 \text { for all } r>0 \tag{6.16}
\end{equation*}
$$

We will use a blowup argument to study the behavior of $u$ near such a point. First fix some good $r_{0}$. Then (6.12) and the definition of $D u^{*}\left(\partial B_{r}\right)$ imply that $D u(x) \in$ $D u^{*}\left(\partial B_{r_{0}}\right)$ for $\mathcal{L}^{2}$ a.e. $x \in B_{r_{0}}(a)$. Since $D u\left(r_{0}\right)$ has bounded variation, $D u^{*}\left(\partial B_{r_{0}}\right)$ is a bounded set, and so $u$ is Lipschitz in $B_{r}(a)$.

Now define $u_{\varepsilon}(x)=\frac{1}{\varepsilon}[u(a+\varepsilon x)-u(a)]$. Note that $u_{\varepsilon}$ is Lipschitz in $B_{r_{0} / \varepsilon}(0)$, with a uniform Lipschitz constant. We claim that there exists a function $u_{0}: \mathbb{R}^{2} \rightarrow \mathbb{R}$, homogeneous of degree 1 and satisfying (6.1), such that

$$
\begin{equation*}
u_{\varepsilon} \rightarrow u_{0} \text { locally uniformly and in the sense of Corollary } 4.3 \tag{6.17}
\end{equation*}
$$

The point is that we do not pass to a subsequence.
Step 4.1. Toward this end, we first observe that

$$
\left[d u_{\varepsilon}\right]=\eta_{\varepsilon \#}[d u], \quad \eta_{\varepsilon}(x, \xi)=\left(\frac{x-a}{\varepsilon}, \xi\right) .
$$

This is checked by verifying that the right-hand side satisfies the defining properties (2.5)-(2.8) of $\left[d u_{\varepsilon}\right]$. The above implies in particular that if $\phi \in C_{c}^{\infty}\left(\mathbb{R}^{2} \times \mathbb{R}^{2}\right)$ and if $\varepsilon$ is small enough that $\phi \circ \eta_{e} \in C_{c}^{\infty}\left(\Omega \times \mathbb{R}^{2}\right)$, then

$$
\begin{align*}
{\left[d u_{\varepsilon}\right]\left(\phi d x_{i} \wedge d \xi_{j}\right) } & =[d u]\left(\phi \circ \eta_{\varepsilon} \frac{d x_{i}}{\varepsilon} \wedge d \xi_{j}\right)  \tag{6.18}\\
& =\int_{\Gamma_{v}} H_{\xi}\left(\phi \circ \eta_{\varepsilon} \frac{d x_{i}}{\varepsilon}\right)\left\langle d \xi_{j}, \tau_{v}(\xi)\right\rangle \mathcal{H}^{1}(d \xi) \\
& =\int_{\Gamma_{v}} f_{\varepsilon}(\xi)\left\langle d \xi_{j}, \tau_{v}(\xi)\right\rangle \mathcal{H}^{1}(d \xi)
\end{align*}
$$

where we write

$$
f_{\varepsilon}(\xi):=H_{\xi}\left(\phi \circ \eta_{\varepsilon} \frac{d x_{i}}{\varepsilon}\right)
$$

Step 4.2. It follows from the explicit description of $H_{\xi}$ in Lemma6.2 that

$$
\begin{aligned}
f_{\varepsilon}(\xi) & =\sum_{i} m^{i}(\xi) \int_{\ell_{i}(\xi)} \phi\left(\frac{x-a}{\varepsilon}, \xi\right)\left\langle\frac{d x_{i}}{\varepsilon}, \tau_{h}(\xi)\right\rangle \mathcal{H}^{1}(d x) \\
& =\sum_{i} m^{i}(\xi) \int_{\eta_{\varepsilon}\left(\ell_{i}(\xi)\right)} \phi(x, \xi)\left\langle d x_{i}, \tau_{h}(\xi)\right\rangle \mathcal{H}^{1}(d x)
\end{aligned}
$$

Define $S_{0}(a)=\bigcap_{r>0} S_{r}(a)$. It follows from (6.9) that (writing $S_{0}$ for short)

$$
\begin{equation*}
S_{0}=\left\{\xi \in \Gamma_{v}: H_{\xi} \text { is well defined and } a \in \bigcup \ell_{i}(\xi)\right\} \tag{6.19}
\end{equation*}
$$

Then by sending $\varepsilon \rightarrow 0$, we find that for $\mathcal{H}^{1}$ a.e. $\xi \in \Gamma_{v}, f_{\varepsilon}(\xi) \rightarrow f_{0}(\xi)=H_{0}\left(\phi d x_{i}\right)$ as $\varepsilon \rightarrow 0$, where

$$
H_{0 \xi}\left(\phi d x_{i}\right):= \begin{cases}m(\xi) \int_{\ell(\xi)} \phi(x, \xi)\left\langle d x_{i}, \tau_{h}(\xi)\right\rangle \mathcal{H}^{1}(d x) & \text { if } \xi \in S_{0}  \tag{6.20}\\ 0 & \text { if not }\end{cases}
$$

where $\ell(\xi)$ is the line through the origin with tangent $\tau_{h}(\xi)$, and $m(\xi)=m^{i}(\xi)$ for the unique $i$ such that $a \in \ell_{i}(\xi)$.

Step 4.3. We now show that $\left|f_{\varepsilon}(\cdot)\right|$ is dominated by a locally integrable function. We fix $M$ such that $\phi$ is supported in $\{(x, \xi):|x| \leq M\}$, and we define

$$
F_{\varepsilon}(\xi):=\frac{1}{\varepsilon} \sum_{i} m^{i}(\xi) \mathcal{H}^{1}\left(\ell_{i}(\xi) \cap B_{\varepsilon M}(a)\right)=\frac{1}{\varepsilon}\left\|H_{\xi}\right\|\left(B_{\varepsilon M}(a) \times \mathbb{R}^{2}\right)
$$

For any straight line $\ell$, it is clear that $\varepsilon \mapsto \frac{1}{\varepsilon} \mathcal{H}^{1}\left(\ell \cap B_{\varepsilon M}(a)\right)$ is a nondecreasing function. Thus if $\varepsilon_{0} \leq r_{0} / M$, and $\varepsilon<\varepsilon_{0}$, then

$$
\left|f_{\varepsilon}(\xi)\right| \leq\|\phi\|_{\infty} F_{\varepsilon}(\xi) \leq\|\phi\|_{\infty} F_{\varepsilon_{0}}(\xi)
$$

Also, $F_{\varepsilon_{0}}(\cdot)$ is nonnegative, and it follows from (3.7) that

$$
\begin{equation*}
\int_{\Gamma_{v}} F_{\varepsilon_{0}}(\xi) \mathcal{H}^{1}(d \xi)<\infty \tag{6.21}
\end{equation*}
$$

Hence we deduce from the dominated convergence theorem and (6.18), (6.20)

$$
\begin{equation*}
\left[d u_{\varepsilon}\right]\left(\phi d x_{i} \wedge d \xi_{j}\right) \rightarrow \int_{S_{0}} H_{0 \xi}\left(\phi d x_{i}\right) \mathcal{H}^{1}(d \xi)\left\langle d \xi_{j}, \tau_{v}(\xi)\right\rangle \tag{6.22}
\end{equation*}
$$

as $\varepsilon \rightarrow 0$. Note from (6.10) and (6.16) that $\mathcal{H}^{1}\left(S_{0}\right)>0$.
Step 4.4. It is easy to check that $\left[d u_{\varepsilon}\right]\left(\phi d \xi_{i} \wedge d \xi_{j}\right)=0$ for $\phi \in C_{c}^{\infty}\left(\eta_{\varepsilon}(\Omega) \times \mathbb{R}^{2}\right)$, so (6.21) and (6.18) imply that $\left\|\left[d u_{\varepsilon}\right]\right\|\left(K \times \mathbb{R}^{2}\right) \leq C(K)<\infty$ for every compact $K$. Since the Lipschitz constants of $\left\{u_{\varepsilon}\right\}$ are locally uniformly bounded, we can pass to a subsequence $\left\{u_{\varepsilon_{n}}\right\}$ that converges to a limit locally uniformly and in the sense of Corollary 4.3. In order to show that the whole sequence converges, we must show that there is a unique such limit $u_{0}: \mathbb{R}^{2} \rightarrow \mathbb{R}$. Thus, suppose we have a different subsequence $\left\{u_{\varepsilon_{n}^{\prime}}\right\}$ converging to a limit $u_{1}$. We claim that $u_{0}=u_{1}$. Since $u_{\varepsilon}(0)=0$ for every $\varepsilon$, clearly $u_{0}(0)=u_{1}(0)=0$. And since $\phi$ was arbitrary in Step 4.3, we deduce that

$$
\begin{equation*}
\left[d u_{0}\right]\left(\phi d x_{i} \wedge d \xi_{j}\right)=\int_{S_{0}} H_{0 \xi}\left(\phi d x_{i}\right)\left\langle d \xi_{j}, \tau_{v}(\xi)\right\rangle \mathcal{H}^{1}(d \xi) \tag{6.23}
\end{equation*}
$$

for all $\phi \in C_{c}^{\infty}\left(\mathbb{R}^{2} \times \mathbb{R}^{2}\right)$ and all $i, j$. Clearly the same holds for $u_{1}$. It follows that $\left\langle\left[d u_{0}\right], \rho, r\right\rangle=\left\langle\left[d u_{1}\right], \rho, r\right\rangle$ for $\mathcal{L}^{1}$ a.e. $r$. Fix $r$ such that (6.7) holds for both [du $\left.u_{0}\right]$
and $\left[d u_{1}\right]$. By inspection of (6.7), we then infer that the $\mathbb{R}^{n}$-valued measures $\partial_{s} D u_{0}(r)$ and $\partial_{s} D u_{1}(r)$ on $\mathbb{R} / 2 \pi r \mathbb{Z}$ are equal, which in particular implies that $D u_{0}\left(x_{r}(\cdot)\right)$ and $D u_{1}\left(x_{r}(\cdot)\right)$ have the same jump sets, and moreover that $D u_{0}^{ \pm}\left(x_{r}(s)\right)=D u_{1}^{ \pm}\left(x_{r}(s)\right)$ at points $s$ in the jump set. We further read off from (6.7) that $D u_{0}\left(x_{r}(s)\right)=D u_{1}\left(x_{r}(s)\right)$ for $\left\|\partial_{s} D u_{0}(r)\right\|$ a.e. $s \in \mathbb{R} / 2 \pi r \mathbb{Z}$ away from the jump set. Also, from (6.22) one can check that $\phi \mapsto\left\langle\left[d u_{0}\right], \rho, r\right\rangle\left(\phi d \xi_{j}\right)$ is nonzero, which implies that $\left\|\partial_{s} D u_{0}(r)\right\|$ is a nonzero measure on $\mathbb{R} / 2 \pi r \mathbb{Z}$. These facts together imply that $D u_{0}\left(x_{r}(s)\right)=$ $D u_{1}\left(x_{r}(s)\right)$ for $\mathcal{L}^{1}$ a.e. $s$. Since this is true for a.e. $r$, we deduce that $D u_{0}=D u_{1}$, $\mathcal{L}^{2}$ a.e. Since $u_{0}(0)=u_{1}(0)=0$, it follows that $u_{0}=u_{1}$ as claimed.

Step 4.5. To finish the proof of (6.17) we only need to show that $u_{0}$ is homogeneous of degree 1 . This, however, follows from the fact that $u_{\varepsilon} \rightarrow u_{0}$ locally uniformly, since

$$
u_{0}(\lambda x)=\lim _{\varepsilon \rightarrow 0} u_{\varepsilon}(\lambda x)=\lambda \lim _{\varepsilon \rightarrow 0} u_{\lambda \varepsilon}(x)=\lambda u_{0}(x)
$$

Step 5. Description of homogeneous solutions. In this step we prove that there exist vectors $p^{+}, p^{-} \in \mathbb{R}^{2}$ such that the blowup limit $u_{0}$ found above satisfies

$$
D u_{0}(x)= \begin{cases}p^{+} & \text {if } x \cdot\left(p^{+}-p^{-}\right)>0  \tag{6.24}\\ p^{-} & \text {if } x \cdot\left(p^{+}-p^{-}\right)<0\end{cases}
$$

This amounts essentially to a classification of homogenous, degree 1 solutions of the equation $\operatorname{det} D^{2} u=0$ in the sense of (6.1).

Step 5.1. We first show that

$$
\begin{equation*}
\text { there exists } x_{0} \neq 0 \text { such that } u_{0}\left(x_{0}\right)=-u_{0}\left(-x_{0}\right) \tag{6.25}
\end{equation*}
$$

Toward this goal, we first notice that

$$
\left\langle\left[d u_{0}\right], \rho_{0}, r\right\rangle\left(\phi d \xi_{j}\right)=\int_{S_{0}}\left\langle H_{0 \xi}, \rho_{0}, r\right\rangle(\phi)\left\langle d \xi_{j}, \tau_{v}(\xi)\right\rangle
$$

for $r>0$. This follows by slicing (6.23) in exactly the same way that we deduced (6.8) from (6.5). (Note that for a homogeneous function, every $r$ is a good radius.) It follows that

$$
\operatorname{supp}\left\langle H_{0 \xi}, \rho_{0}, r\right\rangle \subset \operatorname{supp}\left\langle\left[d u_{0}\right], \rho_{0}, r\right\rangle\left(\phi d \xi_{j}\right) \quad \text { for } \mathcal{H}^{1} \text { a.e. } \xi \in S_{0}
$$

From the explicit form of $H_{0 \xi}$, we see that $\operatorname{supp}\left\langle H_{0 \xi}, \rho_{0}, r\right\rangle=\left\{\left( \pm x_{0}, \xi\right)\right\}$, where $\left\{ \pm x_{0}\right\}=\partial B_{r}(0) \cap \ell(\xi)$. Therefore, to prove (6.25) it suffices to show that

$$
\begin{equation*}
\text { if }(x, \xi) \in \operatorname{supp}\left\langle\left[d u_{0}\right], \rho, r\right\rangle, \text { then } u(x)=x \cdot \xi \tag{6.26}
\end{equation*}
$$

This follows essentially from the description of $\left\langle\left[d u_{0}\right], \rho_{0}, r\right\rangle$ in Lemma 6.3, which implies that

$$
\begin{align*}
\operatorname{supp}\left\langle\left[d u_{0}\right], \rho_{0}, r\right\rangle=\left\{\left(x_{r}(s)\right.\right. & \left., D u_{0}\left(x_{r}(s)\right): s \notin \mathcal{J}_{r}\right\}  \tag{6.27}\\
& \cup\left(\bigcup_{s \in \mathcal{J}_{r}}\left\{x_{r}(s)\right\} \times\left[D u_{0}^{-}\left(x_{r}(s)\right), D u_{0}^{+}\left(x_{r}(s)\right)\right]\right) .
\end{align*}
$$

The homogeneity of $u_{0}$ implies that

$$
\begin{equation*}
u_{0}(x)=x \cdot D u_{0}(x) \text { whenever } x \text { is a Lebesgue point of } D u \tag{6.28}
\end{equation*}
$$

Similarly, if $x \neq 0$ is a jump point of $D u_{0}$, then since $u_{0}$ is continuous, we can pass to limits in (6.28) along sequences of Lebesgue points approaching $x$ from both sides of the jump, to find that
(6.29) $x \cdot D u_{0}^{+}(x)=x \cdot D u_{0}^{-}(x)=u_{0}(x)$ and hence $x \cdot \xi=u_{0}(x)$ for $\xi \in\left[D u_{0}^{+}, D u_{0}^{-}\right]$.

Now (6.26) follows from (6.27), (6.28), and (6.29).
Step 5.2. Now define $v_{0}(x):=u_{0}(x)-\left(x \cdot \frac{x_{0}}{\left|x_{0}\right|^{2}}\right) u_{0}\left(x_{0}\right)$, and note that

$$
\begin{equation*}
v_{0}(\lambda x)=\lambda v_{0}(x) \text { for } \lambda>0, \quad v_{0}\left(x_{0}\right)=v_{0}\left(-x_{0}\right)=0, \quad \operatorname{det} D^{2} v_{0}=0 \tag{6.30}
\end{equation*}
$$

in the sense of (6.1). Let us write $v_{0}(r \cos \theta, r \sin \theta)=r f(\theta)$. We will prove that

$$
\begin{equation*}
\int_{0}^{2 \pi}\left(f^{2}-f^{\prime 2}\right) d \theta=0 \tag{6.31}
\end{equation*}
$$

We first prove this under the assumption that $v_{0}$ is smooth away from the origin. Then the condition $\operatorname{det} D^{2} v_{0}=0$ implies that the constant $A$ in (6.3) must equal zero, where we recall the definition:

$$
\begin{equation*}
A=A\left(v_{0}\right)=\frac{1}{2} \int_{0}^{2 \pi} \gamma(\theta) \wedge \gamma^{\prime}(\theta) d \theta, \quad \gamma(\theta)=D v_{0}(\cos \theta, \sin \theta) \tag{6.32}
\end{equation*}
$$

We will temporarily use the notation $n(\theta)=(\cos \theta, \sin \theta)$ and $t(\theta)=n^{\prime}(\theta)$, so that $\gamma(\theta)=n(\theta) f(\theta)+t(\theta) f^{\prime}(\theta)$. Then we easily compute that $\gamma \wedge \gamma^{\prime}=f\left(f+f^{\prime \prime}\right)$, so that (6.31) follows from the identity $A=0$ via integration by parts.

If $v_{0}$ is not smooth, then let $v_{k}(r \cos \theta, r \sin \theta)=r f_{k}(\theta)$ for a sequence of smooth $2 \pi$-periodic functions $f_{k}$ converging to $f$ in $W^{1,2}$ and such that $\int\left|f_{k}^{\prime \prime}\right| d \theta \leq C$. Such a sequence exists, since $D v_{0} \in B V$. This convergence implies that $v_{k} \rightarrow v$ in the sense of Corollary 4.3 (this is proved in [13, Proposition 4.1]). If we define $A_{k}=A\left(v_{k}\right)$ as in (6.32), then it follows that

$$
0=\lim _{k \rightarrow \infty} 2 A_{k}=\lim _{k \rightarrow \infty} \int_{0}^{2 \pi}\left(f_{k}^{2}-f_{k}^{\prime 2}\right) d \theta=\int_{0}^{2 \pi}\left(f^{2}-f^{\prime 2}\right) d \theta
$$

which proves (6.31) in the general case.
Step 5.3. In view of (6.30), we see that there exists $\alpha \in[0, \pi)$ such that $f(\alpha)=$ $f(\alpha+\pi)=0$. Then Poincaré's inequality implies that

$$
\int_{\alpha}^{\alpha+\pi}\left(f^{2}-f^{\prime 2}\right) d \theta \leq 0, \quad \int_{\alpha-\pi}^{\alpha}\left(f^{2}-f^{\prime 2}\right) d \theta \leq 0
$$

In view of (6.31), equality holds in both integrals, and so the optimality conditions in Poincaré's inequality imply that there exist $a, b$ such that $f(\theta)=a \sin (\theta-\alpha)$ for $\theta \in(\alpha, \alpha+\pi)$ and $f(\theta)=b \sin (\theta-\alpha)$ for $\theta \in(\alpha-\pi, \alpha)$. In other words,

$$
v_{0}(x)= \begin{cases}a x \cdot t(\alpha) & \text { if } x \cdot t(\alpha)>0 \\ b x \cdot t(\alpha) & \text { if } x \cdot t(\alpha)<0\end{cases}
$$

Since $u_{0}$ is the sum of $v_{0}$ and a linear function, (6.24) follows.
Step 6. It follows from Steps 4 and 5 that if $a$ is such that (6.16) holds, then $a \in \mathcal{J}_{D u}$. It also follows from Step 3 that if (6.16) does not hold, then $a \notin \mathcal{J}_{D u}$.

We now prove, continuing to assume (6.16), that there is a line segment $\ell_{a}$ passing through $a$ and meeting $\partial \Omega$ at both endpoints, such that

$$
\begin{equation*}
\ell_{a} \subset \mathcal{J}_{D u}, \quad\left[D u^{-}(a), D u^{+}(a)\right]=\left[D u^{-}(b), D u^{+}(b)\right] \forall b \in \ell_{a} \tag{6.33}
\end{equation*}
$$

Step 6.1. To do this, we will show below that $S_{0}$ as defined in (6.19)

$$
\begin{equation*}
S_{0}=\left[D u^{-}(a), D u^{+}(a)\right] \text { up to a set of } \mathcal{H}^{1} \text { measure } 0 \tag{6.34}
\end{equation*}
$$

and

$$
\begin{equation*}
\text { for } \mathcal{H}^{1} \text { a.e. } \xi \in S_{0}, \tau_{h}(\xi) \text { is tangent to } \mathcal{J}_{D u} \text { at } a . \tag{6.35}
\end{equation*}
$$

First, we demonstrate that these will prove 6.33). To do this, let $\ell_{a}$ be the line segment passing through $a$, tangent to $\mathcal{J}_{D u}$ at $a$, and terminating when it meets $\partial \Omega$. Then (6.35) implies that for $\mathcal{H}^{1}$ a.e. $\xi \in S_{0}$, the associated line segments $\ell_{i}(\xi)$ passing through $a$ all coincide with $\ell_{a}$. Let $b$ denote any other point on $\ell_{a}$. Then $b \in \ell_{i}(\xi)$ for $\mathcal{H}^{1}$ a.e. $\xi \in S_{0}(a)$. It follows that $S_{0}(a) \subset S_{0}(b) \subset S_{r}(b)$ for $r>0$, and hence that $b$ satisfies (6.16). Thus $b \in \mathcal{J}_{D u}$, and (6.34) implies that $\left[D u^{-}(a), D u^{+}(a)\right] \subset\left[D u^{-}(b), D u^{+}(b)\right]$. Reversing the roles of $a$ and $b$ establishes the opposite inclusion and so will prove (6.33), once we have proved (6.34), (6.35).

Step 6.2. We now prove (6.34), (6.35). Since all information about $S_{0}$ and $\tau_{h}(\xi), \xi \in$ $S_{0}$ is recorded in the blowup limit $u_{0}$, we may argue with $u_{0}$, about which we know everything, instead of $u$.

It is convenient to assume that $\mathcal{J}_{D u_{0}}$ is the $x_{2}$-axis. This can be achieved by a change of coordinates. Then there exist numbers $p_{1}^{+}, p_{1}^{-}, p_{2}$, such that

$$
D u_{0}\left(x_{1}, x_{2}\right)= \begin{cases}\left(p_{1}^{-}, p_{2}\right) & \text { if } x_{1}<0 \\ \left(p_{1}^{+}, p_{2}\right) & \text { if } x_{1}>0\end{cases}
$$

Moreover, $u_{0, x_{i} x_{j}}=0$ unless $i=j=1$, and $u_{0, x_{1} x_{1}}$ is a 1-dimensional Hausdorff measure restricted to the $x_{2}$ axis, multiplied by the constant $p_{1}^{+}-p_{1}^{-}$.

By combining (5.1), and (6.24), we arrive at

$$
\left[d u_{0}\right]\left(\phi d x_{i} \wedge d \xi_{j}\right)= \begin{cases}-\int_{\left\{\left(x_{1}, x_{2}\right) \in \mathbb{R}^{2}: x_{1}=0\right\}} \bar{\phi}(x) \mathcal{H}^{1}(d x) & \text { if } i=2, j=1 \\ 0 & \text { if not }\end{cases}
$$

where $\bar{\phi}(x)=\int_{p_{1}^{-}}^{p_{1}^{+}} \phi\left(x,\left(\xi_{1}, p_{2}\right)\right) \mathcal{L}^{1}\left(d \xi_{1}\right)$ for $x \in\left\{\left(x_{1}, x_{2}\right) \in \mathbb{R}^{2}: x_{1}=0\right\}$. By Fubini's Theorem,

$$
\left[d u_{0}\right]\left(\phi d x_{2} \wedge d \xi_{1}\right)=\int_{p_{1}^{-}}^{p_{1}^{+}} \int_{\mathbb{R}} \phi\left(\left(0, x_{2}\right),\left(\xi_{1}, p_{2}\right)\right) d x_{2} d \xi_{1}
$$

On the other hand, from (6.23) and (6.20) we write

$$
\left[d u_{0}\right]\left(\phi d x_{i} \wedge d \xi_{j}\right)=\int_{S_{0}} m(\xi) \int_{\ell(\xi)} \phi(x, \xi)\left\langle d x_{i}, \tau_{h}(\xi)\right\rangle \mathcal{H}^{1}(d x)\left\langle d \xi_{j}, \tau_{v}(\xi)\right\rangle \mathcal{H}^{1}(d \xi)
$$

From the fact that $\left[d u_{0}\right]\left(\phi d x_{i} \wedge d \xi_{j}\right)=0$ for $(i, j) \neq(2,1)$, we conclude that $\tau_{h}(\xi)=$ $\pm e_{2}, \tau_{v}(\xi)= \pm \varepsilon_{1}$ for $\mathcal{H}^{1}$ a.e. $\xi \in S_{0}$, which includes the claim (6.35). And by comparing the above two identities for $\left[d u_{0}\right]\left(\phi d x_{2} \wedge d \xi_{1}\right)$, we deduce (6.34), and as already noted, 6.33) follows.

Step 7. Let $\mathcal{A}:=\{x \in \Omega: u$ is affine in some neighborhood of $x\}$. To complete the proof, it remains to show that if $a \notin\left(\mathcal{A} \cup \mathcal{J}_{D u}\right)$, then there exists a line segment $\ell_{a}$ passing through $a$ and meeting $\partial \Omega$ at both endpoints, such that every point of $\ell_{a}$ is a Lebesgue point for $D u$, and $D u(x)=D u(a)$ for all $x \in \ell_{a}$.
Step 7.1. We first prove that there is a subset of $\Gamma_{v}$ of full $\mathcal{H}^{1}$ measure, say $\Gamma_{v}^{0}$, such that

$$
\begin{equation*}
\text { if } \xi \in \Gamma_{v}^{0} \text {, then } D u(x)=\xi \text { for every } x \in\left(\cup_{i} \ell_{i}(\xi)\right) \backslash \mathcal{J}_{D u} \tag{6.36}
\end{equation*}
$$

We will use the notation $f_{\phi, i}(\xi)=H_{\xi}\left(\phi d x_{i}\right)$. Recall that $f_{\phi, i}: \Gamma_{v} \rightarrow \mathbb{R}$ is $\mathcal{H}^{1}$-measurable for $i=1,2$ and every $\phi \in C_{0}\left(\Omega \times \mathbb{R}^{2}\right)$, which denotes as usual the sup-norm closure of $C_{c}^{\infty}\left(\Omega \times \mathbb{R}^{2}\right)$. Let $\left\{\phi_{k}\right\}_{k=1}^{\infty} \subset C_{c}^{\infty}\left(\Omega \times \mathbb{R}^{2}\right)$ be a countable dense subset of $C_{0}\left(\Omega \times \mathbb{R}^{2}\right)$. We fix $\epsilon>0$, and for each $k=1,2, \ldots$, we apply Lusin's Theorem to find that there exists a set $E_{k} \subset \Gamma_{v}$ such that $\mathcal{H}^{1}\left(E_{k}\right) \leq \epsilon 2^{-k}$, and such that the restriction to $\Gamma_{v} \backslash E_{k}$ of $f_{\phi_{k}, i}$ is continuous for $i=1$, 2. Let $E_{\epsilon}:=\bigcup E_{k}$, and let $\Gamma_{v}^{\epsilon}:=\Gamma_{v} \backslash E_{\epsilon}$. It follows that the restriction to $\Gamma_{v}^{\epsilon}$ of $\xi \mapsto H_{\xi}\left(\phi d x_{i}\right)$ is continuous for every $\phi \in C_{0}\left(\Omega \times \mathbb{R}^{2}\right)$ and $i=1,2$, and also that $\mathcal{H}^{1}\left(E_{\epsilon}\right)=\mathcal{H}^{1}\left(\Gamma_{v} \backslash \Gamma_{v}^{\epsilon}\right) \leq \epsilon$. By discarding a set of $\mathcal{H}^{1}$ measure 0 from $\Gamma_{v}^{\epsilon}$ if necessary, we can arrange that

$$
\begin{equation*}
\mathcal{H}^{1}\left(\left(\Gamma_{v}^{\epsilon}\right) \cap B_{s}(\xi)\right)>0 \quad \forall s>0 \tag{6.37}
\end{equation*}
$$

for every $\xi \in \Gamma_{v}^{\epsilon}$, while preserving the condition $\mathcal{H}^{1}\left(\Gamma_{v} \backslash \Gamma_{v}^{\epsilon}\right) \leq \epsilon$. We claim that

$$
\begin{equation*}
\text { if } \xi \in \Gamma_{v}^{\epsilon} \text {, then } D u(a)=\xi \text { for every } a \in\left(\cup_{i} \ell_{i}(\xi)\right) \backslash \mathcal{J}_{D u} \tag{6.38}
\end{equation*}
$$

This will prove (6.36), since if we define $\Gamma_{v}^{0}=\bigcup_{k} \Gamma_{v}^{\epsilon_{k}}$ for some sequence $\epsilon_{k} \rightarrow 0$, then (6.38) implies that $\Gamma_{v}^{0}$ has the properties required in (6.36).

To prove (6.38), fix $\xi \in \Gamma_{v}^{\epsilon}$ and $a \in\left(\bigcup_{i} \ell_{i}(\xi)\right) \backslash \mathcal{J}_{D u}$. Step 3 shows that in order to prove that $D u(a)=\xi$, it suffices to show that $\xi \in D u^{*}\left(\partial B_{r}\right)(a)$ for every good $r<\operatorname{dist}(a, \partial \Omega)$. Fix some such $r$, and fix a smooth $\phi d x_{i}$ with compact support in $B_{r}(a) \times \mathbb{R}^{2}$ such that $H_{\xi}\left(\phi d x_{i}\right) \neq 0$. This is possible because $a \in \ell_{i}(\xi)$ for some $i$. Since $f_{\phi, i}$ is continuous in $\Gamma_{v}^{\epsilon}$, we deduce that $H_{\xi^{\prime}}\left(\phi d x_{i}\right) \neq 0$ for all $\xi^{\prime} \in \Gamma_{v}^{\epsilon}$ sufficiently close to $\xi$. For such $\xi^{\prime}$, it follows that $\bigcup \ell_{i}\left(\xi^{\prime}\right)$ must intersect $B_{r}(a)$, or equivalently that $\xi^{\prime} \in S_{r}(a)$. Thus (6.37) implies that every neighborhood of $\xi$ intersects $S_{r}(a)$ in a set of positive measure, and hence (in view of (6.11)) every neighborhood of $\xi$ contains points of $D u^{*}\left(\partial B_{r}\right)(a)$. But $D u^{*}\left(\partial B_{r}\right)(a)$ is closed, so $\xi \in D u^{*}\left(\partial B_{r}\right)(a)$, and (6.38) follows.

Step 7.2. Note that if $\xi \in \Gamma_{v}^{0}$ then for every $i$, either $\ell_{i}(\xi) \subset \mathcal{J}_{D u}$ or $\ell_{i}(\xi) \cap \mathcal{J}_{D u}=\varnothing$. Indeed, if $\ell_{i}(\xi) \not \subset \mathcal{J}_{D u}$, then since $\ell_{i}(\xi)$ is a line segment, and $\mathcal{J}_{D u}$ is also a union of line segments, $\ell_{i}(\xi)$ and $\mathcal{J}_{D u}$ can only intersect transversally. But this would imply that $D u$ jumps where $\ell_{i}(\xi)$ crosses $\mathcal{J}_{D u}$, which is impossible, since $D u \equiv \xi$ on $\ell_{i}(\xi) \backslash \mathcal{J}_{D u}$.

Similarly, if $\xi, \xi^{\prime} \in \Gamma_{v}^{0}$ and $\ell_{i}(\xi), \ell_{j}\left(\xi^{\prime}\right)$ do not intersect $\mathcal{J}_{D u}$, then $\ell_{i}(\xi) \cap \ell_{i}\left(\xi^{\prime}\right)=$ $\varnothing$, since if $a \in \ell_{i}(\xi) \cap \ell_{j}\left(\xi^{\prime}\right)$, then $a$ is a Lebesgue point of $D u$ with $D u(a)=\xi=\xi^{\prime}$, which is clearly impossible. Thus if $\xi, \xi^{\prime} \in \Gamma_{v}^{0}$, then $\ell_{i}(\xi)$ and $\ell_{i^{\prime}}\left(\xi^{\prime}\right)$ either coincide or are disjoint.

Step 7.3. Now fix $a \notin \mathcal{J}_{D u} \cup \mathcal{A}$. In view of (6.13), (6.14),

$$
0<\mathcal{H}^{1}\left(D u^{*}\left(\partial B_{r}(a)\right)\right)=\mathcal{H}^{1}\left(S_{r}(a)\right) \rightarrow 0 \quad \text { as } r \rightarrow 0
$$

Let $S_{r}^{*}(a)=S_{r}(a) \cap D u^{*}\left(\partial B_{r}\right)(a) \cap \Gamma_{v}^{0}$, and let

$$
\mathcal{L}_{r}=\left\{\ell_{i}(\xi): \xi \in S_{r}^{*}(a), \ell_{i}(\xi) \cap B_{r}(a) \neq \varnothing\right\}
$$

For $\ell \in \mathcal{L}_{r}$, let $\tau_{\ell}$ denote a unit tangent vector. We have just argued that these segments are pairwise disjoint. This implies that if $r$ is small enough, then the signs of these tangent vectors can be chosen so that $\tau_{\ell} \cdot \tau_{\ell^{\prime}}>\frac{1}{2}$ for all $\ell, \ell^{\prime} \in \mathcal{L}_{r}$, and then it further follows from disjointness that there exists some unit vector $\tau$ such that

$$
\sup _{\ell \in \mathcal{L}_{r}}\left|\tau_{\ell}-\tau\right| \rightarrow 0 \quad \text { as } r \rightarrow 0
$$

Let $\ell_{a}$ denote the segment passing through $a$ with unit tangent $\tau$. Note that $\ell_{a}$ is not a subset of $\mathcal{J}_{D u}$, since by assumption $a \notin \mathcal{J}_{D u}$. So $\ell_{a}$ can only intersect $\mathcal{J}_{D u}$ transversally. If this occurs, then there must exist some $\ell \in \mathcal{L}_{r}$ that intersects $\mathcal{J}_{D u}$ transversally, which is impossible. So every point of $\ell_{a}$ is a Lebesgue point of $D u$.

We finally argue that $D u(x)=D u(a)$ for every $x \in \ell_{a}$. To see this, recall first from Step 3 that

$$
\sup _{\xi \in S_{r}^{*}(a)}|\xi-D u(a)| \leq \sup _{\xi \in D u^{*}\left(\partial B_{r}\right)(a)}|\xi-D u(a)| \rightarrow 0 \quad \text { as } r \rightarrow 0 .
$$

Fix $x \in \ell_{a}$ and, given $\varepsilon>0$, fix $\rho>0$ so small that $|\xi-D u(x)|<\varepsilon$ for all $\xi \in D u^{*}\left(\partial B_{\rho}(x)\right)$. Since the tangents to segments in $\mathcal{L}_{r}$ converge to $\tau$ as $r \rightarrow 0$, it is clear that for $r$ sufficiently small, every $\ell$ in $\mathcal{L}_{r}$ must intersect $B_{\rho}(x)$, which implies that $S_{r}^{*}(a) \subset S_{\rho}(x)$ for $r$ sufficiently small. Then (6.11) implies that $D u^{*}\left(\partial B_{\rho}(x)\right) \cap$ $S_{r}^{*}(a) \neq \varnothing$ for $r$ sufficiently small. Then

$$
\begin{aligned}
|D u(a)-D u(x)| & \leq \sup _{\xi \in S_{r}^{*}(a) \cap D u^{*}\left(\partial B_{\rho}(x)\right)}[|D u(a)-\xi|+|\xi-D u(x)|] \\
& \leq \varepsilon+\sup _{\xi \in S_{r}^{*}(a)}|D u(a)-\xi| \rightarrow \varepsilon \quad \text { as } r \rightarrow 0 .
\end{aligned}
$$

Since $\varepsilon$ is arbitrary, the conclusion follows.

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[^0]:    Received by the editors July 24, 2007. Published electronically December 4, 2009
    The author was partially supported by the National Science and Engineering Research Council of Canada under operating Grant 261955.

    AMS subject classification: 49Q15, 53C24.

